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DEPARTMENT OF PHYSICAL SCIENCES

WATER QUALITY OF DRINKING WATER IN HARYANA

Mohan Singh,

Professor in Chemistry, B.M University, Asthal Bohar, Rohtak

In Haryana particularly Hisar, Sersa, Rohtak, Sonipat due to rapid population growth, unsanitary disposal of waste and other human activities most of the water sources are becoming polluted. The prevailing practices of open defecation, unscientific disposal of human wastes and agricultural practices in most of the rural villages have increased the level of micro biological contamination in the waste from streams, springs and ground sources. Water quality problems caused by physical and chemical parameters have huge impacts on public health when the concentration is high. In many regions the drawdown of the water table every year is another serious problem with some shallow tube wells becoming non-functional. The water quality assess meant of specific quality parameters is essential at the various stages of development. This will help in the development of appropriate planning and remedial action for water quality improvement. There is clear need for an effective water quality monitoring to ensure as safe and sustainable water supply system. The research was carried out in various pasts particularly in villages in Haryana on drinking water by measuring the various water parameters like temperature, pH, electrical conductivity, total dissolved salts, told suspended solids, total alkalinity, hardness etc. From the samples collected. It was found.

(1) Majority of the sample were found with in the limits suggested by BIS as per pH is concerned.
(2) Regarding dissolved solids all the samples were found with in the limit.
(3) All the samples showed the total alkalinity was above the limit of BIS.
(4) The total hardness was also below the desired limit but in some places the hardness was found above permissible values.
(5) Fluoride in all sample were also below the desirable limit.
(6) Nitrate was also below desirable limit.

From the above points it was concluded that the hardness and total dissolved solids are of great importance. On the basis of physic chemical monitoring of the sample studied, it may be concluded that the majority samples collected is hard type and some samples should some moderately hard type.

IMPORTANT STEPS TAKEN FOR DRINKING WATER

The contamination of the drinking water is a serious problem which needs urgent attention. The available information's regarding water impurity can be utilized for policy formulation and management of water. There should be more attention in the analysis and farther use of collected data so that the end product of monitoring is information. If the data collecting is not utilized in framing strategies and management, thus the motive to collect data is lost. Now a day the computer has made it possible for public to use data more effectively and get the logical results of sample data analysis. Water monitoring is basically an information system. The science that serves as a basis for monitoring water system is evolving rapidly and is necessarily broad and complex. It needs to cover the nature of environmental decision making, aquatic ecology, the statistics of analyzing data chemistry of water, the toxicity of chemicals to biological organism , hydrology, data management hardware software and many other areas of science.

Now it is very necessary to apply basic water management in order to meet the requirement of potable water. Thus we need comprehensive water management which

includes rationing and masking. Based on the above facts, I offer the following recommendations to the policy makers and future researchers.

(1) Create awareness among people who frame polices.
(2) Policies should originate from actual survey and authentic works.
(3) Proper urbanization of villagers.
(4) Government should create an independent and autonomous organization water authority.
(5) There should be an effective and proper water management system.
(6) Proper and timely over hauling of the water distribution system.
(7) For water analysis new advanced technology and tools should be used for effective and accurate water analysis.
(8) Waste water drainage should be strengthened in order to avoid pathogens and related diseases.
(9) Change the habits of human being in favour of future adaptations and difficulties and problems.
(10) Water conservation, proper water use and recharge of water are the need of hour. Aware every one to have this mater as primary work of their life.

FOR FUTURE RESEARCHERS

Some important points are still left to farther research. For example:

(1) Include some more water quality parameters related to surface and underground water quality.
(2) To extend the study of quality of water with reference to time and season.
(3) To collect more data that represent time intervals and periods for greater accuracy and results.
(4) Use off some more models and theories to study the quality of water.
(5) To relate the human health with the water quality.
(6) To built the mathematical relationship between water quality and human health.
(7) Comporision and effectiveness of various laws available regarding water issues.
(8) To find and identify the policies and technology which minimize the negative environmental impacts while at the same time producing sufficient and satisfying demands.

ARTICLE ON TARGETED DRUG DELIVERY SYSTEM

Pallavi Bhardwaj, S.P.Sharma and Monika Chahar

Baba Mastnath University, Asthal Bohar Rohtak-124021 Haryana,E-mail: pallavi198277@gmail.com

ABSTRACT

Targeted drug delivery seeks to concentrate the medication in the tissues of interest while reducing the relative concentration of the medication in the remaining tissues. This improves efficacy of the while reducing side effects. It is very difficult for a drug molecule to reach its destination in the complex cellular network of an organism. Targeted delivery of drugs, as the name suggests, is to assist the drug molecule to reach preferably to the desired site. The inherent advantage of this technique has been the reduction in dose & side effect of the drug. Research related to the development of targeted drug delivery system is now a day is highly preferred and facilitating field of pharmaceutical world. A quantum dot is a semiconductor nanostructure which is particularly significant for optical applications due to their theoretically high quantum yield. Transdermal devices allow for pharmaceuticals to be delivered across the skin barrier. Molecules as diverse as small radiodiagnostic imaging agents to large DNA plasmid formulations have been successfully delivered inside FR-positive cells and tissue.

INTRODUCTION

Targeted drug delivery is a method of delivering medication to a patient in a manner that increases the concentration of the medication in some parts of the body relative to others. Drug targeting is the delivery of drugs to receptors or organs or any other specific part of the body to which one wishes to deliver the drugs exclusively. The drug's therapeutic index, as measured by its pharmacological response and safety, relies in the access and specific introduction of the drug with its candidate receptor, whilst minimizing its introduction with non –target tissue. The desired differential distribution of drug its targeted delivery would spare the rest of the body and thus significantly reduce the overall toxicity while maintaining its therapeutic benefits The targeted or site- specific delivery of drugs is indeed a very attractive goal because this provides one of the most potential ways to improve the therapeutic index of the drugs.

Recent approaches Quantam dots: A quantum dot is a semiconductor nanostructure that confines the motion of conduction band electrons, valence band holes, or excitons (Bound pairs of conduction band electrons and valence band holes) in all three spatial directions. The confinement can be due to electrostatic potentials (generated by external electrodes, doping, strain, impurities), the presence of an interface between different semiconductor materials (e.g. in core-shell nanocrystal systems), the presence of the semiconductor surface (e.g. semiconductor nanocrystal), or a combination of these. Quantum dots are particularly significant for optical applications due to their theoretically high quantum yield. The ability to tune the size of quantum dots is advantageous for many applications and it is one of the most promising candidates for use in solid-state quantum computation and diagnosis , drug delivery, Tissue engineering, catalysis, filtration and also textiles technologies.

Transdermal Approach: Transdermal drug delivery system is topically administered medicaments in the form of patches that deliver drugs for systemic effects at a predetermined and controlled rate. A transdermal drug delivery device, which may be of an active or a passive design, is a device which provides an alternative route for administering medication. These devices allow for pharmaceuticals to be delivered across the skin barrier. In theory, transdermal patches work very simply. A drug is applied in a relatively high dosage to the inside of a patch, which is worn on the skin for an extended period of time. Through a diffusion process, the drug enters the bloodstream directly through the skin. Since there is high concentration on the patch

and low concentration in the blood, the drug will keep diffusing into the blood for a long period of time, maintaining the constant concentration of drug in the blood flow.

Folate Targeting: Folate targeting is a method utilized in biotechnology for drug delivery purposes. It involves the attachment of the vitamin, folate (folic acid), to a molecule/drug to form a "folate conjugate". Based on the natural high affinity of folate for the folate receptor protein (FR), which is commonly expressed on the surface of many human cancers, folate-drug conjugates also bind tightly to the FR and trigger cellular uptake via endocytosis.

Molecules as diverse as small radiodiagnostic imaging agents to large DNA plasmid formulations have successfully been delivered inside FR-positive cells and tissues. FA also displays high affinity for the folate receptor (FR), a glycosylphosphatidyinositol-linked protein that captures its ligands from the extracellular milieu and transports them inside the cell via a non-destructive, recycling endosomal pathway. The FR is also a

recognized tumor antigen/biomarker. Because of this, diagnostic and therapeutic methods which exploit the FR's function are being developed for cancer.

Brain targeted drug delivery system: The brain is a delicate organ, and evolution built very efficient ways to protect it. The delivery of drugs to central nervous system (CNS) is a challenge in the treatment of neurological disorders. Drugs may be administered directly into the CNS or administered systematically (e.g., by intravenous injection) for

targeted action in the CNS. The major challenge to CNS drug delivery is the blood-brain barrier (BBB), which limits the access of drugs to the brain substance. Advances in understanding of the cell biology of the BBB have opened new avenues and possibilities for improved drug delivery to the CNS. Various strategies that have been used for

manipulating the blood-brain barrier for drug delivery to the brain include osmotic and chemical opening of the blood-brain barrier as well as the use of transport/carrier systems. Other strategies for drug delivery to the brain involve bypassing the BBB.

Various pharmacological agents have been used to open the BBB and direct invasive methods can introduce therapeutic agents into the brain substance.

It is important to consider not only the net delivery of the agent to the CNS, but also the ability of the agent to access the relevant target site within the CNS. Various routes of administration as well as conjugations of drugs, e.g., with liposomes and nanoparticles, are considered.

Liposomes: These are vesicular concentric structures, range in size from a nanometer to several micrometers, containing a phospholipids bilayer and are biocompatible, biodegradable and non immunogenic. Liposomes have generated a great interest because of their versatility and have played a significant role in formulation of potent drugs to improve therapeutics. Enhanced safety and efficacy have been achieved for a wide range of drug classes, including antitumor agents, antiviral, antimicrobials,

vaccines, gene therapeutics etc. . Recently pharmaceutical science is using liposomes to reduce toxicity and side effect of drugs. The various problems like poor solubility, short half life and poor bioavailability & strong side effect of various drugs can be overcome by employing the concept of liposomes especially in various diseases like cancer etc. Liposomes offer ample opportunities for the investigators to explore the unidentified breakthrough in the field of pharmaceutical technology.

CONCLUSION

Research related to the development of targeted drug delivery system is now a day is highly preferred and facilitating field of pharmaceutical world. It has crossed the infancy period and now touching height of growths from the pharmacy point of view. very difficult for a drug molecule to reach its destination in the complex cellular network of an

organism. Targeted delivery of drugs, as the name suggests, is to assist the drug molecule to reach preferably to the desired site. The inherent advantage of this technique has been the

reduction in dose & side effect of the drug. Overall it may be concluded with the vast database of different studies, the science of site specific or targeted delivery of these drugs has become wiser. Manifestation of these strategies in clinical now seems possible in near future.

REFERENCES

1. Chien Y.W., Novel drug delivery systems, Drugs and the Pharmaceutical Sciences, 50, New York, 797, 992 (2008)
2. Allen T. M. and Cullis P. R., Drug Delivery Systems, Entering the Mainstream Science, 1818 (2004).
3. Nacht S. and Kantz M, A., Novel Topical Programmable Delivery System, Topical Drug Delivery Systems, 299-325 (1992)
4. Won R., Method for delivering an active ingredient by controlled time release utilizing a novel delivery vehicle, 825 (1987)

SYNTHESIS AND APPLICATIONS OF $BaSO_4$ DOPED POLYANILINE

Monika Chahar[1*], Sushil Kumar[2], Pallavi Bhardwaj[1]

1. Baba Mastnath University, Asthal Bohar, Rohtak- 124 021, India. Email: monikachahar2009@gmail.com
2. Chaudhary Devi Lal University, Sirsa-125 055, India.

ABSTRACT

Synthesis of polyaniline (PANI) has been carried out by chemical oxidative polymerization. Chemical doping of $BaSO_4$ is completed in synthesized polyaniline (conducting polymer) by making different concentration of dopant in acidic medium. The synthesized conducting polymer has various potential applications in antistatic materials and they have been incorporated into commercial displays, organic solar cells, printing electronic circuits, organic light-emitting diodes, actuators, electrochromism, supercapacitors, chemical sensors and biosensors, flexible transparent displays, electromagnetic shielding and possibly replacement for the popular transparent conductor indium tin oxide. Another use is for microwave-absorbent coatings, particularly radar-absorptive coatings on stealth aircraft. The optical and structural behavior of doped polyaniline can be studied by the help of spectroscopic techniques such as UV-visible studies, fluorescence and FTIR.

Keywords: polyaniline, chemical doping, FTIR, UV-visible studies.

INTRODUCTION

Conductive polymers or, more precisely, intrinsically conducting polymers (ICPs) are organic polymers that conduct electricity[1]. Such compounds may have metallic conductivity or can be semiconductors. The biggest advantage of conductive polymers is their processability, mainly by dispersion. Conductive polymers are generally not thermoplastics, *i.e.*, they are not thermoformable. But, like insulating polymers, they are organic materials. They can offer high electrical conductivity but do not show similar mechanical properties to other commercially available polymers. The electrical properties can be fine-tuned using the methods of organic synthesis [2] and by advanced dispersion techniques [3]. They are organic conjugated polymers due to extended π-conjugation, and making these materials conducting towards the semi-conducting material range by reacting conjugated polymer with an oxidizing agent, reducing agent or a protonic acid. Conjugated polymers have enormous potential application in various fields such as conducting adhesives, electromagnetic shielding, chemical, biochemical and thermal sensors, OLEDs, and organic solar cells [4]-[5] etc. Alkaline earth's metal doped organic conducting polymers can explore the advanced class of conducting materials for potential applications. Different classes of organic conducting polymers are well studied which includes poly(acetylenes), poly(pyrrole), poly(thiophenes), poly(p-phenylenes sulphide) etc. [6]-[7]. Polyaniline is that conjugated polymer, which have great impression due to its good environmental stability. M. Husain et al have reported synthesis and characterization of polyaniline prepared with the dopant mixture of (ZrO_2/PbI_2) [8], Vazid et al. have reported the metal salts doped studies on conducting polymers and spectroscopic characterization [9]-[10]. In this paper we report a synthetic pathway for the preparation of polyaniline and chemical doping of salt of alkaline earth metal in the polyaniline matrix and its application. The synthesis of polyaniline has been carried out by chemical oxidative polymerization by using McDiarmid method [11].

EXPERIMENTAL DETAILS

Chemicals:

Aniline (Spectrochem Pvt. Ltd. Mumbai, India, Potassium dichromate (S.D. Fine Chemicals Ltd. Bombay L.R.Grade), Hydrochloric acid (Qualigens fine chemicals Ltd. Bombay L.R. Grade), Ammonia solution in water 28% (S.D Fine Chemicals Ltd. Bombay L.R.Grade), Tetrahydrofuran (Merck India Ltd. Bombay L.R.Grade), the doping agent $BaSO_4$ (Hi Media Laboratories Pvt. Ltd.) were used for synthesis and doping in polyaniline.

SYNTHESIS AND CHEMICAL DOPING:

Distilled aniline is used to synthesize polyaniline by chemical oxidative polymerization in acidic medium by adopting Mc Diarmid method. Synthesized polyaniline is dried in oven and grinded to achieve polyaniline powder. Polyaniline grinded powder and $BaSO_4$ (dopant) with different concentration such as 1, 2, 3, 4 and 5% (w/w) have been used in 20 ml THF solution for chemical doping with magnetic stirring about 20 minutes and kept the prepared samples for 40 hours at room temperature 25ºC and than further put the samples into oven at 40ºC for 24 hours and then temperature raises upto 110ºC to achieve moisture free doped polyaniline. The samples were doped with different concentration of dopant such as 1, 2, 3, 4 and 5% (w/w).

RESULTS AND DISCUSSION

The linear-backbone 'polymer blacks' (polyacetylene, polypyrrole, and polyaniline) and their copolymers are the main class of conductive polymers. Poly(p-phenylene vinylene) (PPV) and its soluble derivatives have emerged as the prototypical electroluminescent semiconducting polymers. Today, poly (3-alkylthiophenes) is the archetypical materials for solar cells and transistors.

Conductive polymers are prepared by many methods. Most conductive polymers are prepared by oxidative coupling of monocyclic precursors. Such reactions entail dehydrogenation:

$$n\ H–[X]–H \rightarrow H–[X]_n–H + 2(n–1)\ H^+ + 2(n–1)\ e^–$$

The low solubility of most polymers presents challenges. Some researchers have addressed this through the formation of nanostructures and surfactant-stabilized conducting polymer dispersions in water. These include polyaniline nanofibers and PEDOT:PSS. These materials have lower molecular weights than that of some materials previously explored in the literature. However, in some cases, the molecular weight need not be high to achieve the desired properties. But in my present work we can synthesized $BaSO_4$ doped polyaniline by using McDiarmid method.

They have potential applications in organic solar cells, printing electronic circuits, organic light-emitting diodes, actuators, electrochromism, supercapacitors, chemical sensors and biosensors, flexible transparent displays, electromagnetic shielding and possibly replacement for the popular transparent conductor indium tin oxide. Another use is for microwave-absorbent coatings, particularly radar-absorptive coatings on stealth aircraft. Conducting polymers are rapidly gaining attraction in new applications with increasingly processable materials with better electrical and physical properties and lower costs. The new nanostructured forms of conducting polymers particularly augment this field with their higher surface area and better dispersability. With the availability of stable and reproducible dispersions, PEDOT and polyaniline have gained some large scale applications. While PEDOT (poly(3,4-ethylenedioxythiophene)) is mainly used in antistatic applications and as a transparent conductive layer in form of PEDOT:PSS dispersions (PSS=polystyrene sulfonic acid), polyaniline is widely used for printed circuit board manufacturing – in the final finish, for protecting copper from corrosion and preventing its solderability. Electroluminescence is light emission stimulated by electrical current. In organic compounds, electroluminescence has been known since the early 1950s, when Bernanose and coworkers first produced electroluminescence in crystalline thin films of acridine orange and quinacrine. In 1960, researchers at Dow Chemical developed AC-driven electroluminescent cells using doping. In some cases, similar light emission is observed when a voltage is applied to a thin layer of a conductive organic polymer film. While electroluminescence was originally mostly of academic interest, the increased conductivity of modern conductive polymers means enough power can be put through the device at low voltages to generate practical amounts of light. This property has led to the development of flat panel displays using organic LEDs, solar panels, and optical amplifiers.

CONCLUSIONS

A semi-conducting polymer polyaniline doped with $BaSO_4$ is prepared by chemical oxidation polymerization method. The applications were studied and also we can say that $BaSO_4$ is playing a significant role in the chemical doping in polyaniline matrix because the electrons are responsible for absorption, in the chemical doping process of $BaSO_4$ in polyaniline, which exhibits the charge

transfer like absorption. It is assumed that $BaSO_4$ doped system is behaved as charge transfer complex along with polyaniline chain which can increase the conductivity.

REFERENCES

[1] Inzelt, György (2008). "Chapter 1: Introduction". In Scholz, F. Conducting Polymers: A New Era in Electrochemistry.

[2] Naarmann, Herbert (2000). "Ullmann's Encyclopedia of Industrial Chemistry".

[3] Handbook of Nanostructured Materials and Nanotechnology; Nalwa, H.S., Ed.; Academic Press: New York, NY, USA, 2000, **5**, 501–575.

[4] I. M. Cambell, Introduction to Synthetic Polymers, Oxford Science Publications, **196**, 1994

[5] H. S. Nalwa, Handbook of Organic Conductive Molecules and Polymers, Vol**. 1-4**, Wiley, NewYork, 1997

[6] Inzelt, G. "Conducting Polymers", (2008), Chap. 8, p.265-269.

[7] Herbert Naarmann "Polymers, Electrically Conducting" in Ullman's Encyclopedia of Industrial Chemistry, 2002, Wiley-VCH, Weinheim

[8] Sadia Ameen, G.B.V.S. Lakshmi and M. Husain, J. Phys. D: Appl.Phys. **42**(2009) 105104 (5pp).

[9] Sadia Ameen, Vazid Ali, M. Zulfequar, M. Mazharul Haq, M. Husain, Physica B: Condensed Matter, **403** (2008) 2861.

[10] Vazid Ali, Raminder Kaur, Neel Kamal, Sukhmehar Singh, S.C.Jain, H.P.S Kang, M. Zulfequar And M. Husain, J. Phys. & Chem. Solids, **67** (2006) 659.

[11] K.S. Ryu, B.W. Moon, J. Joo, S. Ho, Chang, Polymer, **42** (2001), 9355.

REVIEW OF NUCLEAR REACTORS

Sunita Dahiya

BMU, Asthal Bohar, Rohtak, e-mail Id: sunitadahiya@gmail.com

ABSTRACT

A nuclear reactor, formerly known as atomic pile, is a device used to initiate and control a sustained nuclear chain reaction. Nuclear reactors are used at nuclear power plants for electricity generation and in propulsion of ships. Heat from nuclear fission is passed to a working fluid (water or gas), which runs through turbines. These either drive a ship's propellers or turnelectrical generators. Nuclear generated steam in principle can be used for industrial process heat or for district heating. Some reactors are used to produce isotopes for medical and industrial use, or for production of weapons-grade plutonium. Some are run only for research. Today there are about 450 nuclear power reactors that are used to generate electricity in about 30 countries around the world. The neutron was discovered in 1932. The concept of a nuclear chain reaction brought about by nuclear reactions mediated by neutrons was first realized shortly thereafter, byHungarian scientist Leó Szilárd, in 1933. He filed a patent for his idea of a simple nuclear reactor the following year while working at the Admiralty in London.[10] However, Szilárd's idea did not incorporate the idea of nuclear fission as a neutron source, since that process was not yet discovered. Szilárd's ideas for nuclear reactors using neutron-mediated nuclear chain reactions in light elements proved unworkable. Inspiration for a new type of reactor using uranium came from the discovery by Lise Meitner, Fritz Strassmann and Otto Hahn in 1938 that bombardment of uranium with neutrons (provided by an alpha-on-beryllium fusion reaction, a "neutron howitzer") produced a barium residue, which they reasoned was created by the fissioning of the uranium nuclei. Subsequent studies in early 1939 (one of them by Szilárd and Fermi) revealed that several neutrons were also released during the fissioning, making available the opportunity for the nuclear chain reaction that Szilárd had envisioned six years previously.

On 2 August 1939 Albert Einstein signed a letter to President Franklin D. Roosevelt (written by Szilárd) suggesting that the discovery of uranium's fission could lead to the development of "extremely powerful bombs of a new type", giving impetus to the study of reactors and fission. Szilárd and Einstein knew each other well and had worked together years previously, but Einstein had never thought about this possibility for nuclear energy until Szilard reported it to him, at the beginning of his quest to produce the Einstein-Szilárd letter to alert the U.S. government. Shortly after, Hitler's Germany invaded Poland in 1939, starting World War II in Europe. The U.S. was not yet officially at war, but in October, when the Einstein-Szilárd letter was delivered to him, Roosevelt commented that the purpose of doing the research was to make sure "the Nazis don't blow us up." The U.S. nuclear project followed, although with some delay as there remained skepticism (some of it from Fermi) and also little action from the small number of officials in the government who were initially charged with moving the project forward. The following year the U.S. Government received the Frisch–Peierls memorandum from the UK, which stated that the amount of uranium needed for a chain reaction was far lower than had previously been thought. The memorandum was a product of the MAUD Committee, which was working on the UK atomic bomb project, known as Tube Alloys, later to be subsumed within the Manhattan Project. Eventually, the first artificial nuclear reactor, Chicago Pile-1, was constructed at the University of Chicago, by a team led by Enrico Fermi, in late 1942. By this time, the program had been pressured for a year by U.S. entry into the war. The Chicago Pile achieved criticality on 2 December 1942[11] at 3:25 PM. The reactor support structure was made of wood, which supported a pile (hence the name) of graphite blocks, embedded in which was natural uranium-oxide 'pseudospheres' or 'briquettes'.

CLASSIFICATIONS

Nuclear Reactors are classified by several methods; a brief outline of these classification methods is provided.

Classification by type of nuclear reaction

- Nuclear fission. All commercial power reactors are based on nuclear fission. They generally use uranium and its product plutonium as nuclear fuel, though a thorium fuel cycle is also possible. Fission reactors can be divided roughly into two classes, depending on the energy of the neutrons that sustain the fission chain reaction:
- Thermal reactors (the most common type of nuclear reactor) use slowed or thermal neutrons to keep up the fission of their fuel. Almost all current reactors are of this type. These contain neutron moderator materials that slow neutrons until their neutron temperature is thermalized, that is, until their kinetic energy approaches the average kinetic energy of the surrounding particles. Thermal neutrons have a far higher cross-section (probability) of fissioning the fissile nuclei uranium-235, plutonium-239, andplutonium-241, and a relatively lower probability of neutron capture by uranium-238 (U-238) compared to the faster neutrons that originally result from fission, allowing use of low-enriched uranium or even natural uranium fuel. The moderator is often also thecoolant, usually water under high pressure to increase the boiling point. These are surrounded by a reactor vessel, instrumentation to monitor and control the reactor, radiation shielding, and a containment building.
- Fast neutron reactors use fast neutrons to cause fission in their fuel. They do not have a neutron moderator, and use less-moderating coolants. Maintaining a chain reaction requires the fuel to be more highly enriched in fissile material (about 20% or more) due to the relatively lower probability of fission versus capture by U-238. Fast reactors have the potential to produce lesstransuranic waste because all actinides are fissionable with fast neutrons,[19] but they are more difficult to build and more expensive to operate. Overall, fast reactors are less common than thermal reactors in most applications. Some early power stations were fast reactors, as are some Russian naval propulsion units. Construction of prototypes is continuing (see fast breeder or generation IV reactors).
- Nuclear fusion. Fusion power is an experimental technology, generally with hydrogen as fuel. While not suitable for power production, Farnsworth-Hirsch fusors are used to produce neutron radiation.

CLASSIFICATION BY MODERATOR MATERIAL

Used by thermal reactors:

- Graphite-moderated reactors
- Water moderated reactors
- Heavy-water reactors (Used in Canada.)
- Light-water-moderated reactors (LWRs). Light-water reactors (the most common type of thermal reactor) use ordinary water to moderate and cool the reactors. When atoperating temperature, if the temperature of the water increases, its density drops, and fewer neutrons passing through it are slowed enough to trigger further reactions. That negative feedback stabilizes the reaction rate. Graphite and heavy-water reactors tend to be more thoroughly thermalized than light water reactors. Due to the extra thermalization, these types can use natural uranium/unenriched fuel.
- Light-element-moderated reactors. These reactors are moderated by lithium or beryllium.
- Molten salt reactors (MSRs) are moderated by a light elements such as lithium or beryllium, which are constituents of the coolant/fuel matrix salts LiF and BeF_2.
- Liquid metal cooled reactors, such as one whose coolant is a mixture of lead and bismuth, may use BeO as a moderator.
- Organically moderated reactors (OMR) use biphenyl and terphenyl as moderator and coolant.

CLASSIFICATION BY COOLANT

- Water cooled reactor. There are 104 operating reactors in the United States. Of these, 69 are pressurized water reactors (PWR), and 35 are boiling water reactors (BWR).[21]
- Pressurized water reactor (PWR) Pressurized water reactors constitute the large majority of all Western nuclear power plants.
- A primary characteristic of PWRs is a pressurizer, a specialized pressure vessel. Most commercial PWRs and naval reactors use pressurizers. During normal operation, a pressurizer is partially filled with water, and a steam bubble is maintained above it by heating the water with submerged heaters. During normal operation, the pressurizer is connected to the primary reactor pressure vessel (RPV) and the pressurizer "bubble" provides an expansion space for changes in water volume in the reactor. This arrangement also provides a means of pressure control for the reactor by increasing or decreasing the steam pressure in the pressurizer using the pressurizer heaters.
- Pressurised heavy water reactors are a subset of pressurized water reactors, sharing the use of a pressurized, isolated heat transport loop, but using heavy water as coolant and moderator for the greater neutron economies it offers.
- Boiling water reactor (BWR)
- BWRs are characterized by boiling water around the fuel rods in the lower portion of a primary reactor pressure vessel. A boiling water reactor uses ^{235}U, enriched as uranium dioxide, as its fuel. The fuel is assembled into rods housed in a steel vessel that is submerged in water. The nuclear fission causes the water to boil, generating steam. This steam flows through pipes into turbines. The turbines are driven by the steam, and this process generates electricity.[22] During normal operation, pressure is controlled by the amount of steam flowing from the reactor pressure vessel to the turbine.
- Pool-type reactor
- Liquid metal cooled reactor. Since water is a moderator, it cannot be used as a coolant in a fast reactor. Liquid metal coolants have included sodium, NaK, lead, lead-bismuth eutectic, and in early reactors, mercury.
- Sodium-cooled fast reactor
- Lead-cooled fast reactor
- Gas cooled reactors are cooled by a circulating inert gas, often helium in high-temperature designs, while carbon dioxide has been used in past British and French nuclear power plants. Nitrogen has also been used.[citation needed] Utilization of the heat varies, depending on the reactor. Some reactors run hot enough that the gas can directly power a gas turbine. Older designs usually run the gas through a heat exchanger to make steam for a steam turbine.
- Molten salt reactors (MSRs) are cooled by circulating a molten salt, typically a eutectic mixture of fluoride salts, such as FLiBe. In a typical MSR, the coolant is also used as a matrix in which the fissile material is dissolved.

CLASSIFICATION BY GENERATION

- Generation I reactor (early prototypes, research reactors, non-commercial power producing reactors)
- Generation II reactor (most current nuclear power plants 1965–1996)
- Generation III reactor (evolutionary improvements of existing designs 1996-now)
- Generation IV reactor (technologies still under development unknown start date, possibly 2030)
- The "Gen IV"-term was dubbed by the United States Department of Energy (DOE) for developing new plant types in 2000. In 2003, the French Commissariat à l'Énergie

Atomique (CEA) was the first to refer to Gen II types in Nucleonics Week; first mentioning of Gen III was also in 2000 in conjunction with the launch of the Generation IV International Forum (GIF) plans.

CLASSIFICATION BY PHASE OF FUEL

- Solid fueled
- Fluid fueled
- Aqueous homogeneous reactor
- Molten salt reactor
- Gas fueled (theoretical)

CLASSIFICATION BY USE

- Electricity
- Nuclear power plants including small modular reactors
- Propulsion, see nuclear propulsion
- Nuclear marine propulsion
- Various proposed forms of rocket propulsion
- Other uses of heat
- Desalination
- Heat for domestic and industrial heating
- Hydrogen production for use in a hydrogen economy
- Production reactors for transmutation of elements
- Breeder reactors are capable of producing more fissile material than they consume during the fission chain reaction (by converting fertile U-238 to Pu-239, or Th-232 to U-233). Thus, a uranium breeder reactor, once running, can be re-fueled with natural or even depleted uranium, and a thorium breeder reactor can be re-fueled withthorium; however, an initial stock of fissile material is required.
- Creating various radioactive isotopes, such as americium for use in smoke detectors, and cobalt-60, molybdenum-99 and others, used for imaging and medical treatment.
- Production of materials for nuclear weapons such as weapons-grade plutonium
- Providing a source of neutron radiation (for example with the pulsed Godiva device) and positron radiation (e.g. neutron activation analysis and potassium-argon dating)
- Research reactor: Typically reactors used for research and training, materials testing, or the production of radioisotopes for medicine and industry. These are much smaller than power reactors or those propelling ships, and many are on university campuses. There are about 280 such reactors operating, in 56 countries. Some operate with high-enriched uranium fuel, and international efforts are underway to substitute low-enriched fuel

REFERENCES

1. Newman, Jay (2008). Physics of the Life Sciences. Springer. p. 652. ISBN 978-0-387-77258-5.
2. "DOE Fundamentals Handbook: Nuclear Physics and Reactor Theory" (PDF). US Department of Energy. Archived from the original (PDF) on 23 April 2008. Retrieved24 September 2008.
3. "Reactor Protection & Engineered Safety Feature Systems". The Nuclear Tourist. Retrieved 25 September 2008.
4. "Bioenergy Conversion Factors". Bioenergy.ornl.gov. Retrieved 18 March 2011.
5. Bernstein, Jeremy (2008). Nuclear Weapons: What You Need to Know. Cambridge University Press. p. 312. ISBN 978-0-521-88408-2. Retrieved 17 March 2011.
6. "How nuclear power works". HowStuffWorks.com. Retrieved 25 September 2008.
7. Jump up to:[a b] "Reactor Protection & Engineered Safety Feature Systems". The Nuclear Tourist. Retrieved 25 September 2008.
8. Jump up^ "Chernobyl: what happened and why? by CM Meyer, technical journalist." (PDF).

9. Jump up^ Tsetkov, Pavel; Usman, Shoaib (2011). Krivit, Steven, ed. Nuclear Energy Encyclopedia: Science, Technology, and Applications. Hoboken, NJ: Wiley. pp. 48; 85. ISBN 978-0-470-89439-2.
10. Jump up^ L. Szilárd, "Improvements in or relating to the transmutation of chemical elements,"British patent number: GB630726 (filed: 28 June 1934; published: 30 March 1936).
11. Jump up^ The First Reactor, U.S. Atomic Energy Commission, Division of Technical Information
12. Jump up^ Enrico, Fermi and Leo, Szilard U.S. Patent 2,708,656 "Neutronic Reactor" issued 17 May 1955
13. Jump up^ Argonne's Nuclear Science and Technology Legacy: Chicago Pile reactors create enduring research legacy
14. Jump up^ Experimental Breeder Reactor 1 factsheet, Idaho National Laboratory
15. Jump up^ "Fifty years ago in December: Atomic reactor EBR-I produced first electricity" (PDF). American Nuclear Society Nuclear news. November 2001.
16. Jump up^ Kragh, Helge (1999). Quantum Generations: A History of Physics in the Twentieth Century. Princeton NJ: Princeton University Press. p. 286. ISBN 0-691-09552-3.
17. Jump up^ "On This Day: 17 October ". BBC News. 17 October 1956. Retrieved 9 November2006.
18. Jump up^ Leskovitz, Frank J. "Science Leads the Way". Camp Century, Greenland.
19. Jump up^ Golubev, V. I.; Dolgov, V. V.; Dulin, V. A.; Zvonarev, A. V.; Smetanin, . Y.; Kochetkov, L. A.; Korobeinikov, V. V.; Liforov, V. G.; Manturov, G. N.; Matveenko, I. P.; Tsibulya, A. M. (1993). "Fast-reactor actinoid transmutation". Atomic Energy 74: 83.doi:10.1007/BF00750983.

REVIEW ARTICLE ON NANOTECHNOLOGY IN ANCIENT HISTORY

Sanjay Kumar

Baba Mastnath University, Asthal Bohar, Rohtak-124001 (sanjaysagar8@gmail.com) M-08607755437.

ABSTRACT:
In the antiquities, nanoparticles were used by the Damascans to create swords with exceptionally sharp edges and the Romans to craft iridescent glassware. So were these archaic artisans also nanotechnologists? And what can today's scientists learn from such historic artefacts? In this article I review how nanotechnology come into existence and how it is used in ancient time.
KEYWORDS: Ancient, antiquities, nanoparticles, artisans.

INTRODUCTION:

The stunning Lycurgus cup reveals a brilliant red when light passes through its sections of glass containing gold-silver alloyed nanoparticles. The ancient empires of the world are remembered for their impressive large-scale feats of engineering: Macchu Picchu in Peru; the pyramids in Egypt; and the Parthenon in Greece to name a few. But the craftsmen of those eras were also skilled at engineering at the opposite end of the spectrum at the nanoscale.

The manipulation of material at the atomic and molecular scale to create new functions and properties sounds like it should be a profoundly modern concept. But artisans from the past also controlled matter at the tiniest scales. By modern-day standards, they were working in a branch of nanotechnology called nanocomposites. These are bulk materials in which nanoscale particles are mixed to improve the properties of the overall or composite material.

There are a number of relatively famous examples of ancient artefacts which were created using nanocomposites. The Lycurgus cup, for example, is a stunning decorative Roman treasure from about AD400; it is made of a glass that changes colour when light is shone through it. The glass contains gold-silver alloyed nanoparticles, which are distributed in such a way to make the glass look green in reflected light but, when light passes through the cup, it reveals a brilliant red.

A corrosion resistant azure pigment known as Maya Blue, first produced in AD800, was discovered in the pre-columbian Mayan city of Chichen Itza. It is complex material containing clay with nanopores into which indigo dye was combined chemically to create an environmentally-stable pigment.

Damascus steel swords from the Middle East were made between AD300 and AD1700 and are known for their impressive strength, shatter resistance and exceptionally sharp cutting edge. The steel blades contain oriented nanoscale wire-and-tube-like structures, which almost certainly enhanced the material's properties.

Pottery from across the Renaissance Mediterranean world was often decorated with an iridescent metallic glaze called lustre, the colour and sheen of which is a down to nanoparticles of copper or silver. So can we call the craftsmen who made these materials nanotechnologists? Ian Freestone at the Institute of Archaeology at University College London, who studied the Lycurgus cup, thinks not. "They were highly skilled but they were not nanotechnologists. They did not know that they were working on the nanoscale," he says.

Peter Paufler from TU Dresden, who led the research on the Damascus sword agrees. "They developed materials by trial and error similar to evolution in biology. They didn't know the processes going on inside the solids."

High-resolution microscopic analysis is used to reveal the nanostructure of these artefacts, but such analysis cannot tell us how they were made. "How did they dissolve these metals into the glass?" says Freestone of the Roman glassmakers who made the Lycurgus cup. "And

how did they get such a homogenous distribution of nanoparticles? We can speculate but we really don't know for sure." However, historical texts from Spain, Italy and the Middle East have revealed how lustreware was made, offering a new procedure for generating metal-glass nanocomposites.

A disadvantage of using high-resolution microscopy is that samples must be milled down to a fraction of their original thickness, destroying part of the artefact. When the material is abundant, such as for weather-resistant Maya Blue, taking a little for analysis is not a big deal. But when these artefacts are rare, it's more difficult to justify. "There's no way it would have been acceptable to remove a sample of the Lycurgus cup," says Freestone. "It's too unique, too valuable. Fortunately for us, some fragments of the glass were found in its metal base several decades ago and were saved."In the case of the steel sword, the curators at the Historical Museum Berne, which donated the sword, weighed up the potential benefits against the loss of one of their collection pieces. "Workers obviously felt that sufficient original blades were available to sacrifice some for research," says Paufler Such sacrifices are worth it. Some of these studies are providing pointers for new nanotechnology research. Understanding the nanoscale mechanism underlying how Maya Blue works has generated a new direction for scientists to investigate stable hybrid nanostructured pigments. Researchers from the French National Centre for Scientific Research have investigated a variety of nanoporous materials in which to insert and stabilise organic dyes.

Researchers from the University of Turin have also investigated the potential of the Maya Blue method for generating environmental-resistant pigments from different coloured dyes.

Similarly, understanding the properties of the Lycurgus cup has led to new nanocomposites with potentially useful optical properties. For example, researchers have developed thin nanocomposite films containing gold nanoparticles which can reflect infra-red while still transmitting light. Ivan Parkin, from UCL, who took part in the research, says: "We effectively tried to replicate the effect seen in the Lycurgus cup to recreate the extraordinary coloration efficiency that the Romans achieved." These films could be used to coat windows in hot countries to reflect heat away while allow light through the glass, thus reducing the need for air conditioning.

CONCLUSION

These historic structures are the results of hundreds of years of trial and error experimentation with craftsmen passing their skills and know-how down through generations. Nanotechnologists can also now build on this ancient wisdom. But they benefit from a modern understanding of the behaviour of atoms and molecules along with state-of-the-art fabrication tools and analytical instruments to achieve exciting new products and devices in a fraction of the time.

DNA MICROARRAY: AN INTERDISCIPLINARY APPROACH AND ITS APPLICATIONS

Ashish Katyal*[1], Neha Katyal[2], Dr. S.P.Sharma[2]

[1] Department of Biotechnology, Meerut Institute of Engineering and Technology, Meerut-250005, India, (aashishkatyaal@gmail.com) M- 09997031901

[2] Baba Mastnath University, Rohtak- 124001, India, (nehakatyal1@gmail.com) M- 09728255584.

ABSTRACT:

Using microarrays for high-throughput DNA detection is one of the fastest growing fields in technology today. However, the field is so new that most users do not know how to incorporate standards and controls into their experiments. DNA microarrays are a powerful emerging technology that scientists use to measure the activity (transcription) of thousands of genes at one time. Initially designed to measure gene transcriptional levels, microarray technologies are now used for comparing other genome features among individuals and their tissues and cells. Results provide valuable information on disease subcategories, prognosis, and their treatment. Microarrays highlight important connections between genetics, cell biology, genes, DNA, chromosomes, gene expression, transcription, cancer biology, proteins, technology, and bioethics. This paper gives a review of DNA microarray technology and its use in gene expression studies, its versatility, and how dramatically it is changing the molecular approach to biomedical research by merging strides in genomics, computer science, and nanotechnology.

INTRODUCTION:

Analysis of gene expression is important in many fields of biological research, since changes in the physiology of an organism or a cell will be accompanied by changes in the pattern of gene expression. Microarray analysis is a powerful new research tool that enables technicians to view and interpret at one time, on one small surface, the extent to which thousands of genes have been expressed in cells. Genes are "differentially regulated": All cells in an organism contain the same genes*, but different genes are expressed (transcribed) in different tissues under different conditions. This is what gives different tissues their different phenotypes (appearance and function).

[Gametes contain half of the genes that somatic cells do, and enucleated cells (such as mature red blood cells) do not contain genes.]

Even genes that are not highly expressed (transcribed) may play an important role in the cell. Studying which genes are active and which are inactive in different cell types helps scientists to understand both how these cells function normally and how they are affected when various genes do not perform properly. In the past, scientists have only been able to conduct these genetic analyses on a few genes at once. With the development of DNA microarray technology, however, scientists can now examine how active thousands of genes are at any given time. The study of gene expression by DNA microarray technology, which is still in development, is based on hybridisation of mRNA to a high-density array of immobilised target sequences, each corresponding to a specific gene.

DNA microarrays are created by robotic machines that arrange minuscule amounts of hundreds or thousands of gene sequences on a single microscope slide. Researchers have a database of over 40,000 gene sequences that they can use for this purpose. When a gene is activated, cellular machinery begins to copy certain segments of that gene. The resulting product is known as messenger RNA (mRNA), which is the body's template for creating proteins. The mRNA produced by the cell is complementary, and therefore will bind to the original portion of the DNA strand from which it was copied.

To determine which genes are turned on and which are turned off in a given cell, a researcher must first collect the messenger RNA molecules present in that cell. The researcher then labels each mRNA molecule by using a reverse transcriptase enzyme (RT) that generates a complementary cDNA to the mRNA. During that process fluorescent nucleotides are

attached to the cDNA. The tumor and the normal samples are labeled with different fluorescent dyes. Next, the researcher places the labeled cDNAs onto a DNA microarray slide. The labeled cDNAs that represent mRNAs in the cell will then hybridize – or bind – to their synthetic complementary DNAs attached on the microarray slide, leaving its fluorescent tag. A researcher must then use a special scanner to measure the fluorescent intensity for each spot/areas on the microarray slide.

If a particular gene is very active, it produces many molecules of messenger RNA, thus, more labeled cDNAs, which hybridize to the DNA on the microarray slide and generate a very bright fluorescent area. Genes that are somewhat less active produce fewer mRNAs, thus, less labeled cDNAs, which results in dimmer fluorescent spots. If there is no fluorescence, none of the messenger molecules have hybridized to the DNA, indicating that the gene is inactive. The use of two differently labelled mRNA samples allows quantitative comparison of gene expression in both samples (Fig. 1). Researchers frequently use this technique to examine the activity of various genes at different times.

MANUFACTURING OF DNA MICROARRAYS

A number of techniques are being developed for manufacturing DNA microarrays. Two main approaches exist. The first approach encompasses direct synthesis of oligonucleotides on a solid surface. Best known is synthesis based on photolithography as has been developed by Fodor et al. (1991). Specific areas of a glass surface, derivatised with linker molecules carrying a photo-labile protective group, are selectively illuminated by using a photomask. Subsequently, the surface is incubated with a solution containing a photo protected nucleotide, which will only be coupled to the light-activated areas. After removal of the excess nucleotide, a second photo mask is used to deprotect other areas on the surface and subsequently another type of nucleotide is coupled to these areas. By repeating this procedure, a defined set of oligonucleotides is synthesised on the surface. The method allows the manufacturing of microarrays with very high densities (\250 000 oligonucleotide spots per cm2)and facilitates the production of large series of identical arrays. Unfortunately, the design of this type of array, also named a DNA chip, is prohibitively expensive, has no flexibility in design and only a limited number of different arrays is on the market yet. The second approach, DNA micro-dispensing, is more flexible and can be performed in a regular molecular biology laboratory. Small quantities of DNA solution, with a minimum volume of approximately 50 pl, are dispensed onto a solid surface. The number of micro-dispensing robots commercially available is quickly increasing and the performance of these machines is continually improving. They can be subdivided into two main categories: passive dispensers and active dispensing units based on ink-jet technology (Fig. 2). Ink-jet based devices use either piezoelectric delivery or delivery by solenoid valves, making direct surface contact redundant. The density of the spots depends on the skills of the dispensing device. DNA dispensing is flexible and allows for constant update of the array. Using micro-dispensing it should also be possible to synthesise oligonucleotides directly on an array or to deposit molecules other than nucleic acids, for example proteins. The DNA molecules can be dispensed onto different kinds of surface (e.g. glass or membranes) to which the DNA is then fixed.

APPLICATIONS

There are various applications using Microarray technology these include:

1) Use for safety assessment of genetic modification in food plants. This should not only be based on the evaluation of the newly introduced trait, but also on possible unintended side effects resulting from the genetic alteration.
2) Use of DNA microarrays for gene expression and discovery. Measuring transcript levels for thousands of genes in parallel is one of the more widespread applications of DNA microarray technology. They documented the stability of the cDNA microarray

technology for profiling diseases and for identifying disease-related genes. The technology provided new targets for drug development and disease therapies. Their data showed that the use of representational difference analysis essentially provided an enriched library of differentially expressed genes, while analysis of the library with microarray technology allowed rapid and reproducible screening of thousands of DNA molecules simultaneously.

3) DNA microarrays are also used for drug discovery and development. Its applications to gene discovery, gene expression, and mapping have been convincingly demonstrated. The suitability of the cDNA microarray for profiling diseases and for identifying disease-related genes has been also well documented.
4) DNA microarray helps in Detection and Supervised Classification of Biomarkers which allows to choose appropriate pharmacological or surgical therapy. It can also identify signature sequences which futher useful for very early detection of diseases by complementing the clinical and histo-pathological analysis. Now more complex Biomarkers having multi-gene nature can also be detected by using DNA microarray.
5) DNA Microarray can do Unsupervised classification that are helpful in establishing a relationship between the molecular state of biological samples. As there are various similar expression profile known as Co-expressed genes which are regulated by same transcriptional factor (i.e. belonging to same metabolic pathways) can be detected which gives a broader overview of potential genes that have similar biological function and can be utilized as potential clinical targets.
6) DNA microarray can be used to detect abnormalities that are associated with single gene or gene copy number which is useful for hybridization experiment also known as Comparative Genomic Hybridization (CGH). Certain other genetic disorders can also be detected using microarray such as Single Nucleotide Polymorphism (SNP) associated diseases like diabetes etc.

CONCLUSION

Microarrays technology provide an effective tools based approach for simultaneously monitoring of gene expression, disease diagnosis, drug discovery and development, analyzing biochemical pathways, functionality monitoring of unknown genes in many areas of biomedical research. There are various software, development platforms and tools are available as freeware on web such as ‘‘bioconductor or BioC’’ which play an important role in monitoring various cancer causing genes. DNA microarray can use genetic markers across the genome to predict individual drug responsiveness. Also, DNA microarrays can analyse genetic linkage analysis which leads to identification of the disease genes under study. Microarrays can rapidly diagnose (in less than 24 hours) gene sequences in the entire genomes of pathogens without need of bacterial and viral cultures with an aim of providing a diagnostic tool that detects expression pattern of antibiotic resistance genes or other specific viral subtypes. DNA microarray majorly useful in analysis of Gene expression ,Genotyping and DNA sequencing which play a potential role in clinical application in biomedical technology.

REFERENCES

[1] Alwine, J.C., Kemp, D.J., Stark, G.R., 1977. Method for detection of specific RNAs in agarose gels by transfer to diazobenzyloxymethyl-paper and hybridization with DNA probes. Proc. Natl. Acad. Sci. USA 74, 5350–5354.

[2] International Human Genome Sequencing Consortium. Initial sequencing and analysis of the human genome. *Nature* 2001;409:860-921.

[3]To affinity . . . and beyond! [editorial]. *Nature Genet* 1996; 14:367-70.

[4] Brown PO, Botstein D. Exploring the new world of the genome with DNA microarrays. *Nature Genet* 1999; 21 (suppl):33-7.

[5] Dudoit Y, Yang Y, Callow MJ, Speed TP. Statistical methods for identifying differentially expressed genes in replicated cDNA microarray experiments. Stat Sin 2002; 12:111-39.

[6] Fan J, Tam P, Vande Woude G, Ren Y. Normalization and analysis of cDNA micro-arrays using within-array replications applied to neuroblastoma cell response to a cytokine. Proc Natl Acad Sci U S A 2004;101: 1135 - 40.
[7] Schena M, Shalon D, Heller R, Chai A, Brown PO, Davis RW. Parallel human genome analysis: Microarray-based expression monitoring of 1000 genes. Proc Natl Acad Sci USA 1996;93: 10614–9.
[8] Johnston, M., 1998. Gene chips: array of hope for understanding gene regulation. Curr. Biol. 8, R171–R174.
[9] Jordan, B.R., 1998. Large-scale expression measurement by hybridization methods: from high-density membranes to 'DNA chips'. J. Biochem. 124, 251–258.
[10] Lemieux, B., Aharoni, A., Schena, M., 1998. Overview of DNA chip technology. Mol. Breed. 4, 277–289.
[11] Lennon, G.G., Lehrach, H., 1991. Hybridization analyses of arrayed cDNA libraries. Trends Genet. 7, 314–317.
[12] Lockhart, D.J., Dong, H., Byrne, M.C., Follettie, M.T., Gallo, M.V., Chee, M.S., Mittmann, M., Wang, C., Kobayashi, M., Horton, H., Brown, E.L., 1996. Expression monitoring by hybridization to high-density oligonucleotide arrays. Nat. Biotechnol. 14, 1675–1680.
[13] Watson, A., Mazumder, A., Stewart, M., Balasubramanian, S., 1998. Technology for microarray analysis of gene expression. Curr. Opin. Biotechnol. 9, 609–614.
[14] Bowtell DD. Options available—from start to finish—for obtaining expression data by microarray. Nat Genet 1999;21(Suppl.):25–32.

APPLICATION IN OPTICAL FIBER COMMUNICATION

Kusum Rani[1] ,Sheetal[1]
[1]Baba Mastnath University, Rohtak- 124001, India, (vatskusum6@gmail.com), M-9034323301

ABSTRACT
This article give an overview of the advancements in silicon photonics focusing on the optical power budget and polarization requirements for applications in optical fiber communications. In this paper we discuss the different generation of optical fibre communication and its important features such as total internal reflection, different types of fibers along with their size and refractive index profile, dispersion and loss mechanism. Also,the general system of optical fiber communication is briefly mentioned along with its advantages and limitations along with its future aspects.
Keywords: Bandwidth; optical fiber; group index; group velocity; soliton v-number; dispersion.

INTRODUCTION

Now we are in the twenty first century, the era of 'Information technology. There is no doubt that information technology has had an exponential growth through the modern telecommunication systems. Particularly, optical fiber communication plays a vital role in the development of high quality and high-speed telecommunication systems. Optical fiber technology was considered to be a major driver behind the information technology revolution and the huge progress on global telecommunications that has been witnessed in recent years. Fiber optic telecommunication is now taken for granted in view of its wide-ranging application as the most suitable singular transmission medium for voice, video, and data signals. There has also been a resurgence of interest amongst researchers to design and fabricate an exotic class of special fibers - fibers in which transmission losses of the material would not be a limiting factor while nonlinearity and dispersion characteristics could be conveniently tailored to achieve certain application-specific fibers, not necessarily for telecommunication applications only. Research targeted towards such fiber designs led to the emergence of a new class of fibers, broadly referred to as microstructured optical fibers (MOF), which are characterized by wavelength-scale periodic refractive index features across its physical cross-section resulting in photonic bandgaps when appropriately designed. These features could be periodically located air holes/low refractive index regions in the cladding region, which surround the central core region, which could be of higher or lower refractive index than the average refractive index of the cladding region. Due to the large degree of design freedom and flexibility and also strong dependence on wavelength of the mode effective index, micro structured fibers have opened up a variety of new applications such as spectral broadening of a short pulse due to extreme nonlinear effects after propagating through a MOF resulting in generation of super continuum light, wide band transmission, high power delivery, endlessly single mode, very large or very small mode effective area, low-loss guidance of light in an air core and so on.

OPTICAL TRANSPARENCY

Loss spectrum, dispersion and nonlinear propagation effects are the three most important propagation characteristics of any signal transmitting single-mode optical fiber in the context of modern optical telecommunication.

1) Loss spectrum

An illustrative example of the loss spectrum of a state-of-the-art commercially available conventional ITU1 recommended standard G.652 type of single-mode fiber .

2) Dispersion spectrum

Chromatic dispersion, whose very name implies that it is dependent on wavelength and whose magnitude is a measure of the information transmission capacity of a single- mode fiber, is another important transmission characteristic (along with loss) and it arises because of the dispersive nature of an optical fiber due to which the group velocity of a propagating

signal pulse becomes a function of frequency which limits the number of pulses that can be sent through the fiber per unit time

EMERGENCE OF AMPLIFIERS

Fiber amplifiers:

Maximum launched optical power into a fiber was below 100 μW , it was difficult to improve system lengths any further and use of electronic repeaters became inevitable. At a repeater, the so-called 3**R**-regeneration functions (reamplification, retiming, and reshaping) are performed in the electric domain on the incoming attenuated as well as distorted (due to dispersion) signals after detection by a photo-detector and before the revamped signals are fed to a laser diode drive circuits, wherefrom these cleaned optical pulses are re-injected in to next section of the fiber link

FIBERS FOR METRO NETWORKS

During the IT bubble burst, there has been a slowing down of business in optical communication due to the so-called huge fiber glut in the long haul networks (typically trans-oceanic). However, the gap between the demand and supply of bandwidth has beenmuch less in the metro sector and in recent years metro optical networks have attracted a great deal of attention due to potentials for high growth. A metro network provides generalized telecommunication services transporting any kind of signal from one point to another in a metro, usually running a couple of hundred kilometers in length. In transport, DWDM is the key enabling technology to expand the capacity of existing and new fiber cables without optical-to-electrical-to-optical conversions.

MICROSTRUCTURED OPTICAL FIBERS (MOF)-DISCOVERY

The state-of-the art in silica-based optical fiber technology could be described as

• Loss close to theoretical limit (0.14 dB/km)

• Dispersion could be tailored close to zero anywhere at a wavelength $\geq$ 1310 nm but not below 1200 nm unless fiber core is significantly reduced through tapering, for example [Birks et al, 2000]

• Minimum nonlinear impairments over distances $\geq$ 100 km

• High quality fiber amplifiers with low noise to compensate for whatever be the transmission loss; noise figure could be close to theoretical minimum of 3 dB

• Demonstration of hero experiments at transmission rates > terabit/s over a single fiber through modulation techniques like CSRZ-DQPSK and polarization mode division multiplexing.

With so much of development it appeared for a while that there was no further research scope for development of newer fibers.

CONCLUSION

In this chapter we have attempted to provide a unified summary description of the most important propagation characteristics of an optical fiber followed by discussion on several variety of special fibers for realizing fiber amplifiers, dispersion compensating fibers, microstructured optical fibers, and so on. Even though huge progress has been made on development of optical fibers for telecom application, a need for developing special fibers, not necessarily for telecom alone, has arisen. This chapter was an effort to describe some of these special fibers. Detailed discussions are given on our own work related to inherently gain-flattened EDFA, DCFs of large mode effective area, index-guided MOF and Bragg fibers for realizing dispersion compensation, for metro network centric applications, and for generating super continuum light.

REFERENCES

[1]Agrawal, G. P. (2007), *Nonlinear Fiber Optics*, Fourth edition, Academic Press, San Diego.

[2]Agrawal, G. P. (2006a), Fiber optic Raman amplifiers in *Guided Wave Optical Components and Devices: basics, Technology, and Applications*, B. P. Pal (Ed.), pp. 1-25, Elsevier Academic Press, Burlington & San Diego.

[3]Argyros, A., Eijkelenborg, M. V., Large, M. and Basset, I. (2006), Hollow core microstructure polymer optical fiber, *Opt. Lett*., Vol. 31, pp. 172-174.
[4]Auguste, J. L., Jindal, R., Blondy, J. M., Clapeau, M., Marcou, J., Dussardier, B., Monnom, G.,Ostrowsky, D. B., Pal, B. P. and Thyagarajan K. (2000), -1800 ps/(nm.km) chromatic dispersion at 1.55 μm in a dual-core fiber, *Electron. Lett*. Vol. 36, pp. 1689-1691.
[5]Birks, T. A., Knight, J. C., and Russel, P. St. J. (1997), Endlessly single-mode photonic crystal fiber, *Opt. Lett*., Vol. 22, pp. 961-963.
[6] Birks, T. A., Wadsworth W. A., and Russel, P. St. J. (2000) Supercontinuum generation in tapered fibers, *Opt. Lett*. Vol. 25, pp. 1415-1417.
[7] Bookey H. T., Dasgupta, S., Bezawada, N., Pal, B. P., Sysoliatin, A., McCarthy, J. E., Salganskii, M., Khopin, V., and Kar, A. K. (2009), Experimental demonstration of spectral broadening in an all-silica Bragg fiber, *Opt. Exp*. Vol. 17, pp. 17130-17135.
[8] Brechet, F., Roy, P., Marcou, J. and Pagnoux, D. (2000), Single-mode propagation in to depressed-core-index photonic bandgap fiber designed for zero-dispersion propagation at short wavelength, *Electron. Lett*., Vol. 36, pp. 514-515.
[9] Bromage, J. (2004), Raman amplification for fiber communication systems, *IEEE J. Lightwave Tech*., Vol. 22, pp. 79-93.
[10]Cox, F.M., Argyros, A. and Large, M .C. J. (2006), Liquid-filled hollow core microstructured polymer optical fiber, *Opt. Exp*., Vol. 14, pp. 4135-4140.
[11] Dasgupta, S., Pal, B. P. and Shenoy, M. R. (2005), Design of dispersion compensating Bragg fiber with ultrahigh figure of merit, *Opt. Lett*., Vol. 30, pp. 1917-1919.
[12] Dasgupta, S. (2006), Personal Communication.
[13] Dasgupta, S., Pal, B. P. and Shenoy, M. R. (2006), Photonic bandgap guided Bragg fibers in *Guided Wave Optical Components and Devices: basics, Technology, and Applications*, B. P. Pal (Ed.), pp. 1-25, Elsevier Academic Press, Burlington & San Diego.

DIFFERENTIAL EQUATION APPROACH FOR LAPLACE TRANSFORM AND LAPLACE INVERSE

G C SHUKLA

Department of Mathematics, Baba Mastnath University, Asthal Bohar, Rohtak (Haryana),
drgcshukla@gmail.com

Abstract

We have shwn that one way to solve laplace transform and is inverse is taking recource to difference equation approach. one easly works out alternative by using integral equation approach of lapce but different has more elegance as will be dissuce by working it into detail.

Introducation

Let f(t) be a function of t defined for all positive values of t , then the Laplace transform of , f(t) , denoted by f(s) is defined as $Lf(t)= \int_{\infty}^{0} \exp(-st)\, f(t)dt$ (1) , provided that the integral exists , S is a parameter real or complex number (Re S>0) .

Further the Laplace transformation of derivatives of f(t) , of various orders are given by ,

$$\left.\begin{aligned} &L[f'(t)] = s\, f(s) - f(0) \\ &L[f''(t)] = s^2 f(s) - s\, f(0) - f'(0) \\ &\text{----------} \\ &L[f^n(t) = s^n f(s) - s^{n-1} f(0) - s^{n-2} f'(0) \ldots\ldots - f^{n-1}(0) \end{aligned}\right\} \quad (2)$$

Further , we have already proved an asymptotic expansion for L f(t) = f(s) = $\sum_{n=0}^{\infty} \frac{nth\ power\ of\ f\ evaluated\ at\ t=0}{n+1\ th\ power\ of\ s}$ (3) [2]

This yields, $f(0) = \lim_{s\to\infty} s\, f(s)$ (4)

$$f^n(0) = \lim_{s\to\infty} (n+1\ th\ power\ of\ s)\, f(s) \text{ where } n= 0,1,2.....)$$

We get Laplace transform of trigonometric, hypergeometric and exponential function by direct evaluation of integral (1) for f(t). The method for this is to evaluate integrals (1) for f(t) is either trigonometric function or hyperbolic functions or exponential function of t.

We differ from this by obtaining differential equations for f(t) and using formula (2) for derivatives. Let us do it step-wise as follows

f(t) = sin at

differentiate it two times with respect to t and we obtain differential equation

$$\left.\begin{aligned} &f''(t) + a^2 f(t) = 0 \\ &f(0) = 0 \\ &f'(0) = a \end{aligned}\right\} \quad (5)$$

Similarly we tabulate differential equation for f(t) = cos at , sin at , sinh at , cosh at , e^{at} , e^{-at} as follows :

f(t)	Differential equation	f(s)
sin at	$f''(t) + a^2 f(t) = 0$,	$\frac{a}{s^2 + a^2}$
	$f(0) = 0$, $f'(0) = a$	
cos at	$f''(t) + a^2 f(t) = 0$,	$\frac{s}{s^2 + a^2}$
	$f(0) = 1$, $f'(0) = 0$	
cosh at	$f''(t) - a^2 f(t) = 0$	$\frac{s}{s^2 - a^2}$
	$f(0) = 1$, $f'(0) = a$	
sinh at	$f''(t) - a^2 f(t) = 0$	$\frac{a}{s^2 - a^2}$
	$f(0) = 0$, $f'(0) = a$	
e^{at}	$f'(t) - a f(t) = 0$,	$\frac{1}{s - a}$
	$f(0) = 1$	
e^{-at}	$f'(t) - a f(t) = 0$	$\frac{1}{s + a}$
	$f(0) = 1$	

(6)

We can use method of differentiation and obtain differential equation for f(t) , whose Laplace transform we have to consider

Laplace Transform

if $L\, f(t) = f(s)$,

then $L^{-1} f(s) = f(t)$ (7)

There are various formulas to work out $L^{-1} f(s)$. We stress differential equation approach developed by us, called convolution differential equation approach, by differentiating convolution integral of Laplace transformation of functions, namely

Convolution Integral for Laplace transformation:

$$L^{-1} f(s)\, g(s) = \int_0^t f(u) g(t-u)\, du = \int_0^t g(u) f(t-u)\, du \quad (8)$$

Call it I(t) . We have just differentiated thus I(t) with respect to t under integral sign and obtain a differential equations which we term convolution differential equation

Nearly 80% Laplace inverse can easily be done with this method. We illustrate as how to obtain convolution differential equation with as examples,

Ex ample 1 $L^{-1}\ s/(s^2 + a^2)$ ⊣ (9)

Let $f(s) = 1/(s^2 + a^2)$, $g(s) = s/(s^2 + a^2)$,

then $f(t) = L^{-1}\ f(s) = \sin at / a$, $g(t) = L^{-1}\ g(s) = \cos at$

The convolution integral for above problem is

$$L^{-1}\ s/(s^2 + a^2) = \int_0^t \sin \frac{au}{a}\ \cos a(t-u)du \quad \} \qquad (10)$$

$$= I(t)$$

Now, differentiate I(t) with respect to t, under integral sign, we get

$$\frac{d\,I(t)}{dt} = \frac{\sin at}{a} - \int_0^t \sin au \sin(at - \cos u)\,du \quad \} \qquad (11)$$

$I(0) = 0$

Again differentiate w. r. t t,

$$\frac{d^2}{dt^2} I(t) = \cos at - a^2 \int_0^t \sin \frac{au}{a} \cos a(t-u)du \quad \} \qquad (12)$$

This yields us desired convolutional differential equation

$$I''(t) + a^2 I(t) = \cos at \quad \dashv \qquad (13)$$

Complementary function or this is $I(t) = A \cos at + B \sin at$ (14)

$I(0) = 0 \implies A = 0$

$I'(0) = 0 \implies B = 0$

Thus Particular integral: $I(t) = \dfrac{\cos at}{D^2 + a^2} = t/2a \sin at$ (15)

Example 2 $L^{-1}\ 1/(s^2-5st+6) = L^{-1}\ f(s) = f(t)$, say (16)

We can solve it as differential equation developed in Example 1, but, we give another method to work out differential equation

We have $L\left\{\dfrac{d^2}{dt^2} f(t)\right\} = s^2 f(s) - s\,f(0) - f'(0)$ (17)

dt^2

For the above f(s) , we get f(0) = 0 , f'(0) = 1

$$\text{Thus } L\left\{ \frac{d^2}{dt^2} f(t) \right\} = \frac{s^2}{s^2-5s+6} \quad (18)$$

$$\frac{d^2f(t)}{dt^2} \text{ here } = e\, L^{-1} \text{ is } \left\{ \frac{s^2}{s^2-5s+6} - 1 \right\}$$

$$= L^{-1}\left\{ \frac{5s-6}{(s-2)(s-3)} \right\}$$

$$= L^{-1}\left\{ \frac{A}{S-2} + \frac{B}{S+3} \right\} \quad (19)$$

The method of partial fraction yields A= 9 , B = -4
Now differential equation of above problem becomes

$$d^2t/dt^2 = Ae^{3t} + Be^{2t} \quad (20)$$

Integrals twice and put values of A & B to get the answers

We get $f(t) = Ae^{3t}/9 + Be^{2t/4}$ or $e^{3t} - e^{2t}$ (21)

Note : 1) Differentiate equation for $L^{-1}f(s)$ can be done by the two methods which we have just exploited
2) There may be many differential equations but identical f(t) as $L^{-1}f(s)$ is unique

REFERENCES

[1] Any standard book on integral transform or Laplace transform

[2] Paper presented by present author for $L^{-1}f(s) = \sum_{n=0}^{\infty} \frac{nth\ power\ of\ f(0)}{n+1\ th\ power\ of\ s}$ in national conference of maths in ITM Gurgaon 2009

[3] Laplace transform and fourier transform book by G C Shukla with Pragati prakashan for publication.

BANACH SPACE AND IT'S BASIC THEORIES

Minakshi

Baba Mastnath University, Asthal Bohar, Rohtak-124021,E-mail: mann80m@gmail.com

ABSTRACT

In mathematics, more specifically in functional analysis, a ***Banach space*** *(pronounced [banax]) is a complete normed vector space. Thus, a Banach space is a vector space with a metric that allows the computation of vector length and distance between vectors and is complete in the sense that a Cauchy sequence of vectors always converges to a well defined limit that is within the space.*

Banach spaces are named after the Polish mathematician Stefan Banach, who introduced and made a systematic study of them in 1920–1922 along with Hans Hahn and Eduard Helly. Banach spaces originally grew out of the study of function spaces by Hilbert, Fréchet, and Riesz earlier in the century. Banach spaces play a central role in functional analysis. In other areas of analysis, the spaces under study are often Banach spaces.

INTRODUCTION

A Banach space is a vector space X over the field **R** of real numbers, or over the field **C** of complex numbers, which is equipped with a norm and which is complete with respect to that norm, that is to say, for every Cauchy sequence $\{x_n\}$ in X, there exists an element x in X such that

$$\lim_{n\to\infty} x_n = x,$$

or equivalently:

$$\lim_{n\to\infty} \|x_n - x\|_X = 0.$$

The vector space structure allows one to relate the behavior of Cauchy sequences to that of converging series of vectors. A normed space X is a Banach space if and only if each absolutely convergent series in X converges,

$$\sum_{n=1}^{\infty} \|v_n\|_X < \infty \quad \text{implies that} \quad \sum_{n=1}^{\infty} v_n \ \text{ converges in } \ X.$$

Completeness of a normed space is preserved if the given norm is replaced by an equivalent one.

All norms on a finite-dimensional vector space are equivalent. Every finite-dimensional normed space over **R** or **C** is a Banach space.

GENERAL THEORY

Linear operators, isomorphisms

If X and Y are normed spaces over the same ground field **K**, the set of all continuous **K**-linear maps $T : X \to Y$ is denoted by $B(X, Y)$. In infinite-dimensional spaces, not all linear maps are continuous. A linear mapping from a normed space X to another normed space is continuous if and only if it is bounded on the closed unit ball of X. Thus, the vector space $B(X, Y)$ can be given the operator norm

$$\|T\| = \sup\{\|Tx\|_Y \mid x \in X,\ \|x\|_X \leq 1\}.$$

For Y a Banach space, the space $B(X, Y)$ is a Banach space with respect to this norm.

If X is a Banach space, the space $B(X) = B(X, X)$ forms a unital Banach algebra; the multiplication operation is given by the composition of linear maps.

If X and Y are normed spaces, they are **isomorphic normed spaces** if there exists a linear bijection $T : X \to Y$ such that T and its inverse T^{-1} are continuous. If one of the two spaces X or Y is complete (or reflexive, separable, etc.) then so is the other space. Two normed spaces X and Y are **isometrically isomorphic** if in addition, T is an isometry, i.e., $\|T(x)\| = \|x\|$ for

every x in X. The Banach-Mazur distance $d(X, Y)$ between two isomorphic but not isometric spaces X and Y gives a measure of how much the two spaces X and Y differ.

BASIC NOTIONS

Every normed space X can be isometrically embedded in a Banach space. More precisely, there is a Banach space Y and an isometric mapping $T : X \to Y$ such that $T(X)$ is dense in Y. If Z is another Banach space such that there is an isometric isomorphism from X onto a dense subset of Z, then Z is isometrically isomorphic to Y.

This Banach space Y is the completion of the normed space X. The underlying metric space for Y is the same as the metric completion of X, with the vector space operations extended from X to Y. The completion of X is often denoted by $\widehat{X}$.

The cartesian product $X \times Y$ of two normed spaces is not canonically equipped with a norm. However, several equivalent norms are commonly used, such as

$$\|(x,y)\|_1 = \|x\| + \|y\|, \qquad \|(x,y)\|_\infty = \max(\|x\|, \|y\|)$$

and give rise to isomorphic normed spaces. In this sense, the product $X \times Y$ (or the direct sum $X \oplus Y$) is complete if and only if the two factors are complete.

If M is a closed linear subspace of a normed space X, there is a natural norm on the **quotient space** X / M,

$$\|x + M\| = \inf_{m \in M} \|x + m\|.$$

The quotient X / M is a Banach space when X is complete. The **quotient map** from X onto X / M, sending x in X to its class $x + M$, is linear, onto and has norm 1, except when $M = X$, in which case the quotient is the null space.

The closed linear subspace M of X is said to be a **complemented subspace** of X if M is the range of a bounded linear projection P from X onto M. In this case, the space X is isomorphic to the direct sum of M and $\mathrm{Ker}(P)$, the kernel of the projection P.

Suppose that X and Y are Banach spaces and that $T \in B(X, Y)$. There exists a **canonical factorization** of T as

$$T = T_1 \circ \pi, \quad T : X \xrightarrow{\pi} X/\operatorname{Ker}(T) \xrightarrow{T_1} Y$$

where the first map π is the quotient map, and the second map T_1 sends every class $x + \mathrm{Ker}(T)$ in the quotient to the image $T(x)$ in Y. This is well defined because all elements in the same class have the same image. The mapping T_1 is a linear bijection from $X / \mathrm{Ker}(T)$ onto the range $T(X)$, whose inverse need not be bounded.

CLASSICAL SPACES

Basic examples of Banach spaces include: the L^p spaces and their special cases, the sequence spaces ℓ^p that consist of scalar sequences indexed by $\mathbf{N}$; among them, the space ℓ^1 of absolutely summable sequences and the space ℓ^2 of square summable sequences; the space c_0 of sequences tending to zero and the space ℓ^∞ of bounded sequences; the space $C(K)$ of continuous scalar functions on a compact Hausdorff space K, equipped with the max norm,

$$\|f\|_{C(K)} = \max\{|f(x)| : x \in K\}, \quad f \in C(K).$$

According to the Banach–Mazur theorem, every Banach space is isometrically isomorphic to a subspace of some $C(K)$. For every separable Banach space X, there is a closed subspace M of ℓ^1 such that $X \cong \ell^1/M$.

Any Hilbert space serves as an example of a Banach space. A Hilbert space H on $\mathbf{K} = \mathbf{R}$, $\mathbf{C}$ is complete for a norm of the form

$$\|x\|_H = \sqrt{\langle x, x \rangle},$$

where

$$\langle \cdot, \cdot \rangle : H \times H \to \mathbf{K}$$

is the inner product, linear in its first argument that satisfies the following:

$$\forall x, y \in H: \quad \langle y, x \rangle = \overline{\langle x, y \rangle},$$
$$\forall x \in H: \quad \langle x, x \rangle \geq 0,$$
$$\langle x, x \rangle = 0 \Leftrightarrow x = 0.$$

For example, the space L^2 is a Hilbert space.

The Hardy spaces, the Sobolev spaces are examples of Banach spaces that are related to L^p spaces and have additional structure. They are important in different branches of analysis, Harmonic analysis and Partial differential equations among others.

BANACH ALGEBRAS

A Banach algebra is a Banach space A over $\mathbf{K} = \mathbf{R}$ or $\mathbf{C}$, together with a structure of algebra over $\mathbf{K}$, such that the product map $(a, b) \in A \times A \to ab \in A$ is continuous. An equivalent norm on A can be found so that $\|ab\| \leq \|a\| \|b\|$ for all $a, b \in A$.

Examples

- The Banach space $C(K)$, with the pointwise product, is a Banach algebra.
- The disk algebra $A(\mathbf{D})$ consists of functions holomorphic in the open unit disk $\mathbf{D} \subset \mathbf{C}$ and continuous on its closure: $\mathbf{D}$. Equipped with the max norm on $\mathbf{D}$, the disk algebra $A(\mathbf{D})$ is a closed subalgebra of $C(\mathbf{D})$.
- The Wiener algebra $A(\mathbf{T})$ is the algebra of functions on the unit circle $\mathbf{T}$ with absolutely convergent Fourier series. Via the map associating a function on $\mathbf{T}$ to the sequence of its Fourier coefficients, this algebra is isomorphic to the Banach algebra $\ell^1(\mathbf{Z})$, where the product is the convolution of sequences.
- For every Banach space X, the space $B(X)$ of bounded linear operators on X, with the composition of maps as product, is a Banach algebra.

- A C*-algebra is a complex Banach algebra A with an antilinear involution $a \to a^*$ such that $\|a^*a\| = \|a\|^2$. The space $B(H)$ of bounded linear operators on a Hilbert space H is a fundamental example of C*-algebra. The Gelfand–Naimark theorem states that every C*-algebra is isometrically isomorphic to a C*-subalgebra of some $B(H)$. The space $C(K)$ of complex continuous functions on a compact Hausdorff space K is an example of commutative C*-algebra, where the involution associates to every function f its complex conjugate $\overline{f}$.

DUAL SPACE

If X is a normed space and $\mathbf{K}$ the underlying field (either the real or the complex numbers), the continuous dual space is the space of continuous linear maps from X into $\mathbf{K}$, or **continuous linear functionals**. The notation for the continuous dual is $X' = B(X, \mathbf{K})$ in this article.Since $\mathbf{K}$ is a Banach space (using the absolute value as norm), the dual X' is a Banach space, for every normed space X.

The main tool for proving the existence of continuous linear functionals is the Hahn–Banach theorem.

Hahn–Banach theorem. Let X be a vector space over the field $\mathbf{K} = \mathbf{R}, \mathbf{C}$. Let further

- $Y \subseteq X$ be a linear subspace,
- $p : X \to \mathbf{R}$ be a sublinear function and
- $f : Y \to \mathbf{K}$ be a linear functional so that $\mathrm{Re}(f(y)) \leq p(y)$ for all y in Y.

Then, there exists a linear functional $F : X \to \mathbf{K}$ so that

$$F|_Y = f, \quad \text{and} \quad \forall x \in X, \quad \mathrm{Re}(F(x)) \leq p(x).$$

In particular, every continuous linear functional on a subspace of a normed space can be continuously extended to the whole space, without increasing the norm of the functional. An important special case is the following: for every vector x in a normed space X, there exists a continuous linear functional f on X such that

$$f(x) = \|x\|_X, \quad \|f\|_{X'} \leq 1.$$

When x is not equal to the **0** vector, the functional f must have norm one, and is called a **norming functional** for x.

The Hahn–Banach separation theorem states that two disjoint non-empty convex sets in a real Banach space, one of them open, can be separated by a closed affine hyperplane. The open convex set lies strictly on one side of the hyperplane, the second convex set lies on the other side but may touch the hyperplane.

A subset S in a Banach space X is **total** if the linear span of S is dense in X. The subset S is total in X if and only if the only continuous linear functional that vanishes on S is the **0** functional: this equivalence follows from the Hahn–Banach theorem.

If X is the direct sum of two closed linear subspaces M and N, then the dual X' of X is isomorphic to the direct sum of the duals of M and N. If M is a closed linear subspace in X, one can associate the *orthogonal of M* in the dual,

$$M^{\perp} = \{x' \in X' : x'(m) = 0, \ \forall m \in M\}.$$

The orthogonal $M^{\perp}$ is a closed linear subspace of the dual. The dual of M is isometrically isomorphic to $X'/M^{\perp}$. The dual of X/M is isometrically isomorphic to $M^{\perp}$.

The dual of a separable Banach space need not be separable, but:

Theorem. Let X be a normed space. If X' is separable, then X is separable.

When X' is separable, the above criterion for totality can be used for proving the existence of a countable total subset in X.

Weak Topologies

The weak topology on a Banach space X is the coarsest topology on X for which all elements x' in the continuous dual space X' are continuous. The norm topology is therefore finer than the weak topology. It follows from the Hahn–Banach separation theorem that the weak topology is Hausdorff, and that a norm-closed convex subset of a Banach space is also weakly closed. A norm-continuous linear map between two Banach spaces X and Y is also **weakly continuous**, i.e., continuous from the weak topology of X to that of Y.

If X is infinite-dimensional, there exist linear maps which are not continuous. The space X^* of all linear maps from X to the underlying field **K** (this space X^* is called the algebraic dual space, to distinguish it from X') also induces a topology on X which is finer than the weak topology, and much less used in functional analysis.

On a dual space X', there is a topology weaker than the weak topology of X', called weak* topology. It is the coarsest topology on X' for which all evaluation maps $x' \in X' \to x'(x)$, $x \in X$, are continuous. Its importance comes from the Banach–Alaoglu theorem.

Banach–Alaoglu Theorem. Let X be a normed vector space. Then the closed unit ball $B' = \{x' \in X' : \|x'\| \leq 1\}$ of the dual space is compact in the weak* topology.

The Banach–Alaoglu theorem depends on Tychonoff's theorem about infinite products of compact spaces. When X is separable, the unit ball B' of the dual is a metrizable compact in the weak* topology.

Examples of dual spaces

The dual of c_0 is isometrically isomorphic to ℓ^1: for every bounded linear functional f on c_0, there is a unique element $y = \{y_n\} \in \ell^1$ such that

$$f(x) = \sum_{n \in \mathbf{N}} x_n y_n, \qquad x = \{x_n\} \in c_0, \quad \text{and} \quad \|f\|_{(c_0)'} = \|y\|_{\ell_1}.$$

The dual of ℓ^1 is isometrically isomorphic to ℓ^{∞}. The dual of $L^p([0, 1])$ is isometrically isomorphic to $L^q([0, 1])$ when $1 \leq p < \infty$ and $1/p + 1/q = 1$.

For every vector y in a Hilbert space H, the mapping

$$x \in H \to f_y(x) = \langle x, y \rangle$$

defines a continuous linear functional f_y on H. The Riesz representation theorem states that every continuous linear functional on H is of the form f_y for a uniquely defined vector y in H. The mapping $y \in H \to f_y$ is an antilinear isometric bijection from H onto its dual H'. When the scalars are real, this map is an isometric isomorphism.

When K is a compact Hausdorff topological space, the dual $M(K)$ of $C(K)$ is the space of Radon measures in the sense of Bourbaki. The subset $P(K)$ of $M(K)$ consisting of non-negative measures of mass 1 (probability measures) is a convex w*-closed subset of the unit ball of $M(K)$. The extreme points of $P(K)$ are the Dirac measures on K. The set of Dirac measures on K, equipped with the w*-topology, is homeomorphic to K.

> Banach-Stone Theorem. If K and L are compact Hausdorff spaces and if $C(K)$ and $C(L)$ are isometrically isomorphic, then the topological spaces K and L are homeomorphic.

The result has been extended by Amir and Camber to the case when the multiplicative Banach–Mazur distance between $C(K)$ and $C(L)$ is < 2. The theorem is no longer true when the distance is = 2.

In the commutative Banach algebra $C(K)$, the maximal ideals are precisely kernels of Dirac mesures on K,

$$I_x = \ker \delta_x = \{f \in C(K) : f(x) = 0\}, \quad x \in K.$$

More generally, by the Gelfand-Mazur theorem, the maximal ideals of a unital commutative Banach algebra can be identified with its characters---not merely as sets but as topological spaces: the former with the hull-kernel topology and the latter with the w*-topology. In this identification, the maximal ideal space can be viewed as a w*-compact subset of the unit ball in the dual A'.

> **Theorem.** If K is a compact Hausdorff space, then the maximal ideal space Ξ of the Banach algebra $C(K)$ is homeomorphic to K.

Not every unital commutative Banach algebra is of the form $C(K)$ for some compact Hausdorff space K. However, this statement holds if one places $C(K)$ in the smaller category of commutative C*-algebras. Gelfand's representation theorem for commutative C*-algebras states that every commutative unital C*-algebra A is isometrically isomorphic to a $C(K)$ space. The Hausdorff compact space K here is again the maximal ideal space, also called the spectrum of A in the C*-algebra context.

Bidual

If X is a normed space, the (continuous) dual X'' of the dual X' is called **bidual**, or **second dual** of X. For every normed space X, there is a natural map,

$$\begin{cases} F_X : X \to X'' \\ F_X(x)(f) = f(x) \quad \forall x \in X, \forall f \in X' \end{cases}$$

This defines $F_X(x)$ as a continuous linear functional on X', i.e., an element of X''. The map $F_X : x \to F_X(x)$ is a linear map from X to X''. As a consequence of the existence of a norming functional f for every x in X, this map F_X is isometric, thus injective.

For example, the dual of $X = c_0$ is identified with ℓ^1, and the dual of ℓ^1 is identified with ℓ^∞, the space of bounded scalar sequences. Under these identifications, F_X is the inclusion map from c_0 to ℓ^∞. It is indeed isometric, but not onto.

If F_X is surjective, then the normed space X is called **reflexive** (see below). Being the dual of a normed space, the bidual X'' is complete, therefore, every reflexive normed space is a Banach space.

Using the isometric embedding F_X, it is customary to consider a normed space X as a subset of its bidual. When X is a Banach space, it is viewed as a closed linear subspace of X''. If X is

not reflexive, the unit ball of X is a proper subset of the unit ball of X''. The Goldstine theorem states that the unit ball of a normed space is weakly*-dense in the unit ball of the bidual. In other words, for every x'' in the bidual, there exists a net $\{x_j\}$ in X so that

$$\sup_j \|x_j\| \leq \|x''\|, \quad x''(f) = \lim_j f(x_j), \quad f \in X'.$$

The net may be replaced by a weakly*-convergent sequence when the dual X' is separable. On the other hand, elements of the bidual of ℓ^1 that are not in ℓ^1 cannot be weak*-limit of *sequences* in ℓ^1, since ℓ^1 is weakly sequentially complete.

BANACH'S THEOREMS

Here are the main general results about Banach spaces that go back to the time of Banach's book (Banach (1932)) and are related to the Baire category theorem. According to this theorem, a complete metric space (such as a Banach space, a Fréchet space or an F-space) cannot be equal to a union of countably many closed subsets with empty interiors. Therefore, a Banach space cannot be the union of countably many closed subspaces, unless it is already equal to one of them; a Banach space with a countable Hamel basis is finite-dimensional.

Banach–Steinhaus Theorem. Let X be a Banach space and Y be a normed vector space. Suppose that F is a collection of continuous linear operators from X to Y. The uniform boundedness principle states that if for all x in X we have $\sup_{T\in F} \|T(x)\|_Y < \infty$, then $\sup_{T\in F} \|T\|_Y < \infty$.

The Banach–Steinhaus theorem is not limited to Banach spaces. It can be extended for example to the case where X is a Fréchet space, provided the conclusion is modified as follows: under the same hypothesis, there exists a neighborhood U of $\mathbf{0}$ in X such that all T in F are uniformly bounded on U,

$$\sup_{T\in F} \sup_{x\in U} \|T(x)\|_Y < \infty.$$

The Open Mapping Theorem. Let X and Y be Banach spaces and $T: X \to Y$ be a continuous linear operator. Then T is surjective if and only if T is an open map.

Corollary. Every one-to-one bounded linear operator from a Banach space onto a Banach space is an isomorphism.

The First Isomorphism Theorem for Banach spaces. Suppose that X and Y are Banach spaces and that $T \in B(X, Y)$. Suppose further that the range of T is closed in Y. Then $X/\mathrm{Ker}(T)$ is isomorphic to $T(X)$.

This result is a direct consequence of the preceding *Banach isomorphism theorem* and of the canonical factorization of bounded linear maps.

Corollary. If a Banach space X is the internal direct sum of closed subspaces M_1, ..., M_n, then X is isomorphic to $M_1 \oplus ... \oplus M_n$.

This is another consequence of Banach's isomorphism theorem, applied to the continuous bijection from $M_1 \oplus ... \oplus M_n$ onto X sending $(m_1, ..., m_n)$ to the sum $m_1 + ... + m_n$.

The Closed Graph Theorem. Let $T: X \to Y$ be a linear mapping between Banach spaces. The graph of T is closed in $X \times Y$ if and only if T is continuous.

REFLEXIVITY

The normed space X is called reflexive when the natural map

$$\begin{cases} F_X : X \to X'' \\ F_X(x)(f) = f(x) \quad \forall x \in X, \forall f \in X' \end{cases}$$

is surjective. Reflexive normed spaces are Banach spaces.

Theorem. If X is a reflexive Banach space, every closed subspace of X and every quotient space of X are reflexive.

This is a consequence of the Hahn–Banach theorem. Further, by the open mapping theorem, if there is a bounded linear operator from the Banach space X onto the Banach space Y, then Y is reflexive.

Theorem. If X is a Banach space, then X is reflexive if and only if X' is reflexive.

Corollary. Let X be a reflexive Banach space. Then X is separable if and only if X' is separable.

Indeed, if the dual Y' of a Banach space Y is separable, then Y is separable. If X is reflexive and separable, then the dual of X' is separable, so X' is separable.

Theorem. Suppose that X_1, ..., X_n are normed spaces and that $X = X_1 \oplus \ldots \oplus X_n$. Then X is reflexive if and only if each X_j is reflexive.

Hilbert spaces are reflexive. The L^p spaces are reflexive when $1 < p < \infty$. More generally, uniformly convex spaces are reflexive, by the Milman–Pettis theorem. The spaces c_0, ℓ^1, $L^1([0, 1])$, $C([0, 1])$ are not reflexive. In these examples of non-reflexive spaces X, the bidual X'' is "much larger" than X. Namely, under the natural isometric embedding of X into X'' given by the Hahn–Banach theorem, the quotient X''/X is infinite-dimensional, and even nonseparable. However, Robert C. James has constructed

an example of a non-reflexive space, usually called "*the James space*" and denoted by J, such that the quotient J''/J is one-dimensional. Furthermore, this space J is isometrically isomorphic to its bidual.

Theorem. A Banach space X is reflexive if and only if its unit ball is compact in the weak topology.

When X is reflexive, it follows that all closed and bounded convex subsets of X are weakly compact. In a Hilbert space H, the weak compactness of the unit ball is very often used in the following way: every bounded sequence in H has weakly convergent subsequences.

Weak compactness of the unit ball provides a tool for finding solutions in reflexive spaces to certain optimization problems. For example, every convex continuous function on the unit ball B of a reflexive space attains its minimum at some point in B.

As a special case of the preceding result, when X is a reflexive space over $\mathbf{R}$, every continuous linear functional f in X' attains its maximum $\|f\|$ on the unit ball of X. The following theorem of Robert C. James provides a converse statement.

James' Theorem. For a Banach space the following two properties are equivalent:

- X is reflexive.
- for all f in X' there exists x in X with $\|x\| \leq 1$, so that $f(x) = \|f\|$.

The theorem can be extended to give a characterization of weakly compact convex sets.

On every non-reflexive Banach space X, there exist continuous linear functionals that are not *norm-attaining*. However, the Bishop–Phelps theorem states that norm-attaining functionals are norm dense in the dual X' of X.

Weak convergences of sequences

A sequence $\{x_n\}$ in a Banach space X is **weakly convergent** to a vector $x \in X$ if $f(x_n)$ converges to $f(x)$ for every continuous linear functional f in the dual X'. The sequence $\{x_n\}$ is a **weakly Cauchy sequence** if $f(x_n)$ converges to a scalar limit $L(f)$, for every f in X'. A sequence $\{f_n\}$ in the dual X' is **weakly* convergent** to a functional $f \in X'$ if $f_n(x)$ converges to $f(x)$ for every x in X. Weakly Cauchy sequences, weakly convergent and weakly* convergent sequences are norm bounded, as a consequence of the Banach–Steinhaus theorem.

When the sequence $\{x_n\}$ in X is a weakly Cauchy sequence, the limit L above defines a bounded linear functional on the dual X', i.e., an element L of the bidual of X, and L is the limit of $\{x_n\}$ in the weak*-topology of the bidual. The Banach space X is **weakly**

sequentially complete if every weakly Cauchy sequence is weakly convergent in X. It follows from the preceding discussion that reflexive spaces are weakly sequentially complete.

Theorem. For every measure μ, the space $L^1(\mu)$ is weakly sequentially complete.

An orthonormal sequence in a Hilbert space is a simple example of a weakly convergent sequence, with limit equal to the **0** vector. The unit vector basis of ℓ^p, $1 < p < \infty$, or of c_0, is another example of a **weakly null sequence**, i.e., a sequence that converges weakly to **0**. For every weakly null sequence in a Banach space, there exists a sequence of convex combinations of vectors from the given sequence that is norm-converging to **0**.

The unit vector basis of ℓ^1 is not weakly Cauchy. Weakly Cauchy sequences in ℓ^1 are weakly convergent, since L^1-spaces are weakly sequentially complete. Actually, weakly convergent sequences in ℓ^1 are norm convergent. This means that ℓ^1 satisfies Schur's property.

Results involving the ℓ^1 basis

Weakly Cauchy sequences and the ℓ^1 basis are the opposite cases of the dichotomy established in the following deep result of H. P. Rosenthal.

> **Theorem.** Let $\{x_n\}$ be a bounded sequence in a Banach space. Either $\{x_n\}$ has a weakly Cauchy subsequence, or it admits a subsequence equivalent to the standard unit vector basis of ℓ^1.

A complement to this result is due to Odell and Rosenthal (1975).

Theorem. Let X be a separable Banach space. The following are equivalent:

- The space X does not contain a closed subspace isomorphic to ℓ^1.
- Every element of the bidual X'' is the weak*-limit of a sequence $\{x_n\}$ in X.

By the Goldstine theorem, every element of the unit ball B'' of X'' is weak*-limit of a net in the unit ball of X. When X does not contain ℓ^1, every element of B'' is weak*-limit of a *sequence* in the unit ball of X.

When the Banach space X is separable, the unit ball of the dual X', equipped with the weak*-topology, is a metrizable compact space K, and every element x'' in the bidual X'' defines a bounded function on K:

$$x' \in K \mapsto x''(x'), \quad |x''(x')| \leq \|x''\|.$$

This function is continuous for the compact topology of K if and only if x'' is actually in X, considered as subset of X''. Assume in addition for the rest of the paragraph that X does not contain ℓ^1. By the preceding result of Odell and Rosenthal, the function x'' is the pointwise limit on K of a sequence $\{x_n\} \subset X$ of continuous functions on K, it is therefore a first Baire class function on K. The unit ball of the bidual is a pointwise compact subset of the first Baire class on K.

Sequences, weak and weak* compactness

When X is separable, the unit ball of the dual is weak*-compact by Banach–Alaoglu and metrizable for the weak* topology, hence every bounded sequence in the dual has weakly* convergent subsequences. This applies to separable reflexive spaces, but more is true in this case, as stated below.

The weak topology of a Banach space X is metrizable if and only if X is finite-dimensional. If the dual X' is separable, the weak topology of the unit ball of X is metrizable. This applies in particular to separable reflexive Banach spaces. Although the weak topology of the unit ball is not metrizable in general, one can characterize weak compactness using sequences.

> Eberlein–Šmulian theorem. A set A in a Banach space is relatively weakly compact if and only if every sequence $\{a_n\}$ in A has a weakly convergent subsequence.

A Banach space X is reflexive if and only if each bounded sequence in X has a weakly convergent subsequence.

A weakly compact subset A in ℓ^1 is norm-compact. Indeed, every sequence in A has weakly convergent subsequences by Eberlein–Šmulian, that are norm convergent by the Schur property of ℓ^1.

REFERENCES

1. Banach, Stefan (1932), *Théorie des opérations linéaires*, Monografie Matematyczne **1**, Warszawa: Subwencji Funduszu Kultury Narodowej, Zbl 0005.20901.
2. Beauzamy, Bernard (1985) [1982], *Introduction to Banach Spaces and their Geometry* (Second revised ed.), North-Holland.
3. Bourbaki, Nicolas (1987), *Topological vector spaces*, Elements of mathematics, Berlin: Springer-Verlag,
4. Carothers, Neal L. (2005), *A short course on Banach space theory*, London Mathematical Society Student Texts **64**, Cambridge: Cambridge University Press,
5. Diestel, Joseph (1984), *Sequences and series in Banach spaces*, Graduate Texts in Mathematics **92**, New York: Springer-Verlag,
6. Dunford, Nelson; Schwartz, Jacob T. with the assistance of W. G. Bade and R. G. Bartle (1958), *Linear Operators. I. General Theory*, Pure and Applied Mathematics **7**, New York: Interscience Publishers, Inc.,
7. Lindenstrauss, Joram; Tzafriri, Lior (1977), *Classical Banach Spaces I, Sequence Spaces*, Ergebnisse der Mathematik und ihrer Grenzgebiete **92**, Berlin: Springer-Verlag,.
8. Megginson, Robert E. (1998), *An introduction to Banach space theory*, Graduate Texts in Mathematics **183**, New York: Springer-Verlag, pp. xx+596,.
9. Ryan, Raymond A. (2002), *Introduction to Tensor Products of Banach Spaces*, Springer Monographs in Mathematics, London: Springer-Verlag, pp. xiv+225,.
10. Wojtaszczyk, Przemysław (1991), *Banach spaces for analysts*, Cambridge Studies in Advanced Mathematics **25**, Cambridge: Cambridge University Press, pp. xiv+382,.

INTRODUCTION OF METRIC SPACE

VINOD BHATIA

Baba Mastnath University, Asthal Bohar,Rohtak-124021, E-mail: bhatiavinod@ymail.com

Before we discuss topological spaces in their full generality, we will first turn our attention to a special type of topological space, a metric space. This abstraction has a huge and useful family of special cases, and it therefore deserves special attention. Also, the abstraction is picturesque and accessible; it will subsequently lead us to the full abstraction of a topological space.

. In mathematics, a **metric space** is a set for which distances between all members of the set are defined. Those distances, taken together, are called a metric on the set.

The most familiar metric space is 3-dimensional Euclidean space. In fact, a "metric" is the generalization of the Euclidean metric arising from the four long-known properties of the Euclidean distance. The Euclidean metric defines the distance between two points as the length of the straight line segment connecting them. A metric on a space induces topological properties like open and closed sets, which lead to the study of more abstract topological spaces.

DEFINITION

A **metric space** is an ordered pair (M, d) where M is a set and d is a metric on M, i.e., a function

$$d\colon M \times M \to \mathbb{R}$$

such that for any $x, y, z \in M$, the following holds:

1. $d(x,y) \geq 0$ (*non-negative*),
2. $d(x,y) = 0 \iff x = y$ (*identity of indiscernibles*),
3. $d(x,y) = d(y,x)$ (*symmetry*) and
4. $d(x,z) \leq d(x,y) + d(y,z)$ (*triangle inequality*) .

The first condition follows from the other three, since: for any $x, y \in M$,

$$\begin{aligned} & d(x,y) + d(y,x) \geq d(x,x) \text{ (by triangle inequality)} \\ \implies & d(x,y) + d(x,y) \geq d(x,x) \text{ (by symmetry)} \\ \implies & 2d(x,y) \geq 0 \text{ (by identity of indiscernibles)} \\ \implies & d(x,y) \geq 0. \end{aligned}$$

The function d is also called *distance function* or simply *distance*. Often, d is omitted and one just writes M for a metric space if it is clear from the context what metric is used.

Examples

- An important example is the discrete metric. It may be defined any non-empty set X as follows $d(x,y) = \begin{cases} 1 & \text{if } x \neq y \\ 0 & \text{if } x = y \end{cases}$
- On the set of real numbers $\mathbb{R}$, define $d(x,y) = |x - y|$ (The absolute distance between x and y).
  To prove that this is indeed a metric space, we must show that *d* is really a metric. To begin with, $d(x,y) = |x - y| \geq 0$ for any rea numbers x and y

- $d(x,y) = |x-y| = 0 \iff (x-y) = 0 \vee -(x-y) = 0 \iff x = y$
.
- $d(x,y) = |x-y| = |y-x| = d(y,x)$
- $d(x,z) = |x-z| = |x-y+y-z| = |(x-y)+(y-z)| \leq |x-y|+|y-z| = d(x,y)+d(y$

 The *positive* real numbers with distance function $d(x,y) = |\log(y/x)|$ is a complete metric space.

 Any normed vector space is a metric space by defining $d(x,y) = \|y-x\|$, see also metrics on vector spaces.(If such a space is complete, we call it a Banach space.)

 The British Rail metric (also called the Post Office metric or the SNCF metric) on a normed vector space is given by $d(x,y) = \|x\| + \|y\|$ for distinct points x and y, and $d(x,x) = 0$. More generally $\|\cdot\|$ can be replaced with a function f taking an arbitrary set S to non-negative reals and taking the value 0 at most once: then the metric is defined on S by $d(x,y) = f(x) + f(y)$ for distinct points x and y, and $d(x,x) = 0$. The name alludes to the tendency of railway journeys (or letters) to proceed via London (or Paris) irrespective of their final destination.

 If (M,d) is a metric space and X is a subset of M , then (X,d) becomes a metric space by restricting the domain of d to $X \times X$.

 The discrete metric, where $d(x,y) = 0$ if $x = y$ and $d(x,y) = 1$ otherwise, is a simple but important example, and can be applied to all non-empty sets. This, in particular, shows that for any non-empty set, there is always a metric space associated to it. Using this metric, any point is an open ball, and therefore every subset is open and the space has the discrete topology.

 A finite metric space is a metric space having a finite number of points. Not every finite metric space can be isometrically embedded in a Euclidean space.

 The hyperbolic plane is a metric space. More generally:

 If M is any connected Riemannian manifold, then we can turn M into a metric space by defining the distance of two points as the infimum of the lengths of the paths (continuously differentiable curves) connecting them.

 If X is some set and M is a metric space, then, the set of all bounded functions $f: X \to M$ (i.e. those functions whose image is a bounded subset of M) can be turned into a metric space by defining $d(f,g) = \sup_{x \in X} d(f(x), g(x))$ for any two bounded functions f and g (where $\sup$ is supremum.[4] This metric is called the uniform metric or supremum metric, and If M is complete, then this function space is complete as well. If *X* is also a topological space, then the set of all bounded continuous functions from X to M (endowed with the uniform metric), will also be a complete metric if *M* is.

 If G is an undirected connected graph, then the set V of vertices of G can be turned into a metric space by defining $d(x,y)$ to be the length of the shortest path connecting the vertices x and y. In geometric group theory this is applied to the Cayley graph of a group, yielding the word metric.

 Given a metric space (X,d) and an increasing concave function $f: [0,\infty) \to [0,\infty)$ such that $f(x) = 0$ if and only if $x = 0$, then $f \circ d$ is also a metric on X.

 Given an injective function f from any set A to a metric space (X,d), $d(f(x), f(y))$ defines a metric on A.

 Using T-theory, the tight span of a metric space is also a metric space. The tight span is useful in several types of analysis.
 The set of all m by n matrices over some field is a metric space with respect to the rank distance $d(X, Y) = \operatorname{rank}(Y - X)$.

TYPES OF METRIC SPACES

Complete spaces

A metric space M is said to be **complete** if every Cauchy sequence converges in M. That is to say: if $d(x_n, x_m) \to 0$ as both n and m independently go to infinity, then there is some $y \in M$ with $d(x_n, y) \to 0$.

Every Euclidean space is complete, as is every closed subset of a complete space. The rational numbers, using the absolute value metric $d(x, y) = |x - y|$, are not complete.

Every metric space has a unique (up to isometry) completion, which is a complete space that contains the given space as a dense subset. For example, the real numbers are the completion of the rationals.

If X is a complete subset of the metric space M, then X is closed in M. Indeed, a space is complete iff it is closed in any containing metric space

Compact spaces

A metric space *M* is compact if every sequence in *M* has a subsequence that converges to a point in *M*. This is known as sequential compactness and, in metric spaces (but not in general topological spaces), is equivalent to the topological notions of countable compactness and compactness defined via open covers.

Examples of compact metric spaces include the closed interval [0,1] with the absolute value metric, all metric spaces with finitely many points, and the Cantor set. Every closed subset of a compact space is itself compact.

A metric space is compact iff it is complete and totally bounded. This is known as the Heine–Borel theorem. Note that compactness depends only on the topology, while boundedness depends on the metric.

Lebesgue's number lemma states that for every open cover of a compact metric space *M*, there exists a "Lebesgue number" δ such that every subset of *M* of diameter < δ is contained in some member of the cover.

Connectedness

A metric space M is connected if the only subsets that are both open and closed are the empty set and M itself.

A metric space M is path connected if for any two points $x, y \in M$ there exists a continuous map $f: [0, 1] \to M$ with $f(0) = x$ and $f(1) = y$. Every path connected space is connected, but the converse is not true in general.

There are also local versions of these definitions: locally connected spaces and locally path connected spaces.

Simply connected spaces are those that, in a certain sense, do not have "holes"

Separable spaces

A metric space is separable space if it has a countable dense subset. Typical examples are the real numbers or any Euclidean space. For metric spaces (but not for general topological spaces) separability is equivalent to second countability and also to the Lindelöf property

CONVERGENCE

Definition

First, Lets translate the calculus definition of convergence, to the "language" of metric spaces: We say that a sequence x_n converges to x if for every $\epsilon > 0$ exists N that for each

$n^* > N$ the following holds: $d(x_{n^*}, x) < \epsilon$.
Equivalently, we can define converges using Open-balls: A sequence x_n converges to x If for every $\epsilon > 0$ exists N that for each $n^* > N$ the following holds: $x_{n^*} \in B_\epsilon(x)$.
The latter definition uses the "language" of open-balls, But we can do better - We can remove the ϵ from the definition of convergence, thus making the definition more topological. Let's *define* that x_n **converges** to x (and mark $x_n \to x$) , if for **every ball** B around x , exists N_B that for each $n^* > N_B$ the following holds: $x_{n^*} \in B(x)$ x is called the *limit* of the sequence.
The definitions are all the same, but the latter uses topological terms, and can be easily converted to a topological definition later.

Uniform Convergence

A sequence of functions $\{f_n\}$ is said to be uniformly convergent on a set S if for any $\epsilon > 0$, there exists an N such that when a and b are both greater than N, then $d(f_a(x), f_b(x)) < \epsilon$ for any $x \in S$

Product metric spaces

If $(M_1, d_1), \ldots, (M_n, d_n)$ are metric spaces, and *N* is the Euclidean norm on R^n, then $\Big(M_1 \times \ldots \times M_n, N(d_1, \ldots, d_n)\Big)$ is a metric space, where the product metric is defined by

$$N(d_1, ..., d_n)\Big((x_1, \ldots, x_n), (y_1, \ldots, y_n)\Big) = N\Big(d_1(x_1, y_1), \ldots, d_n(x_n, y_n)\Big),$$

and the induced topology agrees with the product topology. By the equivalence of norms in finite dimensions, an equivalent metric is obtained if *N* is the taxicab norm, a p-norm, the max norm, or any other norm which is non-decreasing as the coordinates of a positive *n*-tuple increase (yielding the triangle inequality).
Similarly, a countable product of metric spaces can be obtained using the following metric

$$d(x, y) = \sum_{i=1}^{\infty} \frac{1}{2^i} \frac{d_i(x_i, y_i)}{1 + d_i(x_i, y_i)}.$$

An uncountable product of metric spaces need not be metrizable. For example, $\mathbf{R}^{\mathbf{R}}$ is not first-countable and thus isn't metrizable

Generalizations of metric spaces

- Every metric space is a uniform space in a natural manner, and every uniform space is naturally a topological space. Uniform and topological spaces can therefore be regarded as generalizations of metric spaces.
- If we consider the first definition of a metric space given above and relax the second requirement, we arrive at the concepts of a pseudometric space or a dislocated metric space. If we remove the third or fourth, we arrive at a quasimetric space, or a semimetric space.
- If the distance function takes values in the extended real number line $\mathbf{R} \cup \{+\infty\}$, but otherwise satisfies all four conditions, then it is called an *extended metric* and the corresponding space is called an ∞*-metric space*. If the distance function takes values in some (suitable) ordered set (and the triangle inequality is adjusted accordingly), then we arrive at the notion of *generalized ultrametric*.[10]
- Approach spaces are a generalization of metric spaces, based on point-to-set distances, instead of point-to-point distances.

NOTES

1. B. Choudhary (1992). *The Elements of Complex Analysis*. New Age International. p. 20. ISBN 978-81-224-0399-2.

2. Nathan Linial. *Finite Metric Spaces—Combinatorics, Geometry and Algorithms*, Proceedings of the ICM, Beijing 2002, vol. 3, pp573–586
3. *Open problems on embeddings of finite metric spaces*, edited by Jirīi Matoušek, 2007
4. Searcóid, p. 107.
5. PlanetMath: a compact metric space is second countable

REFERENCES

1. Victor Bryant, *Metric Spaces: Iteration and Application*, Cambridge University Press, 1985, ISBN 0-521-31897-1.
2. Dmitri Burago, Yu D Burago, Sergei Ivanov, *A Course in Metric Geometry*, American Mathematical Society, 2001, ISBN 0-8218-2129-6.
3. Athanase Papadopoulos, *Metric Spaces, Convexity and Nonpositive Curvature*, European Mathematical Society, First edition 2004, ISBN 978-3-03719-010-4. Second edition 2014, ISBN 978-3-03719-132-3

ARTICLE ON VERMICOMPOSITION

Veena Maurya and Shalini

Baba Mastnath University, AsthalBohar, Rohtak, (124021) Haryana.

INTRODUCTION

Vermicomposting technology is one of the best optionsavailable for the treatment of organics-rich solid wastes. The term vermicomposting is coined from the Latin word '*Vermis'*meaning to the 'worms'. Vermicomposting refers to composting or natural conversion of biodegradable garbage into high quality manure with the help of earthworms. Earthworms play a key role in soil biology; they serve as versatile natural bioreactors to harness energy and destroy soil pathogens. The worms do so by feeding voraciously on all biodegradable refuse such as leaves, paper (nonaromatic), kitchen waste, vegetable refuse etc. Earthworms have been used forwaste stabilization for many years, especially in Southeast Asian and Europeancountries.Highlighting the role of earthworms, Charles Darwin called them the unheraldedsoldiers of the soil. From then on, different experimental studies have been carriedout to study the role of earthworms in maintaining the soil fertility and also in thedegradation of the organic matter present in the soil. Different scholars have triedthe possibility of utilizing earthworms for the break down of organic wastes such asanimal wastes, vegetable wastes and municipal sludge. Earthworms maintainaerobic conditions in the mixture, ingest solids, and convert a portion of the organicmatter into worm biomass and respiration products, and expel the remainingpartially stabilized matter as discrete material (castings). Ronald and Donald(1977a) have reported that the earthworms and the microorganisms actsymbiotically to accelerate the decomposition of organic matter.

TYPES OF EARTHWORMS

Ronald and Donald (1977b) have described the six common types of earthworms found in Europe. These are:

1. The native night crawler, or Lumbricusterrestris
2. The common field worm, or Helodriluscaliginosus
3. The green worm, or Helodriluschloroticus
4. The manure worm, or Eiseniafoetida
5. The slim earthworm, or Diplocardiaverrucusa
6. The redworm or Lumbricusrubellus.

The most common types of earthworms used for vermicomposting are banded worms (Eiseniafetida) and redworms or red wigglers (Lumbricusrubellus). Often found in aged manure piles, they generally have alternating red and buff-coloredstripes. They are not to be confused with the common garden or field earthworm (Allolobophoracaliginosa) and other species.

REPRODUCTION

Earthworms that are sexually mature have a prominent band around their body, which is called as the Clitellum. This is usually visible around 8 - 12 weeks of age. During copulation, the worms will join together at the clitellum (sometimes for quite a long period of time). Reproductive material is exchanged. When the worms separate, a ring of mucus material forms at the clitellum of each worm. This process is known as copulation. Sperm from the other worm is stored in sacs. As the mucus slides over the worm, it encases the sperm and eggs inside. After slipping free from the worm, both ends seal, forming a lemon-shape cocoon approximately 3.2 mm long. Two or more baby worms will hatch from one end of the

cocoon in approximately 3 weeks. Baby worms are whitish to almost transparent and are 12 to 25 mm long. Redwormstake 4 to 6 weeks to become sexually mature.

ROLE OF EARTHWORMS IN ORGANIC MATTER RECYCLING

The role of earthworms in humificaton and breakdown of plant litter in natural soilhas been known since the time of Darwin, but their potential to stabilize the organic refuse into useful components has been known only recently. Edwards (1998)reported five earthworm species (D. veneta, E. eugeniae, P. excavatus and ***P.hawayana*** and E. fetida) to be the most potential earthworms for breakdown of organic refuse. Generally most organic wastes can be broken down as such, except for those, which might need some degree of pretreatment prior to feeding. Earthworms are highly adaptable to different types of organic waste, provided, the physical structure, pH and the salt concentration are not above the tolerance level (Seenappa, et al., 1995). In most of the cases, the feedstock is thermophilicallycomposted in windrows (turned twice weekly), for 15 to 30 days before being fed to earthworms.

VERMITECHNOLOGY AND VERMICULTURE

Vermitechnology represents a relatively new and environmentally sound approach in the management of Municipal refuse (Loehr et al., 1988). Earthworms have the potential to be used in Vermitechnology systems for industrial or municipal applications. Such systems require significant investments of capital up front. Their commercial viability depends on what payments a producer of waste will pay as well as what price can be obtained for the vermicast and associated products that come out at the end of the process. The demand for compost worms from this source really depends on additional facilities being set up. Approximately 60% ormore of household waste in Asian region is of an organic type that could be recycled using Vermitechnology. Many Governments in this region have committed to reduce the amount of organic wastes going to landfill. There are thus environmental, economic and regulatory reasons for an increase in demand for compost worms. One area that is poised for development in future is 'contract waste management' using vermiculture. Vermitech is an integrated operation and quality assurance process which focuses on product quality and public health. Vermitech System has a number of benefits:

- □ No odour
- □ Cost effective Pollution free Valuable end product
- □ Destruction of pathogens
- □ Low green house emissions
- □ Established on site - no cartage
- □ Scalable to suit any volume
- □ Environmentally Sustainable process

ANIMAL MANURES

Use of animal manure as primary feed for earthworms is very common in Vermitech systems. For instance, Vermiculture Production Centre in Pinar del Rio Province, one of the largest of Cuba's 170 vermicomposting centres uses cow manure as its primary feedstock for earthworms, in addition to pig and sheep manure, sugar cane pulp, coffee pulp, and other crop residues. Cattle solids are the most suitable of all animal wastes for earthworm biomass increase They usually do not have materials that deter the growth of earthworms. Cow dung slurry is a suitable substrate for vermicomposting, both when mixed with solid materials or on applying to the surface of bedding materials containing earthworms. Hand et al. (1988) have reported that the mixture of slurry with paper 8% and 15%.

VERMITECH FOR SLUDGE PROCESSING

This is a relatively new process known as vermicomposting or verminstablisation of sewage sludge (bio solids). It is not a true composting, as the process does not transformation. Given the nature of the worm behavior and the bed design and management, the resultant product has a higher stabilization and soil supplement value than traditional composting which relies

on mechanical incorporation of sludge with green waste in large compost heaps. Maximum reduction of the volatile solids is a goal of any sludge stabilization system. If earthworms are to be useful in stabilizing sludge they must increase the rate of volatile solids reduction, thereby increasing the stabilization rate. Vermistablisation studies. According to them E. foetida increases the rate of volatile solid sludge destruction when present in aerobic sludge. There are many fundamental factors that have to be evaluated to assure the technical and economic success of sludge conversion. Pre-treatment of the sludge prior to feeding and the appropriate loading rate are vital to ensure ideal environment for worm activity ensuring the conversion of all wastes. Under favorable conditions, earthworms and microorganisms act symbiotically to accelerate and enhance the decomposition of the organic matter. Increase in the sludge solids destruction rate reduces the probability of putrefaction occurring in the sludge due to anaerobic conditions. The rapid degradation of organic matter may be due to the increased aeration and other factors brought about by the earthworms (Loehr, et al., 1988). Bhiday (1995) reported that the aerobic and the anaerobic stages of the sludge help convert the organic matter into the right form for rapid consumption and digestion by the earthworms.

RESEARCH POTENTIAL

There is a greater need to find out an alternate solution for the sustainable solid waste management in tropical countries. The consumption of organic wastes by earthworms is an ecologically safe method in the natural conversion of many of our organic wastes into an extremely environmentally beneficial product. The tinycreatures' ability to devour virtually any organic waste-livestock manure, rotten food, even ratty T-shirts and excrete it as premium organic fertilizer (dubbed "black gold" by organic farmers for its nitrogen richness) is proving profitable for a host of non-squeamish entrepreneurs. Research findings, developmental programs and application aspects of MSW composting by earthworms in tropical countries need to be propagated and commercialized.

CONCLUSION

Vermicomposting for resource recovery and recycling of MSW is one of the fastest growing sectors in waste management. The application of Vermitechnology and vermiculture is not new for composting, as it is a natural contributor for farming and gardening. In North America, Europe, Asia and African regions, and in several other countries earthworms are being used for various waste treatment options. They could help waste managers for minimizing waste input to landfills and saving precious groundwater resources. In addition, vermicomposting will be helpful for managing domestic solid waste problems and could stabilize wastes with low toxicity, pathogens and heavy metals. The eco-solid waste management could successfully promote vermicomposting as a viable alternative for the disposal of solid wastes.

REFERENCES

1. Aiyyar, M. E. and Ananthakrishnan, T. N. (1993). A manual of Zoology, Volume-I,S. Viswanathan Publishers and Printers, Chennai, India Bhiday, M.H. (1995). Wealth from Waste. 'Vermiculturing' Tata Energy ResearchInstitute, New Delhi, India.ISBN 81-85419-11-6.
2. Bhawalkar U. S, (1995), 'VermicultureEcotechnology', Bhawalkar Earthworm
3. Research Institute, Pune, India Butt, K. R. (1999) Inoculation of Earthworms into Reclaimed Soils: The UK Experience. Land Degradation and Development 10, 565-575.
4. CAPART (1988), 'Vermicompost', Council for Advancement of People's Action and Rural Technology, Centre for Technology Development, New Delhi. India Daniel, T and Karmegam, N.(1999). Bio- conversion of selected leaf litters using an African epigeic earthworm, E. eugeniae.Eco.Env.& Cons. 5 (3): pp 271- 275.
5. Daniel, T and Karmegam, N. (2000). Asian Journal of Microbiology, Biotechnology and Environmental Sciences, Global Science Publications, India, Vol. 2 No. (1-2), pp 63-66.
6. Datar M T, Rao M N, and Reddy S. (1997). Vermicomposting - a technological option for solid waste management. J. Solid Waste Tech. and Management 24(2): 89–93.

7. Edwards, C.A. and Lofty J.R., (1976), Biology of Earthworms. Bookworm Publishing Company, Crawfordsville, Indiana. ISBN 0-916302-20-2. Publisher details to be checked Edwards, C.A., (1998). Breakdown of animal, vegetable and industrial organic waste by earthworms, 'Earthworms in waste and environmental management.' edited by Edwards, C. A &Neuhauser, E. F SPB Academic Publishing, Netherlands. ISBN 90-5103-017-7.
8. Govindan, V.S., (1998). Vermiculture and Vermicomposting, Ecotechnology for pollution control and Environmental Management. pp. 48- 57.
9. Hand, P., Hayes, W. A., Satchell, J.E., Frankland, J.C.,(1988). The vermicomposting of cow slurry, 'Earthworms in waste and environmental management.' edited by Edwards, C. A. &Neuhauser, E. F. SPB Academic Publishing, Netherlands. ISBN 90-5103-017-7.
10. Ismail, S.A., (1997). Vermicolgy 'Biology of earthworms' Orient Longman Limited, Chennai, India. ISBN 81-250-10106.
11. Kale, R. D. (1998). Earthworm: Nature's gift for utilization of organic wastes, 'Earthworm Ecology', edited by Edwards, C.A. St. Lucie Press New York. ISBN 1- 884015-74-3. pp 355- 376.
12. Kale, R. D. and Sunitha, N.S., (1995). Efficiency of earthworm E. eugeniae in converting the solid waste from aromatic oil extraction units into vermicompost, Journal IAEM, Vol. 22, No. 1, pp 267-269
13. Lakshmi, B.L and Vijayalakshmi, G.S., (2000). Vermicomposting of sugar factory filterpressmud using African earthworm species E. eugeniae, Pollution Research, Vol.19, No.3, pp.481-483.
14. Loehr, R.C., Martin, J.H., Neuhauser, E.F., (1998). Stabilization of liquid municipal sludge using earthworms, 'Earthworms in waste and environmental management.' edited by Edwards, C. A &Neuhauser, E. F. SPB Academic Publishing, Netherlands. ISBN 90-5103-017-7.

TREMENDOUS ACTIVITY OF HERBAL DRUGS AGAINST NEPHROLITHIASIS

Mohd Tasleem[1], Snigdha Tiwari[2], Mamta Baunthiyal[3], Shalini[4], Veena Maurya[5]

[1,2,3]Ph.D scholar, G B Pant engineering college, PauriGarhwal, Uttarakhand.
Mohdtasleem99@gmail.com
[4,5] Baba Mastnath university, AsthalBohar, Rohtak (124021), Haryana.

Urolithiasis (nephrolithiasis) or kidney stone is formation of urinary calculi at any level of urinary tract. They are crystal aggregations formed in the kidneys. It is estimated that 12% of world population experiences renal stone disease with a recurrence rate of 70-80% in male and 47-60% in female (Sundararajan et al., 2006). Urinary calculi are the third most common affliction of the urinary tract which is exceeded by the urinary tract infections and prostate diseases (Hamid et al., 2007). Renal calculi can be broadly classified in two large groups: tissue attached and unattached. Attached calculi are mainly integrated by calcium oxalate monohydrate (COM) renal calculi, with a detectable attachment site to the renal papilla and basically consisting of a core located near to the attachment site (concave zone) and radially striated concentrically laminated peripheral layers. Unattached calculi, with no detectable site of attachment to papilla, are developed in renal cavities of low or reduced urodynamic efficacy and can exhibit diverse composition and structures (Grases et al., 2002; Kuo et al., 2003). Several reports have been published since Randall's first description of papillary calcifications and their possible active role in the genesis of COM papillary calculi (Low et al., 2000; Evan et al., 2005). The most common type of kidney stones worldwide contains calcium. For example, calcium-containing stones represent about 80% of all cases in the United States; these typically contain calcium oxalate either alone or in combination with calcium phosphate in the form of apatite orbrushite (Coe et al., 2005). About 10–15% of urinary calculi are composed of struvite (Ammonium magnesium phosphate). Struvite stones (also known as "infection stones", urease or triple-phosphate stones), form most often in the presence of infection by urea-splitting bacteria. Cystine kidney stones are due to cystinuria, an inherited disorder of the transport of an amino acid called cystine that results in an excess of cystine in the urine (cystinuria) and the formation of cystine stones. The presence of renal calculi is diagonised by the symptoms explained by the patients and the stones are recognized in the body with the help of X-rays. The analytical markers in urine and serum that are responsible for the clinical diagnosis of the urologic disorders are calcium, albumin, creatinine, urate and oxalate (Suman et al., 2011). The problem of stone formation produces pain and obstruct the flow of urine as the stones formed are unable to travel through ureter, It also causes, severe back ache (the worst pain known as colicky pain is produced in the lower back), sickness, urge for urination, burning sensation during urination, fever, less urine volume, change in urinary pH, and infections (Kalpana et al., 1993). An attempt has been made during the last decade to study the identical, chemistry, pharmacology and clinical investigations of Pashanbheda plants used for dissolving kidney stones. Herbal drugs have created interest among the people by its clinically proven effects like immunomodulation, adaptogenic and antimutagenic. Number of medicinal plants shows antiurolthiatic activity and play vital role in prevention of disease. Here an attempt has been made to emphasis on potent indigenous herb for urinary stone (Havagiray et al., 2010).

Herbal medicines have several phytoconstituent and exert their beneficial effects urolithiasis by multiple mechanisms like:

- Helps in spontaneous passage of calculi by increasing urine volume, pH and anti-calcifying activity (Diuretic activity).

- Balance the Inhibitor and promoter of the crystallization in urine and affects the crystal nucleation, aggregation and growth (Crystallization inhibition activity).
- Relieves the binding mucin of calculi (lithotriptic activity).
- Improved renal function.
- Regulation of oxalate metabolism.
- Regulates the crystalloid colloid imbalance and improve renal function, thus prevents recurrence of urinary calculi.
- Improve renal tissue antioxidant status and cell membrane integrity and prevent reoccurrence (Antioxidant activity).
- ACE and Phospholipase A2 Inhibition.
- Exerts significant anti-infective action in against the major causative organisms (Antimicrobial activity).
- Reveals marked improvement in symptoms of urinary calculi like pain, burning micturition and haematuria (Analgesic and anti-inflammatory activity).

We have studied of two medicinal herbal plants Macrotylomauniflorum and Bergeniaciliata. Horse Gram is scientifically known as Macrotylomauniflorum. It also goes by the name Dolichosuniflorus due to a lot of confusion in the Dolichos category the right name for the horse gram scientifically is Macrotylomauniflorum. According to USDA (United States Department of Agriculture) database both the name's Macrotylomauniflorum and Dolichosuniflorus mean the same. Horse Gram is native to the old world tropics. BergeniaciliataWall., (B. ciliata) commonly known as Paashaanbhed in the Indian systems of medicine. Bergeniaciliata a small perennial herb found throughout temperate Himalayas from Bhutan to Kashmir at an altitude between 2000-3000 m and in Khasia hills upto 1200 m altitude. It is perennial herb upto 35 cm tall and very commonly found in rocks. It is native to central Asia and found in Afghanistan, China and in Himalaya. In Uttarakhand it is found in both region of Garhwal and Kumaun. In our study, the aqueous extract of B. ciliata rhizomes and M. uniflorum seeds useful in the treatment of urinary calculi, was evaluated for its safety and efficacy. Both plants B. ciliata and M. uniflorum were found to be effective in dissolving kidney stone. According to our study, B. ciliata showed maximum Litholytic activity as compare to M. uniflorum dissolving renal stones. The observed beneficial effects in the management of urolithiasis following an herbal formulation treatment could be due to the prevention of urinary supersaturation, inhibition of mineralization of stone forming constituents, normalization of cellular function in renal oxidative stress, correction of crystalloid-colloid balance as well as the beneficial effects such as anti-inflammatory, antimicrobial, diuretic, antispasmodic, litholytic, and anticalcifying activities of individual ingredients. Urolithiasis (Mutrashmari) in Ayurveda is described as a painful disease which needs to be addressed early. In the beginning stages, disease can be effectively managed with medical intervention, while in the later stages with enlarged stones, the only successful measure would be lithotripsy. The medical management would include the administration of lithotriptic herbs and substances besides the measures to manage the bio-energies that control the disease manifestations. With this line of treatment, Ayurveda envisages that the stone forming substances are controlled, which prevents recurrence of the stone.

Key words: Nephrolithiasis, Macrotylomauniflorum, Bergeniaciliata, Lithotripsy, Cystinuria and Herbal drugs.

REFERENCES

1. Umashankar, D., Chandra, R., Chawla, A.S., Deepak, M., Singh, D. and Handa, S.S. (1999). High pressure liquid chromatographic determination of bergenin and (+)-afzelechin from different parts of Paashaanbhed (Bergenialigulata). Phytochem Anal, 10(1): 44–7.
2. Venkatadri, R., Guha, G. and Rangasamy, A. K. (2011). Anti-neoplastic activity of Bergeniaciliata rhizome. J Pharmacy Res, 4(2): 443-445.

3. WHO, in Progress Report by the Director General, Document NO. A44/20, 22 March 1991, World health organization, Geneva.
4. Sinha, S., Murugesan T., Maiti, K., Gayen, J. R., Pal, B., Pal, M. and Saha, B. P. (2001). Antibacterial activity of Bergenia ciliate rhizome.Fitoterapia 72(5): 550-552.
5. Sinha, S., Murugesan, T., Pal M. and Saha B. P. (2001). Evaluation of anti-tussive activity of BergeniaciliataSternb.rhizome extract in mice. Phytomedicine 8(4): 298-301.
6. Rana, Gopal, S., Sanjeev, Kumar, B. and Rakesh, K. (2010). Litholytic Property of Kulattha (DolichousBiflorus) vs. Potassium Citrate in Renal Calculus Disease: A Comparative Study. JAPI, 58: 287.
7. Ravishankar, K. and Vishnu, Priya P. S. (2012). Evaluation of diuretic effect of ethanolic seed extracts of Macrotylomauniflorum and Cucumismelo in rats. Int J Pharm Bio Sci, 3(3): 251- 255.
8. Koji, S., Kenji, K. and Ryuzo, T. (1999). Formation and growth inhibition of calcium oxalate crystals by Takusha (Alismatisrhizoma). Scanning Microscopy, 13(2-3): 183-189.
9. Kore, K. J., Shete, R. V., Jadhav, P. J. and Kabra, M. P. (2011). Antiurolithiatic effects of hydroalcoholic extract of lawsoniainermis leaves. International Journal of Universal Pharmacy and Life Sciences, 1(2); 103-105.
10. Krishna, K.L., Paridhavi, M. and Jagruti, A.P. (2008). Review on nutritional, medicinal and pharmacological properties of Papaya (Carica papaya Linn.). Natural Product Radiance, 7(4): 364-373.

BIOELECTRONICS: A NOVEL CONCEPT IN BIOMEDICINE

Neha Katyal*[1], Ashish Katyal[2],Dr. S.P.Sharma[1]
[1] Department of Physics, Baba Mastnath University, Rohtak- 124001, India
(nehakatyal1@gmail.com) M- 09728255584.
[2]Department of Biotechnology, Meerut Institute of Engineering and Technology, Meerut-250005, India (aashishkatyaal@gmail.com) M- 09997031901

ABSTRACT:
Today, the organic electronics field stands at the interface with biology. The novel field of organic bioelectronics has grown tremendously over the last few years and it is very exciting today to see how the advantages and characteristics of this technology have paved the way for novel applications in nanomedicine. This paper gives the review of concept of bioelectronics and its multidisplinary application by adaption of methods of micro-fabrication to make structures for the investigation of the interface between biological molecules and electronic materials. Such structures may be made from either conducting or non-conducting materials, and can be patterned over a wide area with sub-micron scale accuracy, either to investigate electrical or electrochemical systems or to control cell movement.

INTRODUCTION

With the rapid development of nanotechnology, the suitable integration of numerous inorganic nanomaterials with outstanding optical, electronic, and magnetic performance into polymeric systems has attracted significant attention in the research community. Future information technologies is aimed at the miniaturization of devices down to ultimate limits as determined by basic physics and quantum mechanical principles. An often considered example concerns the *human brain* as compared with the *man-made computer*. Technologically unmatched performances of the brain concern high information density, low power consumption, high flexibility, excellent association memory, etc. Biologically unmatched performances of the computer concern quantitative information processing, high reproducibility, etc. With increasing complexity and demands for future information technologies, a trend is to be seen towards the design of "smart" nanostructures which will be interfaced to silicon or other substrates. These structures may consist either of chemically synthesized units such as molecules, supramolecules and biologically active (biomimetic) recognition centers, or of natural and hence vey complex biomolecular function units with synthesized units such as molecules, supramolecules and biologically active (biomimetic) recognition centers, or of natural and hence vey complex biomolecular function units with high molecular weight which may be extracted from biological systems. Bioelectronics is aimed at the direct coupling of biomolecular function units of high molecular weight and extremely complicated molecular structure with electronic or optical transducer devices. This requires the development of structures for signal uptake, transduction, amplification, processing and conversion.

The tremendous biochemical and biotechnological progress in tailoring new biomaterials by genetic engineering or bioengineering provides unique and novel means to synthesize new enzymes and protein receptors, and to engineer monoclonal antibodies or aptamers for nonbiological substrates (such as explosives or pesticides) and DNA-based enzymes. All these materials provide a broad platform of functional units for their integration with electronic elements. The latter electronic elements may involve, for example, electrodes, field-effect transistor devices, piezoelectric crystal, magneto resistance recording media, scanning tunnelling microscopy (STM) tips and others. The bioelectronics devices, Figure 1,

may operate in dual directions: In one configuration, the biological event alters the interfacial properties of the electronic element, thus enabling the readout of the bioreaction by monitoring the performance of the electronic unit such as the readout of the potential, impedance, charge transport, or surface resistance of electrodes or field-effect transistors, or by following the resonance frequencies of piezoelectric crystals. The second configuration of bioelectronic systems uses the electronic units to activate the biomaterials toward desired functions.

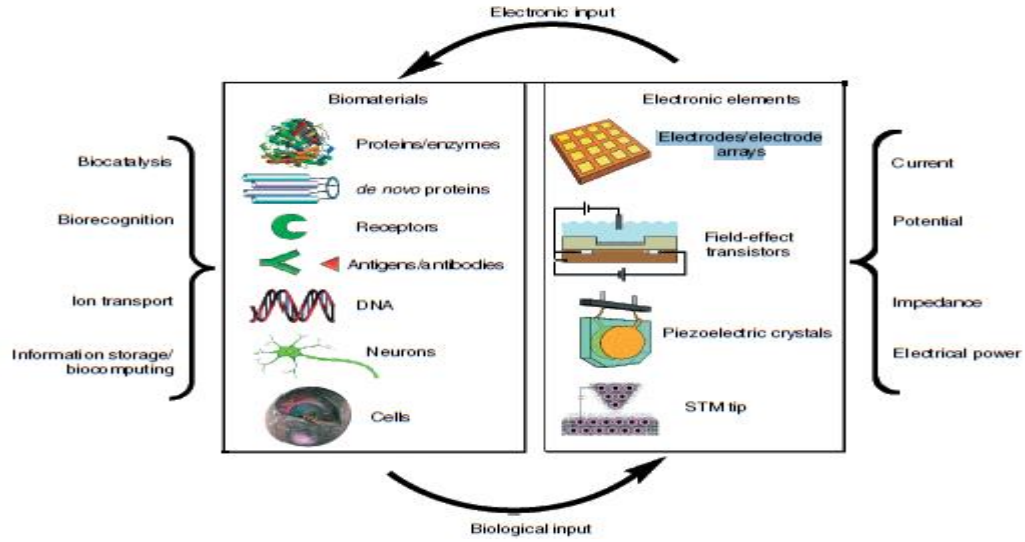

Fig 1: Integrated system for biomaterials and electronic elements for bioelectronic applications

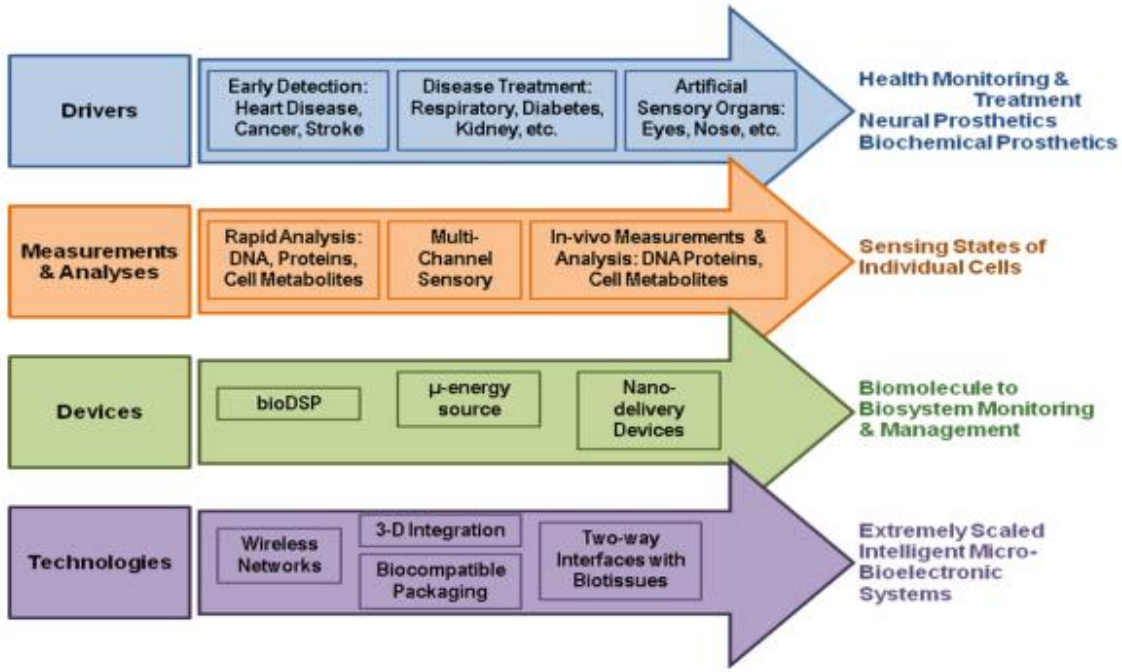

Fig2: Basic application of Bioelectronic devices

APPLICATIONS:

The electrical contacting between biomolecules and electrodes is an essential feature for most bioelectronic systems. There are various applications of Bioelectronic systems:

1) The virgin field of systems biology uses systems engineering approaches to analyze the cellular function in Real-Time and resulted into **Massively Parallel Molecular and Cellular Characterization** for Systems Biology. Systems biology comprise of new way from which biological systems and knowledge can be culled from its approaches could lead to advances in medicine and security.
2) Cells and their components can be used as biological transducers for measurements or as components in building novel materials or circuits and can act as **Biological Sensors.**
3) New ways for **Protection and Restoration** of Health. As advances has been observed in miniaturization and power transmission, storage, and generation allows implantable medical devices to emerge. For example The artificial retina, which is a first step toward restoring sight in people with degenerative diseases of the retina, is representative of these advancements.
4) Biomolecular measurements are increasingly focused on **higher bandwidth measurements as biomolecules** are generally present in low concentration so it it difficult to measure and analyse them correctly. By the use of Bioelectronics we can overcome this problem by Rapidly determine whole genome and RNA sequences, Identify biomarkers for disease states, Simultaneously determine membrane protein structure and function and Monitoring multiple intracellular events in real-time etc.
5) Biosensors will play a key role in meeting future bioelectronics demands and improvements to increase bandwidth and lower detection limits are needed. New methods are needed to **fabricate structures** reliably at the nanoscale, and new metrology and standards are needed to characterize these structures. By the use of Bioelectronics we can do better Fabrication resulted into Integrating high-density photodiode arrays into substrates, Immobilization of electrode surfaces for control of antibody and antigen binding, Improving fabrication techniques for on-chip integrated optical excitation and detection in nanoliter-femtoliter volumes and Developing electronics with a sensitivity of < 1 pA and a bandwidth of hundreds of MHz etc.
6) By the use of Bioelectronics we can make **biocompatible materials** that do not kill tissue or cells and maintain the native structure of biomolecules, Limiting implant rejection by maintaining and stimulating tissue integrity.
7) Bioelectronics is useful in Improving the efficiency of RF power transmission to devices within the body. It can also Improve battery energy density. It can also Develop alternative **energy generation** strategies such as power harvested from body motion and synthetic photosynthesis etc.
8) Bioelectronics is helpful in **Forensics** by generate new lab-on-a-chip devices that can perform rapid DNA analysis for example It allows law enforcement officials to conduct testing at the scene of the crime.
9) Bioelectronics is useful in **Medicine** by creating artificial retina, which records light intensity and then transmits this information to nerves within the retina. These devices promise to allow selective stimulation of particular regions of the brain to restore patient function.

CONCLUSION:

In early days, the semiconductor industry has focused on applications and products for which it can provide significant added value, for example in information and communications technologies. Today, the field of bioelectronics is essential for exponential growth in certain critical areas of science and technology, including sensors, forensics, medicine, energy generation, formation of biocompatible materials, fabricate new structure at nanoescale by used of nanoelectronics, and coordinated effort of academia and industries. It is now clear that the eventual commercialization of

bioelectronics will require expertise from both the biomedical and electronics industries, that can understands the healthcare market and its regulatory framework and business model.

REFERENCES

[1] Willner, I. and Katz, E. (eds) (2005) *Bioelectronics: From Theory to Applications*, Wiley-VCH Verlag GmbH, Weinheim.

[2] Pethig, R.R. and Smith, S. (2012) *Introductory Bioelectronics: For Engineers and Physical Scientists*, John Wiley & Sons, Ltd, Chichester.

[3] Sarpeshkar, R. (2010) *Ultra Low Power Bioelectronics: Fundamentals, Biomedical Applications, and Bio-Inspired Systems*,Cambridge University Press, Cambridge.

[4] Banica, F.-G. (2012) *Chemical Sensors and Biosensors: Fundamentals and Applications*, John Wiley & Sons, Ltd, Chichester.

[5] Li, S., Singh, J., Li, H., and Banerjee, I.A. (eds) (2011) *Biosensor Nanomaterials*, Wiley-VCH Verlag GmbH, Weinheim.

[6] C.A. NICOLINI (Ed.), *Biophysics of Electron Transfer and Molecular Bioelectronics*, Plenum Press, New York, **1998**.

[7] K.-H. HOFFMANN (Ed.), *Coupling of Biological and Electronic Systems*, Springer-Verlag, Berlin, **2002**.

[8] L. HABERM ¨ ULLER, M.MOSBACH, W. SCHUHMANN, *Fresenius' J. Anal. Chem.*, **2000**, *366*, 560–568.

[9] A. HELLER, *Acc. Chem. Res.*, **1990**, *23*, 128–134.

[10] F.A. ARMSTRONG, G.S. WILSON, *Electrochim. Acta*, **2000**, *45*, 2623–2645.

ELECTRONIC AND OPTICAL PROPERTIES OF SEMICONDUCTOR

Kusum Rani and Sheetal

Baba Mastnath University, Asthal Bohar, Rohtak- 124021 India(Vatskusum6@gmail.com),M-9034323301

ABSTRACT

For strained self-assembled InAs quantum dots on GaAs or InP substrate atomic positions and strain distribution are described using valence-force field approach and continuous elasticity theory. The strain is coupled with the effective mass, k · p, effective bond-orbital and atomistic tight-binding models for the description of the conduction and valence band states. The single-particle states are used as input to the calculation of optical properties, with electron-electron interactions included via configuration interaction (CI) method. This methodology is used to describe multiexciton complexes in quantum dot lasers, and in particular the hidden symmetry as the underlying principle of multiexciton energy levels, manipulating emission from biexcitons for entangled photon pairs, and optical control and detection of electron spins using gates. The self-assembled quantum dots are compared with graphene quantum dots, one carbon atom-thick nanostructures. It is shown that the control of size, shape and character of the edge of graphene dots allows to manipulate simultaneously the electronic, optical, and magnetic properties in a single material system.Here we start with a brief discussion of the important bulk and epi-taxial crystal growth techniques. We then discuss the important semiconductor crystal structures. We also discuss strained lattice structures and the strain tensor for such crystals. Strained epitaxy and its resultant consequences are now widely exploited in semiconductor physics and it is important to examine how epitaxial growth causes dis-tortions in the crystal lattice.

INTRODUCTION

Characterization of electronic properties of localized electronic states (shallow or deep) in bulk semiconductor materials; interface states in semiconductor-insulator structures and electron properties of two-dimensional systems at the same boundary; defects and impurities (bulk and interface) in heterostructures (single junctions and multilayer); super lattices and layered compounds and thin slabs and quantum wells of these materials. The investigated materials are extremely important for device engineering for informatics and communication technique such as: silicon; semiconductors A3B5 and their the many compounds (GaAs, InSb, AlGaAs, InGaAs); layered compounds (GaSe,InSe); heterostructures Si/SiO2, InSb/SiO2, GaAs/AlGaAs, superlattices and microstructures from these materials. Electrical and optical characterization methods are used. The following equipment isavailable:1) Raman spectroscopy, Photoluminescence (PL) and Photoconductivity. A high resolution double monochromator (SPEX Model 1404, f=0.85m) is used for Raman and PL spectroscopy. A set of holographic gratings (1800 grooves/mm) and another two sets of gratings (600 gr/mm and 300 gr/mm blazed at 1000 and 2000 nm respectively) are available. Both visible and NIR detection are possible. An RCA C31034photomultiplier (GaAs photocathode) in photon counting mode is used in the range 180-900nm. A PbS detector cooled with liquid nitrogen covers the range 900-4500nm.Another low resolution monochromator (SPEX Minimate, f=0.22m) coupled with thePbS detector is used to record the deep level PL of several materials (e.g. GaAs). This monochromator is equipped with a set of gratings covering the range from visible up to 40Characterization of electronic properties of localized electronic states (shallow ordeep) in bulk semiconductor materials; interface states in semiconductor-insulator structures and electron properties of two-dimensional systems at the same boundary; defects and impurities (bulk and interface) in heterostructures (single junctions and multilayers); superlattises and layered compounds and thin slabs and quantum wells of these materials. The investigated materials are extremely important for device engineering for informatics and communication technique such as: silicon; semiconductors A3B5and their the mary compounds (GaAs, InSb, AlGaAs, InGaAs); layered compounds (GaSe,InSe); heterostructures Si/SiO2, InSb/SiO2, GaAs/AlGaAs, superlattises and microstructures from these materials. Electrical and optical characterization methods are used. The following equipment isavailable:1) Raman spectroscopy, Photoluminescence (PL) and Photoconductivity. A high resolution double monochromator (SPEX Model 1404, f=0.85m) is used for Raman and PL spectroscopy. A set of holographic gratings (1800 grooves/mm) and another two sets of gratings (600 gr/mm and 300 gr/mm blazed at 1000 and

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$$\Delta C/C \cong 10\text{-}7$$

and the following modes of peration: DLTS, DDLTS, C-V, C-T, C-t, N-W,N-V. It is intended to develop different optical DLTS modes as DLOS.

ELECTRONIC PROPERTIES OF SEMICONDUCTOR

The semiconductor materials described here are single crystals; i.e., the atoms are arranged in a three-dimensional periodic fashion. Each silicon atom in the crystal is surrounded by four of its nearest neighbours. Each atom has four electrons in its outer orbit and shares these electrons with its four neighbours. Each shared electron pair constitutes a covalent bond. The force of attraction between the electrons and both nuclei holds the two atoms together. For isolated atoms (e.g., in a gas rather than a crystal), the electrons can have only discrete energy levels. However, when a large number of atoms are brought together to form a crystal, the interaction between the atoms causes the discrete energy levels to spread out into energy bands. When there is no thermal vibration (i.e., at low temperature), the electrons in an insulator or semiconductor crystal will completely fill a number of energy bands, leaving the rest of the energy bands empty. The highest filled band is called the valence band. The next band is the conduction band, which is separated from the valence band by an energy gap (much larger gaps in crystalline insulators than in semiconductors). This energy gap, also called a bandgap, is a region that designates energies that the electrons in the crystal cannot possess. Most of the important semiconductors have bandgaps in the range 0.25 to 2.5 electron volts (eV). The bandgap of silicon, for example, is 1.12 eV, and that of gallium arsenide is 1.42 eV. In contrast, the bandgap of diamond, a good crystalline insulator, is 5.5 eV. At low temperatures the electrons in a semiconductor are bound in their respective bands in the crystal; consequently, they are not available for electrical conduction. At higher temperatures thermal vibration may break some of the covalent bonds to yield free electrons that can participate in current conduction. Once an electron moves away from a covalent bond, there is an electron vacancy associated with that bond. This vacancy may be filled by a neighbouring electron, which results in a shift of the vacancy location from one crystal site to another. This vacancy may be regarded as a fictitious particle, dubbed a "hole," that carries a positive charge and moves in a direction opposite to that of an electron. When an electric field is applied to the semiconductor, both the free electrons (now residing in the conduction band) and the holes (left behind in the valence band) move through the crystal, producing an electric current. The electrical conductivity of a material depends on the number of free electrons and holes (charge carriers) per unit volume and on the rate at which these carriers move under the influence of an electric field. In an intrinsic semiconductor there exists an equal number of free electrons and holes. The electrons and holes, however, have different mobilities; that is, they move with different velocities in an electric field. For example, for intrinsic silicon at room temperature, the electron mobility is 1,500 square centimetres per volt-second ($cm^2/V{\cdot}s$)—i.e., an electron will move at a velocity of 1,500 centimetres per second under an electric field of one volt per centimetre—while the hole mobility is 500 $cm^2/V{\cdot}s$. The electron and hole mobilities in a particular semiconductor generally decrease with increasing temperature.

Electrical conduction in intrinsic semiconductors is quite poor at room temperature. To produce higher conduction, one can intentionally introduce impurities (typically to a

concentration of one part per million host atoms). This is called doping, a process that increases conductivity despite some loss of mobility. For example, if a silicon atom is replaced by an atom with five outer electrons, such as arsenic, four of the electrons form covalent bonds with the four neighbouring silicon atoms. The fifth electron becomes a conduction electron that is donated to the conduction band. The silicon becomes an *n*-type semiconductor because of the addition of the electron. The arsenic atom is the donor. Similarly, if an atom with three outer electrons, such as boron, is substituted for a silicon atom, an additional electron is accepted to form four covalent bonds around the boron atom, and a positively charged hole is created in the valence band. This creates a *p*-type semiconductor, with the boron constituting an acceptor.

OPTICAL PROPERTIES OF SEMICONDUCTOR:

We note that in GaAs the top of the valence band and the bottom of the conduction band both have the same wavevector In this case we say that the semiconductor has a direct gap On the other hand, in Si the top of the valence band is at, while the bottom of the condution band is at ∼0.8 In this case we say that Si has an indirect gap The nature of the band gap (direct or indirect) bears important implications on the optical properties of a semiconductor. In fact, when we shine light on a semiconductor in the frequency range between the infrared and the visible (1 eV to 4 eV, i.e. wavelenghts from 1200 nm to 300 nm), light absorption can take place only if the energy of the photons can be transferred to the electrons. This happens if the electrons can be promoted from the valence band to the conduction band by overcoming the band gap. It is possible to use "perturbation theory" in order to calculate the absorption spectra of semiconductors (see for instance Yu and Cardona, Fundamentals Of Semiconductors). The derivation is rather di cult, therefore we only discuss the qualitative results. In the absorption process energy and momentum (wavevector) are conserved.

REFERENCES

1. Milton Ohring Reliability and failure of electronic materials and devices Academic Press, 1998 ISBN 0-12-524985-3, p. 310
2. John Dakin, Robert G. W. Brown Handbook of optoelectronics, Volume 1, CRC Press, 2006 ISBN 0-7503-0646-7 p. 57
3. Safa O. Kasap, Peter Capper (2006). *Springer handbook of electronic and photonic materials*. Springer. pp. 54,327. ISBN 0-387-26059-5.
4. Y. Tao, J. M. Boss, B. A. Moores, C. L. Degen (2012). Single-Crystal Diamond Nanomechanical Resonators with Quality Factors exceeding one Million. arXiv:1212.1347
5. Kittel, Charles. *Introduction to Solid State Physics, 7th Edition*. Wiley.
6. http://www.matweb.com/search/datasheet.aspx?matguid=64d7cf04332e428dbca9f755f4624a6c
7. Abass, A. K.; Ahmad, N. H. (1986). "Indirect band gap investigation of orthorhombic single crystals of sulfur". *Journal of Physics and Chemistry of Solids* **47** (2): 143. doi:10.1016/0022-3697(86)90123-X. edit
8. Evans, D A; McGlynn, A G; Towlson, B M; Gunn, M; Jones, D; Jenkins, T E; Winter, R; Poolton, N R J (2008). "Determination of the optical band-gap energy of cubic and hexagonal boron nitride using luminescence excitation spectroscopy". *Journal of Physics: Condensed Matter* **20** (7): 075233. Bibcode:2008JPCM...20g5233E. doi:10.1088/0953-8984/20/7/075233.

ROLE OF MATHEMATICS IN MEDICAL SCIENCE

Minakshi

Baba Mastnath University, Asthal Bohar, Rohtak-124021E-mail: mann80m@gmail.com

ABSTRACT

Today in 21st century is crucial to have the best experts possible. Societies that focus on human resources have chances to answer the challenges of modern time we live in. To accomplish the wanted level of scientific and technological literacy of professionals, it is needed to reconstruct the old school system into new one, where we could change points of view, concepts and relationships. First of all, we must make a new concept of a teacher, student, study, teaching and new relationship between teachers and students. It is also important to modernize the relations among all the participants in the process of education. Teachers should always have in mind that the only way to come to scientific results is in corporation with each other within a certain project. In developed countries, this way of communication among scientists, is at the highest level, as the most important national priority. The mathematics itself is in the basis of technological and scientific literacy, and this article is going to explain these relationships, focusing medical sciences.

INTRODUCTION

The application of mathematics in the field of science has been in place since times immemorial. However, its actual importance and applicability in medical sciences has become more visible in today's world. To understand it medically, one can say that the application of mathematics in medical sciences has been sporadic. The study of mathematics and the subject, in general, has been contributing to biology and the study of life sciences in many ways.

Today, biologists and professionals in the field of medical sciences use the concepts of mathematics, its theorems, and formula, among others in obtaining quantitative measures and analyzing their studies in quantity along with quality. Many scientific medical equipments and machines have been designed keeping the applications of mathematics in mind. Biologists use mathematics to derive quantitative data from any medical study, such as estimating the number of genes that are involved in inheriting a particular trail from the parents to the child.

Today, the concepts and theories of mathematics are being applied to almost all aspects of medical sciences. However, there are some fields of study under medical sciences where mathematics has found greater scope. These fields are clinical research, biotechnology, and genetics. To discuss first, let's talk about how mathematics and its sub branches play an important role in the study of genetics.

In the study of genetics, the two most important sub sections of mathematics that play a major role are statistics and probability. Through probability, its formula and applications, people working in genetics understand the mechanism of meiosis, forming of egg cells and the sperm, and the process of inheritance, among others. They also get an understanding of how phenotypes (observable diseases) and genotypes (the DNA sequences) are related to each other through the study of probability distributions. They also conduct the analysis and study of genetic determinants through the applications of statistics and its methods and principles. Owing to this reason, a student who wishes to pursue a higher education degree or course in genetic studies should have studied Mathematics at class 12 level.

Another advantage that the study of mathematics in medicine and biology provides is to mathematical modelers. Mathematical modeling refers to looking and studying the biological systems and the medical phenomena through the study of mathematics by applying its principles and theorems. Mathematics has also shown its worth in testing out new ideas in the

field of medicine. This has been or high importance in the study of cancerous and tumor cells. Medical practitioners have been successful in understanding the direction in which cancerous cells potentially grow inside the living body with the help of a model strictly based on mathematics.

Many health organizations are now using online systems and software applications to manage their patient records and keep a check on the remedial steps for each patient, coordinating with practitioners and other hospital staff on a regular basis along with many more things. All these online systems and software applications are developed keeping mathematics as the platform.

In biotechnology as well, mathematics is seen to play a major role. The study of mathematics is needed to understand concepts such as concentration/dilution, calibration, molarity, molality, solution preparation, serial dilution, radioactive decay and decomposition, absorption, and cell growth, among others. In clinical research, the understanding of DNA, thumb impressions, and markups, among others is also carried out based on the study of mathematics.

Looking at the new avenue of career opportunities that these fields provide, many students are now choosing to club together mathematics along with medical sciences to open doors to a new skill set that is much in demand and gives lucrative job options to them. Some of the career options available are in the field of biotechnology, clinical research, and genetics. Thus, a student who wants to pursue a career in any of these mentioned fields should necessarily have studied both mathematics and biological sciences in class 12. In addition to these three core branches, there are many other branches, which are open to students that want to make a career out of this combination. Other such sub branches are biosciences, health professions, investigation and research, quality assurance, quality checks, biopharmaceutical, regulatory sciences, and clinical sciences, among others. Most of these courses are available at the post-graduation level. Thus, a student wanting to pursue a career in mathematical sciences along with biological sciences should take up a basic bachelor of science (B.Sc.) at graduation level in any of the science streams.

Once successfully completing the basic B.Sc., a student can sit in the entrance tests of universities and autonomous institutions that provide courses in these subjects at post-graduation level. However, some universities also provide a direct five-year M.Tech. course in fields such as Biotechnology where graduation and post-graduation are clubbed together, and a student is directly awarded a post-graduation degree after successful complete of the course.

CONCLUSION

The structure of mathematical theories is universal for any of the scientific theories, and also for the set of scientific theories usually called Medical Sciences. At this point, it is important, for anyone who has an intention to deal with scientific work in medicine, should get familiar with the structure of basic mathematical theories. Also, medical students must be familiar with the basis of mathematical calculus, so that every day medical practice becomes more accurate and beneficial.

REFERENCES

1 .Bellman ER. Mathematical Methods in Medicine. World Scientifing Publishing Co., Inc. River Edge, NJ, USA

2. Feinstein RA. Two Centuries of Conflict Collaboration Between Medicine and Mathematics J. Clin. Epidemiol

3. Breslow EN. Are Statistical Contributions to Medicine Undervalued? Biometrics 2003.

4. Grobe SJ. Nursing informatics competencies. Methods of Information in Medicine.

INTRODUCTION OF CAUCHY'S INTEGRAL FORMULA

VINOD BHATIA

Baba Mastnath University, Asthal Bohar,Rohtak-124021E-mail: bhatiavinod@ymail.com

In mathematics, **Cauchy's integral formula**, named after Augustin-Louis Cauchy, is a central statement in complex analysis. It expresses the fact that a holomorphic function defined on a disk is completely determined by its values on the boundary of the disk, and it provides integral formulas for all derivatives of a holomorphic function. Cauchy's formula shows that, in complex analysis, "differentiation is equivalent to integration": complex differentiation, like integration, behaves well under uniform limits – a result denied in real analysis

THEOREM

Suppose U is an open subset of the complex plane **C**, $f : U \to \mathbf{C}$ is a holomorphic function and the closed disk $D = \{ z : | z - z_0| \leq r\}$ is completely contained in U. Let γ be the circle forming the boundary of D. Then for every a in the interior of D:

$$f(a) = \frac{1}{2\pi i} \oint_\gamma \frac{f(z)}{z-a}\, dz$$

where the contour integral is taken counter-clockwise.

The proof of this statement uses the Cauchy integral theorem and similarly only requires f to be complex differentiable. Since the reciprocal of the denominator of the integrand in Cauchy's integral formula can be expanded as a power series in the variable $(a - z_0)$, it follows that holomorphic functions are analytic. In particular f is actually infinitely differentiable, with

$$f^{(n)}(a) = \frac{n!}{2\pi i} \oint_\gamma \frac{f(z)}{(z-a)^{n+1}}\, dz.$$

This formula is sometimes referred to as **Cauchy's differentiation formula**.

The circle γ can be replaced by any closed rectifiable curve in U which has winding number one about a. Moreover, as for the Cauchy integral theorem, it is sufficient to require that f be holomorphic in the open region enclosed by the path and continuous on its closure.

By using the Cauchy integral theorem, one can show that the integral over C (or the closed rectifiable curve) is equal to the same integral taken over an arbitrarily small circle around a.

Since $f(z)$ is continuous, we can choose a circle small enough on which $f(z)$ is arbitrarily close to $f(a)$. On the other hand, the integral

$$\oint_C \frac{1}{z-a}\, dz = 2\pi i,$$

over any circle C centered at a. This can be calculated directly via a parametrization (integration by substitution) $z(t) = a + \varepsilon e^{it}$ where $0 \leq t \leq 2\pi$ and ε is the radius of the circle.

Letting $\varepsilon \to 0$ gives the desired estimate

$$\left|\frac{1}{2\pi i}\oint_C \frac{f(z)}{z-a}\,dz - f(a)\right| = \left|\frac{1}{2\pi i}\oint_C \frac{f(z)-f(a)}{z-a}\,dz\right|$$

$$= \left|\frac{1}{2\pi i}\int_0^{2\pi}\left(\frac{f(z(t))-f(a)}{\varepsilon\cdot e^{i\cdot t}}\cdot\varepsilon\cdot e^{t\cdot i}i\right)\,dt\right|$$

$$\leq \frac{1}{2\pi}\int_0^{2\pi}\frac{|f(z(t))-f(a)|}{\varepsilon}\,\varepsilon\,dt$$

$$\leq \max_{|z-a|=\varepsilon}|f(z)-f(a)| \xrightarrow[\varepsilon\to 0]{} 0.$$

Example

Let

$$g(z) = \frac{z^2}{z^2+2z+2},$$

and let C be the contour described by $|z| = 2$ (i.e. the circle of radius 2).

To find the integral of $g(z)$ around the contour C, we need to know the singularities of $g(z)$. Observe that we can rewrite g as follows:

$$g(z) = \frac{z^2}{(z-z_1)(z-z_2)}$$

where $z_1 = -1+i, z_2 = -1-i.$

Thus, g has poles at z_1 and z_2 . The moduli of these points are less than 2 and thus lie inside the contour. This integral can be split into two smaller integrals by Cauchy-Goursat theorem; that is, we can express the integral around the contour as the sum of the integral around z_1 and z_2 where the contour is a small circle around each pole. Call these contours C_1 around z_1 and C_2 around z_2.

Now, each of these smaller integrals can be solved by the Cauchy integral formula, but they first must be rewritten to apply the theorem. For the integral around C_1, define f_1) as $f_1)(z)=(z-z_1)g(z)$. This is analytic (since the contour does not contain the other singularity). We can simplify f_1) to be:

$$f_1(z) = \frac{z^2}{z-z_2}$$

and now

$$g(z) = \frac{f_1(z)}{z-z_1}.$$

Since the Cauchy integral theorem says that:

$$\oint_C \frac{f_1(z)}{z-a}\,dz = 2\pi i\cdot f_1(a),$$

we can evaluate the integral as follows:

$$\oint_{C_1} g(z)\,dz = \oint_{C_1}\frac{f_1(z)}{z-z_1}\,dz = 2\pi i\frac{z_1^2}{z_1-z_2}.$$

Doing likewise for the other contour:

$$f_2(z) = \frac{z^2}{z - z_1},$$

$$\oint_{C_2} g(z)\,dz = \oint_{C_2} \frac{f_2(z)}{z - z_2}\,dz = 2\pi i \frac{z_2^2}{z_2 - z_1}.$$

The integral around the original contour *C* then is the sum of these two integrals:

$$\begin{aligned}\oint_C g(z)\,dz &= \oint_{C_1} g(z)\,dz + \oint_{C_2} g(z)\,dz \\ &= 2\pi i \left(\frac{z_1^2}{z_1 - z_2} + \frac{z_2^2}{z_2 - z_1}\right) \\ &= 2\pi i(-2) \\ &= -4\pi i.\end{aligned}$$

An elementary trick using partial fraction decomposition:

$$\oint_C g(z)dz = \oint_C \left(1 - \frac{1}{z - z_1} - \frac{1}{z - z_2}\right) dz = 0 - 2\pi i - 2\pi i = -4\pi i$$

CONSEQUENCES

The integral formula has broad applications. First, it implies that a function which is holomorphic in an open set is in fact infinitely differentiable there. Furthermore, it is an analytic function, meaning that it can be represented as a power series. The proof of this uses the dominated convergence theorem and the geometric series applied to

$$f(\zeta) = \frac{1}{2\pi i} \int_C \frac{f(z)}{z - \zeta}\,dz.$$

The formula is also used to prove the residue theorem, which is a result for meromorphic functions, and a related result, the argument principle. It is known from Morera's theorem that the uniform limit of holomorphic functions is holomorphic. This can also be deduced from Cauchy's integral formula: indeed the formula also holds in the limit and the integrand, and hence the integral, can be expanded as a power series. In addition the Cauchy formulas for the higher order derivatives show that all these derivatives also converge uniformly.The analog of the Cauchy integral formula in real analysis is the Poisson integral formula for harmonic functions; many of the results for holomorphic functions carry over to this setting. No such results, however, are valid for more general classes of differentiable or real analytic functions. For instance, the existence of the first derivative of a real function need not imply the existence of higher order derivatives, nor in particular the analyticity of the function. Likewise, the uniform limit of a sequence of (real) differentiable functions may fail to be differentiable, or may

be differentiable but with a derivative which is not the limit of the derivatives of the members of the sequence.

Another consequence is that if $f(z) = \sum a_n z^n$ is holomorphic in $|z| < R$ and $0 < r < R$ then the coefficients a_n satisfy **Cauchy's inequality**[1]

$$|a_n| \leq r^{-n} \sup_{|z|=r} |f(z)|.$$

GENERALIZATIONS

Smooth functions

A version of Cauchy's integral formula is the Cauchy-Pompeiu formula,[2] and holds for smooth functions as well, as it is based on Stokes' theorem. Let D be a disc in $\mathbf{C}$ and suppose that f is a complex-valued C^1 function on the closure of D. Then (Hörmander 1966, Theorem 1.2.1)

$$f(\zeta) = \frac{1}{2\pi i} \int_{\partial D} \frac{f(z)dz}{z-\zeta} + \frac{1}{2\pi i} \iint_D \frac{\partial f}{\partial \bar{z}}(z) \frac{dz \wedge d\bar{z}}{z-\zeta}.$$

One may use this representation formula to solve the inhomogeneous Cauchy–Riemann equations in D. Indeed, if φ is a function in D, then a particular solution f of the equation is a holomorphic function outside the support of μ. Moreover, if in an open set D,

$$d\mu = \frac{1}{2\pi i} \phi \, dz \wedge d\bar{z}$$

for some $\varphi \in C^k(D)$ ($k \geq 1$), then $f(\zeta, \bar{\zeta})$ is also in $C^k(D)$ and satisfies the equation

$$\frac{\partial f}{\partial \bar{z}} = \phi(z, \bar{z}).$$

The first conclusion is, succinctly, that the convolution $\mu * k(z)$ of a compactly supported measure with the **Cauchy kernel**

$$k(z) = \text{p. v.} \frac{1}{z}$$

is a holomorphic function off the support of μ. Here p.v. denotes the principal value. The second conclusion asserts that the Cauchy kernel is a fundamental solution of the Cauchy–Riemann equations. Note that for smooth complex-valued functions f of compact support on $\mathbf{C}$ the generalized Cauchy integral formula simplifies to

$$f(\zeta) = \frac{1}{2\pi i} \iint \frac{\partial f}{\partial \bar{z}} \frac{dz \wedge d\bar{z}}{z-\zeta},$$

and is a restatement of the fact that, considered as a distribution, $(\pi z)^{-1}$ is a fundamental solution of the Cauchy-Riemann operator $\partial/\partial \bar{z}$.[3] The generalized Cauchy integral formula can

be deduced for any bounded open region X with C^1 boundary ∂X from this result and the formula for the distributional derivative of the characteristic function χ_X of X:

$$\frac{\partial \chi_X}{\partial \bar{z}} = \frac{i}{2} \oint_{\partial X} dz,$$

where the distribution on the right hand side denotes contour integration along ∂X.[4]

Several variables

In several complex variables, the Cauchy integral formula can be generalized to polydiscs (Hörmander 1966, Theorem 2.2.1). Let D be the polydisc given as the Cartesian product of n open discs $D_1, ..., D_n$:

$$D = \prod_{i=1}^{n} D_i.$$

Suppose that f is a holomorphic function in D continuous on the closure of D. Then

$$f(\zeta) = \frac{1}{(2\pi i)^n} \int \cdots \iint_{\partial D_1 \times \cdots \times \partial D_n} \frac{f(z_1, \ldots, z_n)}{(z_1 - \zeta_1) \ldots (z_n - \zeta_n)} \, dz_1 \cdots dz_n$$

where $\zeta = (\zeta_1,...,\zeta_n) \in D$.

In real algebras

The Cauchy integral formula is generalizable to real vector spaces of two or more dimensions. The insight into this property comes from geometric algebra, where objects beyond scalars and

vectors (such as planar bivectors and volumetric trivectors) are considered, and a proper generalization of Stokes theorem.

Geometric calculus defines a derivative operator $\nabla = \hat{e}_i \partial_i$ under its geometric product—that is, for a k-vector field $\psi(\vec{r})$, the derivative $\nabla\psi$ generally contains terms of grade $k+1$ and $k-1$. For example, a vector field ($k = 1$) generally has in its derivative a scalar part, the divergence ($k = 0$), and a bivector part, the curl ($k = 2$). This particular derivative operator has a Green's function:

$$G(\vec{r}, \vec{r}') = \frac{1}{S_n} \frac{\vec{r} - \vec{r}'}{|\vec{r} - \vec{r}'|^n}$$

where S_n is the surface area of a unit ball in the space (that is, $S_2 = 2\pi$, the circumference of a circle with radius 1, and $S_3 = 4\pi$, the surface area of a sphere with radius 1). By definition of a Green's function, $\nabla G(\vec{r}, \vec{r}') = \delta(\vec{r} - \vec{r}')$. It is this useful property that can be used, in conjunction with the generalized Stokes theorem:

$$\oint_{\partial V} d\vec{S} \, f(\vec{r}) = \int_V d\vec{V} \, \nabla f(\vec{r})$$

where, for an n-dimensional vector space, $d\vec{S}$ is an $(n-1)$-vector and $d\vec{V}$ is an n-vector. The function $f(\vec{r})$ can, in principle, be composed of any combination of multivectors. The proof of Cauchy's integral theorem for higher dimensional spaces relies on the using the generalized Stokes theorem on the quantity $G(\vec{r}, \vec{r}') f(\vec{r}')$ and use of the product rule:

$$\oint_{\partial V'} G(\vec{r}, \vec{r}') \, d\vec{S}' \, f(\vec{r}') = \int_V \left([\nabla' G(\vec{r}, \vec{r}')] f(\vec{r}') + G(\vec{r}, \vec{r}') \nabla' f(\vec{r}') \right) \, d\vec{V}$$

when $\nabla \vec{f} = 0$, $f(\vec{r})$ is called a *monogenic function*, the generalization of holomorphic functions to higher-dimensional spaces—indeed, it can be shown that the Cauchy–Riemann condition is just the two-dimensional expression of the monogenic condition. When that condition is met, the second term in the right-hand integral vanishes, leaving only

$$\oint_{\partial V'} G(\vec{r}, \vec{r}') \, d\vec{S}' \, f(\vec{r}') = \int_V [\nabla' G(\vec{r}, \vec{r}')] f(\vec{r}') = -\int_V \delta(\vec{r} - \vec{r}') f(\vec{r}') \, d\vec{V} = -i_n f(\vec{r})$$

where i_n is that algebra's unit n-vector, the pseudoscalar. The result is

$$f(\vec{r}) = -\frac{1}{i_n} \oint_{\partial V} G(\vec{r}, \vec{r}') \, d\vec{S} \, f(\vec{r}') = -\frac{1}{i_n} \oint_{\partial V} \frac{\vec{r} - \vec{r}'}{S_n |\vec{r} - \vec{r}'|^n} \, d\vec{S} \, f(\vec{r}')$$

Thus, as in the two-dimensional (complex analysis) case, the value of an analytic (monogenic) function at a point can be found by an integral over the surface surrounding the

point, and this is valid not only for scalar functions but vector and general multivector functions as well.

NOTES

Hörmander 1983, p. 63,81

Hörmander 1983, pp. 62–63

REFERENCES

1. Ahlfors, Lars (1979), *Complex analysis* (3rd ed.), McGraw Hill, ISBN 978-0-07-000657-7.
2. D. Pompeiu, *Sur la continuité des fonctions de variables complexes*, Annales de la faculté des sciences de Toulouse Sér. 2, 7 no. 3 (1905), p. 265–315
3. Titchmarsh, E.C. (1939), *Theory of functions* (2nd ed.), Oxford University Press
4. Hörmander, Lars (1966), *An introduction to complex analysis in several variabl*

ARTICLE ON ALGAL BLOOM

Veena and Shalini

Baba MastNath University, Asthal Bohar Rohtak- 124021(Haryana). veena.gzb@gmail.com

An algal bloom is a rapid increase or accumulation in the population of algae (typically microscopic) in an aquatic system. Cyanobacteria blooms are often called blue-green algae. Algal blooms may occur in freshwater as well as marine environments. Typically , only one or a small number of phytoplankton species are involved and some blooms may be recognized by discolouration of the water resulting from the high density of pigmented cells.

The fresh water and oceans blooms have been found to be anthropogenic, directly caused by nutrient pollution of nitrogen and phosphates from agriculture and intensive animal farm fertilizers, and household cleaning products. An EPA report in 2013 concluded that changing environmental conditions such as more algae growth is associated with current climate change and may negatively impact the environment human health and the economy for communities across the US and around the world.

BLOOMING

Although there is no officially recognized threshold level, algae can be considered to be blooming at concentrations of hundreds to thousands of cells per millimetre, depending on the severity. Algal bloom concentrations may reach millions of cells per millilitre. Algal blooms are often green, but they can also be other colors such as yellow- brown or red, depending on the species of algae.

Bright green blooms are a result of cyanobacteria such as Microcystis. Blooms may also consists of macro algal (non-phytoplankton) species. These blooms are recognizable by large blades of algae that may wash up onto the shoreline.

Of particular note are harmful algal blooms (HABs), which are algal bloom events involving toxic or otherwise harmful plankton such as dinoflagellates of the genus Alexandrium and Karenia or diatoms of the genus Pseudo-nitzschia. Such blooms often take on a red or brown hue and are known as red tides.

FRESH WATER ALGAL BLOOMS

Fresh water algal blooms are the result of an excess of nutrients, particularly some phosphates. The excess of nutrients may originates from fertilizers that are applied to land for agriculture or recreational purposes. They may also originate from household cleaning products containing phosphorous. These nutrients can then enter watersheds through water runoff. Excess carbon and nitrogen have also been suspected as causes. Presence of residual sodium carbonate acts as catalysts for the algae to bloom by providing dissolved carbon di oxide for enhanced photosynthesis in the presence of nutrients.

When phosphates are introduced into water systems, higher concentrations cause increased growth of algae and plants. Algae tend to grow very quickly under high nutrient availability, but each alga is short-lived and the result is a high concentration of dead organic matter which start to decay. The decay process consumes dissolved oxygen in the water, resulting in hypoxic conditions. Without sufficient dissolved oxygen in the water, animals and plants may die off in large numbers.

Blooms may be observed in freshwater aquariums when fish are overfed and excess nutrients are not absorbed by plants. These are generally harmful for fish and the situation can be correct by changing the water in the tank and then reducing the amount of food given.

HARMFUL ALGAL BLOOMS

A harmful algal bloom (HAB) is an algal bloom that causes negative impacts to other organisms via production of natural toxins, mechanical damage to other organisms or by other means. For example, the toxicity of microcystin produced by cyanobacteria, can harm

water quality or human health. HABs are often associated with large scale mortality events and have been associated with various types of shellfish poisonings.

Examples of common harmful effects of HABs include:

1. The production of neurotoxins which cause mass mortalities in fish, seabirds, sea turtles and marine mammals.
2. Human illness or death via consumption of seafood contaminated by toxic algae.
3. Mechanical damage to other organisms, such as disruption of epithelial gill tissues in fish, Resulting in asphyxiation.
4. Oxygen depletion of the water column (hypoxia or anoxia) from cellular respiration and bacterial degradation.

   Due to their negative economic and health impacts. HABs are often carefully monitored.

INCIDENCE AND DAMAGE BY INSECT PEST ON VEGETABLE CROPS IN ROHTAK

Shalini and Veena Maurya

Department of Zoology, Baba Mastnath University, Rohtak. E-mail: skshalinikhetarpal@gmail.com

Insects are the great majority of known animals and many of them are enormously abundant as individuals. It is important to know when numbers are great enough to justify artificial controls, to be able to evaluate the effectiveness of controls and to have same basis for anticipating outbreaks. Population size is a reflection of feeundity, which is subject to several sources of variation, and of mortality, which may occur at any stage of the life cycle and result from many different environmental or other influences. Large number of young insects can often be tolerated in a crop because natural mortality agents cause significant decrease on pest abundant before the insects become large enough to inflict serious injury. Most of the vegetable crops are short duration crops, can be produced in succession on the same plot and all the family labour of the vegetables grower can be usefully employed throughout the year. In addition vegetable growing add to the income of farmers this is the way the suburban farmers near to the metro cities prefer to grow all types of vegetables for continuous income throughout the year.

Hemchandra and singh (2003) observed at the net reproductive rate was worked out 27.19 per female. Female generation of *Plutella xylostella*. this value of *P. Xylostella* Indicate that the insect can survive and well distributed among the cold crops.

Bijya et al (2006) reported spatial distribution of the aphids *Myzus persical* was studies on cauliflower, *Brassica oleracea* Various indicres of dispersion viz. variance mean ratio, dispersion parwas ammeter -k, mean crowding lioyds index of patchiness Iwao's regression fitting and Taylor's Power law suggested that aphids species exhibited contagious population distribution in this field. Value of Iwao's regression analysis showed that aggregation was mainly due to aphids. colonies rather than individual.

The incidence of jassid started in 2nd week of April with mean initial population of 4.8/3 leaves and reached to its week of May with a mean of 22.3 jassid per 3 leaves and then declined gradually and smallest no. of jassid was found through out the rainy season.

Sinha *et al* (2007) observed the peak incidence of insect pest from mid September to November attributing season and growth of Okra plant or crop as observed by Kumari *et al* (2012).

Mahesh and Men (2007) found that the maximum and minimum temperature showed that positive non significant correlation with fruit and shoot borer population. However Malto and Yadav reported that the pest build up showed a significant positive correlation with the date and growth of okra plant.

Naik et al (2008) reported that the correlation studies indicated non significant relation between the shoot damage and the major parameter. Borer infestation show a significant positive correlation if maximum temperature minimum humidity or rainfall occurs.

MATERIALS & METHODS

Present investigation was carried out at farmers field growing four vegetables (Okra, brinjal, Tomato, and Cabbage) to find out various fauna associated with the four Crops. To record the seasonal incidence of insect pest hundred plants of each Crop were sweeped with insect collecting net at weekly interval starting from initiation of the infestation. Daring the Collection of insects, observation on the damaged plant will also be recorded out of 100 plants.

For biological studies the predominant species of insect in each crop will be brought to the lab and male and female will be kept in a vials and mouth was covered with muslin cloth tied with rubber band. After mating, the insect were taken out of the vials and were kept alone with suitable host leaves for feeding. The dominant species was cultured in under laboratory conditions giving suitable food for multiplication.

For management of insect under field condition 2m * 2m plate has been taken with FYM (Farm yard manure). Local variety of okra has been collected, sown in the 2m * 2m plate.

For Tomato, brinjal and cabbage, seeds were sown in nursery beds and after 30 days seedlings were transplanted in main plate and all agronomical practices such as seeding, irrigation & fertilizer application had given in time.

OBSERVATION

In tomato, higher population (2-5 number of ***Helicoverpa armigera*** and ***Bemisia tabacci***) were recorded during the month of Feb-March when the temperature start rising. In brinjal fruit and shoot borer maximum population was recorded during the month of Oct-Nov. When the fruiting started, whereas in case of cabbage ***Plutella xylostella*** has been found causing maximum damage during month of Jan to March and the highest population was recorded during the month of February. In okra, ***Earias vitella*** causing maximum damage of the crop during summer month from the month of June to August.

As per observation, it was observed that when the crop was 1-2 month old the maximum population of each insect was recorded. This was probably due to maximum numbers of host plants for feeding was available. However, significant correlation was find out with the rise in temperature, rainfall and humidity.

REFERENCES

1. Hemchandra, O. and Singh, T.K. (2003) Life table, rate of increase and stable distribution of Plutella xylostella (Linneaus) on cauliflower. Ann. Pl. Protec. Sci 11(2):269-273.
1. Bijya, p. Debjani, P. Subhrani, S. and Singh, T.K. (2006) spatial distribution of Myzus persica (Sulza) on cauliflower. Ann. Pl. Protec. Sci.14(1): 38-40.
2. Sinha, S.R. Singh, Rai and Sharma, R.K. (2007). Management of insect pest of Okra through Insecticides and inter cropping. Ann. Pl. Protec. Sci. 15:321-324.
3. Mahesh, P. and Menu, U.B. (2007) Seasonal incidences of Leucinodes orbonalis on brinjal, Ann. Pl. Protec. Sci. 15:498-499.
4. Ishar A.K. Bhagat, R.M, Arora A.K and Monobrullah, Md. (200) Effect of abiotic factor on Leucinodes orbonalis in brinjal. Ann. Pl. Protec. Sci. 15:483-484.
5. Naik, V.B, Rao, A.P, Krishnoyya, P.V and rao, V.S. (2008) Seasonal incidences and management of Leucinodes orbonalis Guence on brinjal. Ann. P. Protec. Sci. 16: 329-332.

DEPARTMENT OF PHARMACEUTICAL SCIENCES

ROLE OF PRODRUGS IN SOLUBILITY ENHANCEMENT OF DRUGS

Minakshi Gupta*, Anusha Rohilla

Deptt of Pharm Sciences, Baba Mast Nath University, Asthal Bohar (Rohtak) Haryana e-mail: minakshi.sbmnipsr@gmail.com

ABSTRACT

Solubility is one of the essential parameters to achieve bioavailability of drug in systemic circulation to produce expected therapeutic response. Nearly 40% of new drug molecules face solubility challenge. Though several techniques like micronization, crystal engineering, hydrotrophy, solid dispersion and so forth were available for solubility enhancement of poorly soluble drugs based on physical and chemical modification, prodrug approach is a vital technique amongst the other. Prodrugs are biologically inactive compound which can be metabolized in the body to produce active drug. It is estimated that currently about 10% of all world-wide approved drugs are prodrugs. This review extensively confers the role of prodrugs and the kind of promoieties used for solubility enhancement.

INTRODUCTION

The poor solubility and low dissolution rate of poorly water soluble drugs in the aqueous gastrointestinal fluids often cause insufficient bio availability. The major challenge with the design of oral dosage forms lies with their poor bioavailability. Low aqueous solubility is the major problem encountered with formulation development of new chemical entities as well as generic development. More than 40% NCEs (new chemical entities) developed in pharmaceutical industry are practically insoluble in water. As for BCS class-II drugs rate limiting step is drug release from the dosage form and solubility and not the absorption, so increasing the solubility in turn increases the bioavailability for BCS class-II drugs[1]. Prodrug is biologically inactive compound which can be metabolized in the body produce drug.[2] The prodrug approach is emerged tool to overcome the various obstacles in drug formulation and targeting the chemical instability, low aqueous solubility, local irritation and toxicity.[3] The term "prodrug" was introduced 55 years ago in 1958 by Albert & started to gain the popularity in 1960's.[4] Albert described "prodrug" or "proagent" as a pharmacologically inactive compound that is transformed by the mammalian system into an active substance by either chemical (non enzymatic) or metabolic (enzymatic) means. Later these compounds have been referred to as the "latentiated drugs" and the process as "drug latentiation" by Harper, which he described as the chemical modification of a biologically active compound to form a new compound which, upon in vivo enzymatic attack liberates the parent compound.[5] These compounds were also called as "congeners" and "bioreversible derivatives", but 'prodrug' is now the most commonly accepted term.[6-7] The first prodrugs were not necessarily meant to be prodrugs; this feature was unintentional. The antibiotic prontosil produced by Bayer was found to release the active agent sulfanilamide by reductive enzymes. Later it was marketed in 1935.[8]

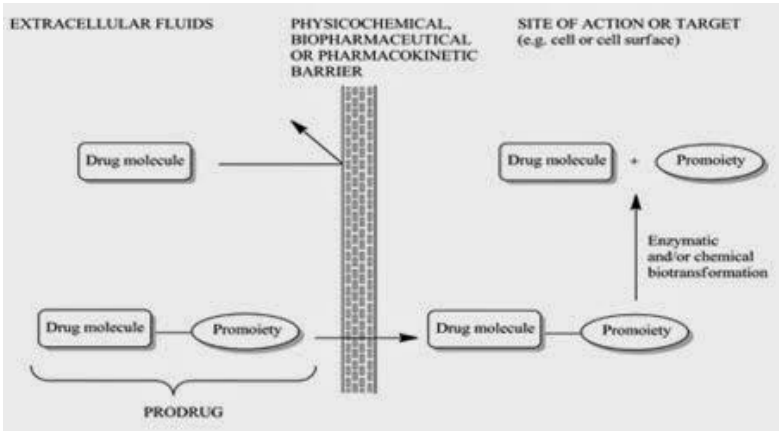

Figure No.1. A simplified illustration of the prodrug concept

CLASSIFICATION

Depending upon the constitution, lipophilicity, method of bioactivation and the catalyst involved in bioactivation, prodrugs are classified into two categories; Carrier linked prodrug and Bioprecursor prodrug.[9]

A. Carrier linked prodrugs: Carrier linked prodrug consists of the active drug covalently linked to an inert carrier or transport moiety, generally ester or amide. Such prodrugs have greatly modified lipophilicity due to the attached carrier. The active drug is released by hydrolytic cleavage either chemically or enzymatically. The Prodrug and carrier released after in vivo enzymatical or non-enzymatical attack must be nontoxic.[10]

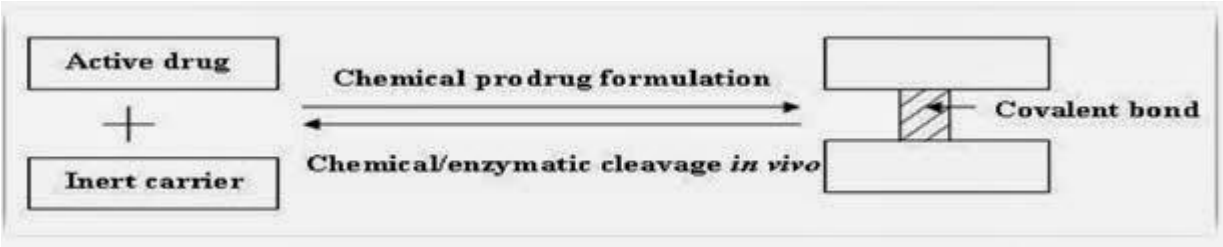

Figure No.2. Carrier linked prodrug

B. Bioprecursors: They are inert molecules obtained by chemical modification of the active drug but do not contain a carrier. Such a moiety has almost the same lipophilicity as the parent drug and is bioactivated generally by redox biotransformation only enzymatically. e.g. Aryl acetic acid NSAID such as Fenbufen from aroyl propionic acid precursors.

Carrier linked prodrug consists of the attachment of a carrier group to the active drug to alter its physicochemical properties. The subsequent enzymatic or non-enzymatic mechanism releases the active drug moiety. Hence, the carrier linked prodrugs have a major drawback that they are linked through covalent linkage with specialized nontoxic protective groups or carriers or promoieties in a transient manner to alter or eliminate undesirable properties in the parent molecule.

Depending upon the nature of carrier, the carrier linked prodrug may further be classified into,[11]

Double prodrug: Double prodrug also termed as 'Pro-prodrug', is a prodrug further derivatized in a fashion such that only enzymatic conversion to prodrug is possible before the later can cleave to release the active drug. e.g. Cefpodoxime proxetil.

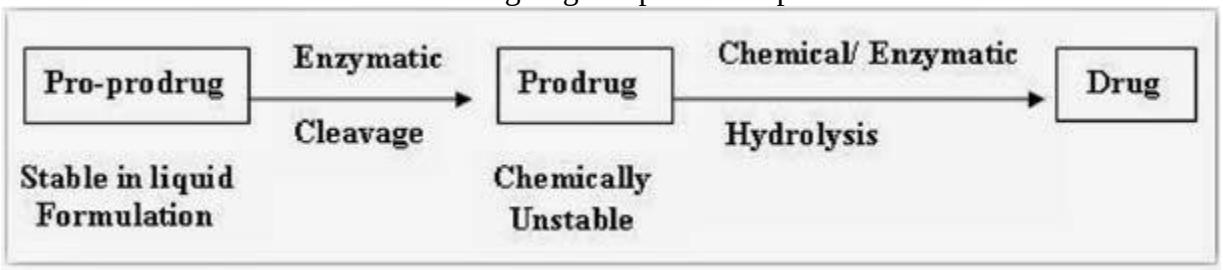

Figure No.3. Double prodrug

Macromolecular prodrug: Macromolecules like polysaccharides dextrans, cyclodextrins, proteins, peptides and polymers may be used as carriers to form the macromolecular prodrugs. e.g. Naproxen-2-glyceride.

Site specific prodrug: In this approach, prodrug is designed using a carrier which acts as a transporter of the active drug to a specified targeted site. e.g. Progabide-Diethyl stilbesterol.[12]

Mutual Prodrug: A mutual prodrug consists of two pharmacologically active agents coupled together so that each acts as a promoiety for the other agent and vice versa.[13] e.g. Estramustine is a mutual prodrug composed of a phosphorylated steroid (17- α- estradiol) linked to Nor-mustard, an anti-androgenic drug through a carbamate linkage. The steroid

portion of the molecule helps to concentrate the drug in prostrate. The carrier selected may have the same biological action as that of the parent drug to give synergistic action or some additional biological action that is lacking in the parent drug, thus ensuring some additional benefits. The carrier can also be a drug that might help to targetthe parent drug to a specific site or organ or cells or may improve site specificity of a drug. The carrier drug may be used to overcome some side effect of the parent drugs as well.[14]

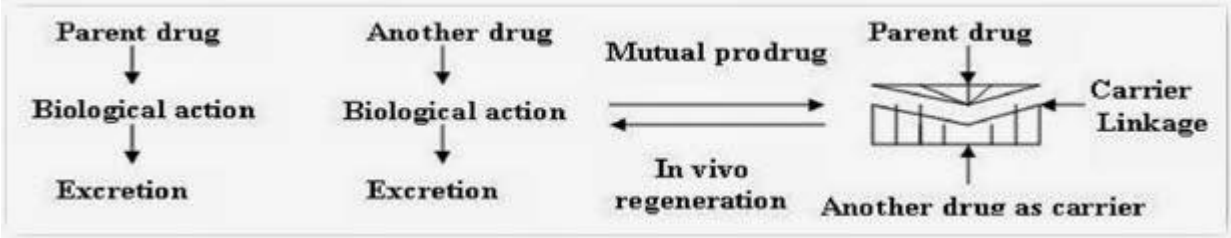

Figure No.4. Mutual prodrug

CONCLUSION

The prodrug strategy is one of the most promising approaches to enhance the solubility and in turn bioavailability of drug. Introduction of new chemical entity in the market is very expensive and time consuming process. Thus R&D oriented pharma companies are trying to alleviate solubility problems of existing drugs. Prodrugs remain an effective tool to overcome those barriers. Hence legitimate choice of promoieties surmounts pharmacokinetic barrier in formulating a chemical entity. In prospective view emergence of prodrug is a boon for pharma industry and to mankind.

REFERENCES

1. Sharma D., Soni M., Kumar S., and Gupta G.D; Solubility enhancement- eminent role in poorly soluble drugs; Research Journal of Pharmacy and Technology; 2009; 2(2); 220–224
2. Kristiina Huttunen M., Hannu Raunio., and Jarkko Rautio; Prodrugs-from Serendipity to Rational Design; Pharmacol Rev; 2011; 63(3); 750-771
3. Higuchi T., Stella V., Pro-drugs: An overview and definition, ACS Symposium Series, Washington DC, Ame. Chem. Soc; 1975; 1-115
4. Albert A; Chemical aspects of selective toxicity, Nature; 1958; 182; 421-422
5. Bentley R; Different roads to discovery; Prontosil (hence sulfa drugs) and penicillin (hence beta-lactams). J. Ind. Microbiol Biotechnol; 2009; 36; 775–786
6. Harper N.J; Drug Latentiation; Journal of Medicinal Chemistry; 1959; 1; 467-500
7. Kalgutkar A.S., Marnett A.B., Crews B.C., Remel R.P., Marnet L.J; J. Med. Chem; 2000; 43(15); 2860-2870
8. Timmins G.S., and Deretic V; Mechanisms of action of isoniazid; Mol. Microbiol. 2006; 62; 1220–1227
9. Singh G., Wadhwa L.K., Sharma S; J. Sci. Ind. Res; 1996; 55; 497-510
10. Bhosale D., Bharambe S., Gairula N., and Dhaneshwar S.S; Ind. J. Pharm. Sci; 2006; 68(3); 286-294
11. Kopecek J., and Duncan R; Targetable Polymeric Prodrugs; J. Contr. Rel; 2003; 6; 315–337
12. Fleisher D; Improved oral drug delivery solubility limitations overcome by the use of prodrugs; Adv. Drug Delivery Rev; 1996; 115
13. Wise R; After pro-drugs-mutual pro-drugs; J. Antimicrob. Chemother; 1981; 7, 503-504
14. Bhosle D., Bharambe S., Gairola N., and Dhaneshwar S.S; Mutual prodrug concept: Fundamentals and applications; Indian Journal of Pharmaceutical Sciences; 2006; 68; 286-294.

OBESITY – IS FOOD, A FRIEND OR ENEMY?

Balvinder Singh*, Pawan Jalwal, Rajiv Kumar Arora

Deptt. of Pharma. Sciences, Baba Mastnath University, Asthalbohar (Rohtak) Haryana. E-mail-balvindersinghpharmaco@gmail.com

ABSTRACT

Worldwide obesity has more than doubled since 1980. In 2014, more than 1.9 billion adults, 18 years and older, were overweight. Of these over 600 million were obese. 39% of adults aged 18 years and over were overweight in 2014, and 13% were obese. Most of the world's population live in countries where overweight and obesity kills more people than underweight. 42 million children under the age of 5 were overweight or obese in 2013. Obesity is preventable.

KEYWORDS: Obesity, overweight, Body Mass Index (BMI), fats, exercise.

INTRODUCTION

The word Obesity is a Latin word *obesitas*, which means "stout, fat, or plump". *The Oxford English Dictionary* documents its first usage in 1611 by Randle Cotgrave. Obesity is a medical condition in which excess body fat has deposited to the extent that it has a negative effect on health, leading to reduced life expectancy and/or increased health problems.[1][2] Obesity is a result of excessive food energy intake, lack of physical activity, and genetic susceptibility, although a few cases are caused primarily by genes, endocrine disorders, medications, or psychiatric illness. Limited evidence is there to support the view that some obese people eat little yet gain weight due to a slow metabolism. On an average, obese people have a greater energy expenditure than their thin counterparts due to the energy required to maintain an increased body mass.[3][4]

The two main treatments for obesity are Dieting and Exercising. Diet quality can be improved by reducing the consumption of high energy foods, such as fat and sugars, and by increasing the intake of dietary fiber. If diet, exercise, and medication are not effective, a gastric balloon may assist with weight loss, or surgery may be performed to reduce stomach volume and/or bowel length, leading to feeling full earlier and a reduced ability to absorb nutrients from food.[5][6]

It is one of the most serious public health problems of the 21st century.[7] Obesity is stigmatized in much of the modern world (particularly in the Western world), though it was widely seen as a symbol of wealth and fertility at other times in history and still is in some parts of the world.[2][8] In 2013, the American Medical Association classified obesity as a disease.[9][10]

In Western countries, people are considered obese when their body mass index (BMI),[11] exceeds 30 kg/m^2.

$$\mathrm{BMI} = \frac{m}{h^2}$$

where *m* and *h* are the subject's weight and height respectively.
Body mass index (BMI) is a simple index of weight-for-height. It is defined as a person's weight in kilograms divided by the square of his height in meters (kg/m^2).
The WHO definition is:

- a BMI greater than or equal to 25 is overweight
- a BMI greater than or equal to 30 is obesity.

FACTS ABOUT OBESITY AS PER WHO

In 2014, more than 1.9 billion adults, 18 years and older, were overweight. Of these over 600 million were obese. Overall, about 13% of the world's adult population (11% of men and 15% of women) were obese in 2014. In 2014, 39% of adults aged 18 years and over (38% of men and 40% of women) were overweight. The worldwide prevalence of obesity more than doubled between 1980 and 2014. In 2013, 42 million children under the age of 5 were overweight or obese. Overweight and obesity are linked to more deaths worldwide than underweight. [12]

CAUSES

- an increased intake of energy-rich foods that are high in fat and sugar; and
- an increase in physical inactivity due to the increasingly sedentary nature of many forms of work.

HEALTH CONSEQUENCES

- cardiovascular diseases (mainly heart disease and stroke), which were the leading cause of death in 2012;
- diabetes;
- musculoskeletal disorders (especially osteoarthritis - a highly disabling degenerative disease of the joints);
- some cancers (endometrial, breast, and colon).

PREVENTION

- limit energy intake from total fats and sugars;
- increase consumption of fruit and vegetables, as well as legumes, whole grains and nuts;
- engage in regular physical activity (60 minutes a day for children and 150 minutes per week for adults).

CONCLUSION

Increased dietary intake, sedentary lifestyle, reduced physical exercises makes our life full of disorders and diseases. The need of the hour is to use public transport, stairs, go to parks for walking for at least an hour, use less of Airconditioners, Junk Foods, fried items, sugar rich sweets, Aerated Drinks etc. We have to come out of our comfort zones, if we wish to be healthy and want to spend less on the treatment of obesity.

REFERENCES

1. WHO 2000 p.6
2. Haslam DW, James WP (2005). "Obesity". *Lancet* (Review) 366 (9492): 1197–209. doi:10.1016/S0140-6736(05)67483-1. PMID 16198769.

3. Kushner, Robert (2007). *Treatment of the Obese Patient (Contemporary Endocrinology)*. Totowa, NJ: Humana Press. p. 158. ISBN 1-59745-400-1. Retrieved April 5, 2009.
4. Adams JP, Murphy PG (July 2000). "Obesity in anaesthesia and intensive care". *Br J Anaesth* 85 (1): 91–108. doi:10.1093/bja/85.1.91. PMID 10927998.
5. NICE 2006 p.10–11
6. Imaz I, Martínez-Cervell C, García-Alvarez EE, Sendra-Gutiérrez JM, González-Enríquez J (July 2008). "Safety and effectiveness of the intragastric balloon for obesity. A meta-analysis". *Obes Surg* 18 (7): 841–6. doi:10.1007/s11695-007-9331-8. PMID 18459025.
7. Barness LA, Opitz JM, Gilbert-Barness E (December 2007). "Obesity: genetic, molecular, and environmental aspects". *American Journal of Medical Genetics* 143A (24): 3016–34. doi:10.1002/ajmg.a.32035. PMID 18000969.
8. Woodhouse R (2008). "Obesity in art: A brief overview". *Front Horm Res*. Frontiers of Hormone Research 36: 271–86. doi:10.1159/000115370. ISBN 978-3-8055-8429-6. PMID 18230908.
9. Pollack, Andrew (June 18, 2013). "A.M.A. Recognizes Obesity as a Disease". *New York Times*. Archived from the original on June 18, 2013.
10. Weinstock, Matthew (June 21, 2013). "The Facts About Obesity". *H&HN*. American Hospital Association. Retrieved June 24, 2013.
11. WHO 2000 p.9
12.Fact sheet N°311, Updated January 2015

QUANTITATIVE STRUCTURE ACTIVITY RELATIONSHIP (QSAR)

Arun Kumar, Dr. Rajiv Tonk, Pawan Jalwal

Faculty of Pharmaceutical Sciences, Baba Mastnath University, Asthal Bohar, Rohtak-124021

INTRODUCTION

QSAR involves the derivation of mathematical formula which relates the biological activities of a group of compounds to their measurable physicochemical parameters. These parameters have major influence on the drug's activity. QSAR derived equation take the general form:
Biological activity= function (parameters)
Activity is expressed as log (1/c). C is the minimum concentration required to cause a defined biological response. Quantitative structure–property relationships (QSPRs), are mathematical models that attempt to relate the structure-derived features of a compound to its biological or physicochemical activity. Similarly, quantitative structure–toxicity relationship (QSTR) or quantitative structure–pharmacokinetic relationship (QSPkR) is used when the modeling applies on toxicological or pharmacokinetic systems. QSAR (also QSPR, QSTR, and QSPkR) works on the assumption that structurally similar compounds have similar activities. Therefore, these methods have predictive and diagnostic abilities. They can be used to predict the biological activity (e.g., IC50) or class (e.g., inhibitor versus noninhibitors) of compounds before the actual biological testing. They can also be used in the analysis of structural characteristics that can give rise to the properties of interest.

PARAMETERS ‰

The parameter is the measure of the potential contribution of its group to a particular property of the parent drug. Various parameters used in QSAR studies are

1. Lipophilic parameters: partition coefficient, π-substitution constant
2. Polarizability parameters: molar refractivity, parachor
3. Electronic parameters: Hammet constant, dipole moment.
4. Steric parameters: Taft's constant.
5. Miscellaneous parameters: molecular weight, geometric parameters.

LIPOPHILIC PARAMETERS

Lipophilicity is partitioning of the compound between an aqueous and non-aqueous phase. Partition coefficient:

- P = [drug] in octanol / [drug] in water
- Typically over a small range of log P, e.g. 1-4, a straight line is obtained e.g. log 1/C = 0.75 log P + 2.30
- If graph is extended to very high log P values, then get a parabolic curve log 1/C = - k1 (log P) 2 + k 2 log P + k3
- When P small, dominated by log P term
- When P large, log P squared dominates & so activity decreases

π -SUBSTITUENT CONSTANT OR HYDROPHOBIC SUBSTITUENT CONSTANTS:

- The π-substituent constant defined by hansch and co-workers by the following equation.
- px = log Px - log PH
- A positive π value indicates that the π substituent has a higher lipophilicity than hydrogen and the drug favours the organic phase.
- A negative π value indicates that the π substituent has a lower lipophilicity than hydrogen and the drug favours the aqueous phase.

ELECTRONIC PARAMETERS

- The Hammett constant (σ); s x = log (K x/Kbenzoic)
- Electron Withdrawing Groups
- Equilibrium shifts Right & K x > Kbenzoic
- Since s x = log K x – log Kbenzoic, then s will be positive .
- Hammett constant takes into account both resonance and inductive effects; thus, the value depends on whether the substituent is para or meta substituted -ortho not measured due to steric effects.

STERIC SUBSTITUTION CONSTANT

- It is a measure of the bulkiness of the group it represents and it effects on the closeness of contact between the drug and receptor site. much harder to quantitate
- Examples are:
- Taft's steric factor (Es) (~1956), an experimental value based on rate constants
- Molar refractivity (MR)--measure of the volume occupied by an atom or group--equation includes the MW, density, and the index of refraction--
- Verloop steric parameter--computer program uses bond angles, van der Waals radii, bond lengths.

HANSCH ANALYSIS

- Proposed that drug action could be divided into 2 stages: 1) Transport & 2) Binding
- Each of these stages depend upon the physical and chemical properties of the drug.
- Log 1/C = k1P = k 2 P 2 + k 3s + k 4Es + k 5
- Look at size and sign for each component of the equation.
- Values of r << 0.9 indicate equation not reliable.
- Accuracy depends on using enough analogs, accuracy of data, & choice of parameters Applications: used to predict the activity of an as yet unsynthesized analouge.

FREE WILSON ANALYSIS

- This method is based on the assumption that the introduction of a particular substituent at a particular molecular position , always leads to a quantitatively similar effect on biological potency of the whole molecules and expressed by the equation as
- BA= μ + Σ aj
- Application:
- Easy to apply
- Simple method
- The substituent which can not fulfill the principle of additivity can be recognized
- Effective when substituent constants are not available.

TOPLISS METHOD ™

- This approach is completely non-mathematical and nonstatistical and does not need computerization of the data. ™
- A Topliss scheme is a flow diagram that in a series of steps directs the medicinal chemist to produce a series of analogues, some which have greater activity than lead used to start the tree. ™ There are two topliss schemes
    1. For the aromatic substituents
    2. For the aliphatic side chain substituents.
- Applications: This method can be used if synthetic route might be difficult and only a very few structures can be made in a limited time

CONCLUSION

Despite all the successful stories & methods in QSAR modeling, in our opinion, computational methods may not and should not completely replace conventional in vitro or in vivo testing methods, but should be further developed as an important and essential

complementary tool in the drug development process. Moreover, there are still many research opportunities in the areas of data harmonization, model applicability domain, model validation, inclusion of nonstructural information in modeling, similarity or pharmacophore research, consensus modeling. All QSAR models & methods developed should be presented with the following information : . A defined endpoint;. An unambiguous algorithm; . A defined domain of applicability; . Appropriate measures of goodness-of-fit, robustness and predictivity; and. a mechanistic interpretation if possible. In essence, the guidelines demonstrate the importance of a transparent validation process and reliable QSAR for its acceptance in regulatory context. The readers are to refer to the guideline document, available on the OECD website (http://www.oecd. org), for a comprehensive explanation and interpretation of the guidelines.

REFERENCES

1. Snodin, D.J. (2002) Toxicology Letters, 127, 161–168.
2. Kraljevic, S., Stambrook, P.J., and Pavelic, K. (2004) EMBO Reports, 5, 837–842.
3. Yap, C.W., Xue, Y., Li, Z.R., and Chen, Y.Z. (2006) Current Topics in Medicinal Chemistry, 6, 1593–1607.
4. Guido, R.V., Oliva, G., and Andricopulo, A.D. (2008) Current Medicinal Chemistry, 15, 37–46.
5. Schwaighofer, A., Schroeter, T., Mika, S., and Blanchard, G. (2009) Combinatorial Chemistry and High Throughput Screening, 12, 453–468.
6. Valerio, L.G. Jr. (2009) Toxicology and Applied Pharmacology, 241, 356–370.
7. Worth, A.P., Bassan, A., De Bruijn, J., Gallegos Saliner, A., Netzeva, T., Patlewicz, G., Pavan, M., Tsakovska, I., and Eisenreich, S. (2007) SAR and QSAR in Environmental Research, 18, 111–125.
8. Lilienblum, W., Dekant, W., Foth, H., Gebel, T., Hengstler, J., Kahl, R., Kramer, P.J., Schweinfurth, H., and Wollin, K.M. (2008) Archives of Toxicology, 82, 211–236.
9. Yap, C.W. and Chen, Y.Z. (2005) Journal of Pharmaceutical Sciences, 94, 153–168.
10. Xue, Y., Li, H., Ung, C.Y., Yap, C.W., and Chen, Y.Z. (2006) Chemical Research in Toxicology, 19, 1030–1039.
11. Liew, C.Y., Ma, X.H., Liu, X., and Yap, C.W. (2009) Journal of Chemical Information and Modeling, 49, 877–885.
12. Liew, C.Y., Ma, X.H., and Yap, C.W. (2010) Journal of Computer-Aided Molecular Design, 24, 131–141.
13. Todeschini, R. and Consonni, V. (2000) Handbook of Molecular Descriptors, vol. 11, Wiley-VCH, Weinheim.
14. Guyon, I. and Elisseeff, A. (2003) The Journal of Machine Learning Research, 3, 1157–1182.
15. Tropsha, A., Gramatica, P., and Gombar, Vijay K. (2003) The QSAR & Combinatorial Science, 22, 69–77. Schultz, T.W., Netzeva, T.I., and Cronin, M.T. (2003) SAR and QSAR in Environmental Research, 14, 59–81.
16. Scior, T., Medina-Franco, J.L., Do, Q.T., Martinez-Mayorga, K., Yunes Rojas, J.A., and Bernard, P. (2009) Current Medicinal Chemistry, 16, 4297–4313.
17. Topliss, J.G. and Edwards, R.P. (1979) Journal of Medicinal Chemistry, 22, 1238–1244.
18. Benigni, R., Giuliani, A., Franke, R., and Gruska, A. (2000) Chemical Reviews, 100, 3697–3714.
19. Cash, G.G. (2001)Mutation Research, 491, 31–37.
20. Abraham, M.H., Hassanisadi, M., JalaliHeravi, M., Ghafourian, T., Cain, W.S., and Cometto-Muniz, J.E. (2003)
21. OECD, Guidance Document on the Validation of (Quantitative) Structure– Activity Relationships Models, Organisation for Economic Co-operation and Development (2007).

DNA FINGER PRINTING OF HERBALS

Upma, Shilpi Arora

Department of Pharmaceutical Sciences and research, Baba Mastnath University, Asthal bohar, Rohtak

Herbal medicine refers to the long historical use of medicinal plants, which is well established and widely known to be safe and effective. Moreover, and nowadays, approximately 70% of "synthetic" medicines are derived from medical plants. But, with no doubt,the adulteration and/or substitution of medicinal plants remain the main problem treating the drug industry worldwide. Adulteration is defined as an intentional substitution or addition of another plant species or foreign substance, in order to increase the weight, potency of the product, and/or to decrease its cost. For many decades, many methods were elaborated for the differentiation of the spurious and authentic plants. Most of these methods were based on plant morphological observations, microscopy and chemical markers. The limitations of these methods generated a need for newer and supportive techniques.

DNA FINGERPRINTING:

Deoxyribonucleic acid (DNA) is the fundamental building component of all living cells. Our characteristics, traits and physical features are determined by the specific arrangement of DNA base-pair sequences in the cell. It is this distinct arrangement of adenine, guanine, thymine and cytosine (called DNA nucleotides) that regulates the production of specific proteins and enzymes . It is this specific 3-D arrangement of DNA that confers upon us our uniqueness in this world. DNA forms the basic genotype (genetic identity) of an organism, which in turn determines the phenotype (physical features) of the organism.

This concept of fingerprinting has been increasingly applied in the past few decades to determine the ancestry of plants, animals and other microorganisms. Genotypic characterization of plant species and strains is useful as most plants, though belonging to the same genus and species, may show considerable variation between strains. A good example of this is the fraudulent adulteration of Chianti wines with inferior quality grapes 1.This is also the case with medicinal plants, where the amounts of active chemicals may vary from plant to plant. Factors such as soil, climate and adaptability dictate the viability of a particular species and subsequently its drug content. In such cases, there are observed variations in the genetic composition of the plant, in addition to varying amounts of the active drug compound. When used commercially, two factors affect the final drug quality:

(1) The variability with respect to strain-specific drug content.

(2) The potential adulteration of plant drugs with extracts from plants that have lower drug content.

Such discrepancies are very difficult to detect using conventional methods of morphology and microscopy. Cinchona bark, from which quinine is obtained, is One case where the DNA fingerprinting technique could be useful. The bark of Cinchona grown in the plains contains quinine, which is therapeutically active. The same species of tree grown on hilltops and slopes looks morphologically similar but has no active quinine. Fingerprinting of DNA is dictated by several factors; sequence or restriction site data, taxonomic level of study, the level at which the study is being done (species, genera, etc.), robustness and reproducibility of the method, effectiveness in terms of cost and time, and availability of DNA.

TYPES OF DNA BASED MARKERS:

Various types of DNA based techniques are available to evaluate DNA polymorphism. These are hybridization based methods, polymerase chain reaction (PCR) based methods and sequencing based methods. 18 Generally accepted classification of markers is as follows: 19,20 Hybridization based • Restriction Fragment Length Polymorphism (RFLP) • Variable Number Tandem Repeat (VNTR) • Probe hybdriduzation with Micro and Minisatellite •

Random Genomic Clone • cDNA Clone PCR based • Inter Simple Sequence Repeat (ISSR) • Random Amplification Polymorphic DNA (RAPD)/Arbitrary Primed PCR • Amplified Fragment Length Polymorphism (AFLP) • DNA Amplification Fingerprinting (DAF) Sequence based • Simple Sequence Repeats (SSR) • Sequence Characterized Amplified Region (SCAR) • Cleaved Amplified Polymorphic Sequence (CAPS) • Single Nucleotide Polymorphism (SNP).

METHODOLOGY OF DNA FINGERPRINTING:

Few techniques in molecular biology have received so much attention and popular acceptance as PCR. Invented by Kary Mullis in 1983, PCR is a method used to generate billions of copies of genomic DNA within a very short time. This amplification is useful in criminal cases where there are miniscule amounts of DNA available. Today PCR finds application in almost all aspects of biomedicine. PCR has been used for the detection of many pathogenic organisms, from bacteria to viruses.

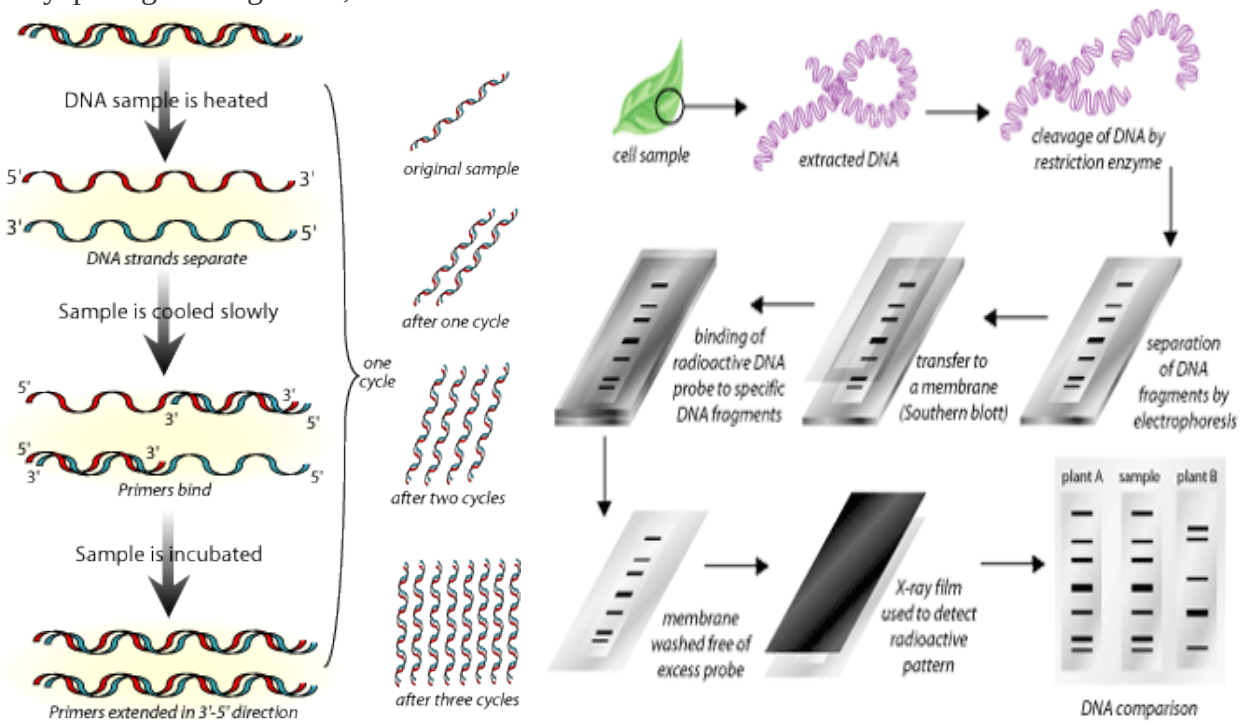

Fig :1The Polymerase Chain Reaction is used to amplify a sample of DNA.

Fig 2: RFLP is one of the DNA fingerprinting techniques that is used to determine plant strain and purity in nutraceutical and herb production.

1. Microsatellites are simple sequence repeats (SSRs), 1 to 6 nucleotides in length, which show a high degree of polymorphism. Specific microsatellites can be isolated using hybridized probes followed by their sequencing. Like any DNA fragment, SSRs can be detected by specific dyes or by radiolabelling using gel electrophoresis. The advantage of using SSRs as molecular markers is the extent of polymorphism shown, which enables the detection of differences at multiple loci between strains [3].Coupled with chemical and morphological data, we can identify the plant species or strain of interest. The main advantage of using SSRs for fingerprinting is that small amounts of DNA are required compared to the restriction fragment length polymorphisms (RFLP) method. This is due to the large amounts of SSRs present in any genome. Further, assays involving SSRs are more robust than random amplified polymorphic DNA (RAPDs), making them up to seven times more efficient. A drawback to using SSRs is the need to develop separate SSR primer sets for each species. The latest research suggests that SSRs will be involved in new methods of detection of alterations of specific sequences in the DNA.

2. Restriction fragment length polymorphisms are unequal lengths of DNA fragments obtained by cutting Variable Number of Tandem Repeat (VNTRs) sequences up to 30 sequences long with restriction enzymes at specific sites. VNTRs vary between plant species, as do the number and location of restriction enzyme-recognition sites. On an agarose gel, RFLPs can be visualized using radiolabeled complementary DNA sequences. There is no need for PCR amplification of DNA in this method. A routine southern blot experiment is used instead. Normally, RFLPs are used to identify the origins of a particular plant species, setting the stage for mapping its evolution. There are some problems with the RFLP method of DNA fingerprinting. First, the results do not specifically indicate the chance of a match between two organisms. Secondly, the process involves a lot of money and labor, which not many laboratories can afford. Finally, unlike the microsatellites, a few loci in the assay must suffice.

3. Amplified fragment length polymorphism (AFLP) is a PCR-based derivative method of RFLP in which sequences are selectively amplified using primers. It is a reliable and efficient method of detecting molecular markers. DNA is cut with two restriction enzymes to generate specific sequences, which are then amplified suitably. The mere addition or deletion of bases at the 3′ end determines the selectivity and complexity of the amplification 4. By using AFLP, it is possible to evaluate more loci than with RFLP or RAPD. AFLP is also capable of determining a large number of polymorphisms. Similar to SSRs, AFLP-based assays are cost-effective and can be automated.

4. Random amplified polymorphic DNA is one of the most commonly used primary assays for screening the differences in DNA sequences of two species of plants. RAPD consists of fishing for the sequence using random amplification. Here, plant genomic DNA is cut and amplified using short single primers at low annealing temperatures, resulting in amplification at multiple loci. By running a 2-dimensional electrophoresis gel, it is possible to determine the change in sequence pattern by superimposing the 2 gels. Once the band of interest is identified, the gel is cut, and the DNA is isolated and sequenced. Using this target, DNA from other cultivars can be assessed using other techniques such as AFLP or SSRs. It is also more cost effective than RFLPs. RAPDs lack specificity, however, due to low annealing temperatures and easier reaction conditions.

5. Other Methods include the use of single nucleotide polymorphs (SNPs) DNA amplification fingerprinting (DAF) and their offshoots. Although these techniques vary slightly from each other, they operate on the same principle.

APPLICATION

The practical applications of DNA fingerprinting are numerous. DNA fingerprints are used in pedigree analysis and establishing paternity and maternity. The patterns are so specific that half of the DNA fragments (RFLPs) will be common with those of the father and half with those of the mother (see figure). This is because the person inherits his or her RPLP pattern from his or her parents. It is interesting to know that a parental RFLP pattern can be reconstructed even if only the children's RFLP patterns are available.

CONCLUSION

DNA fingerprinting, apart from identifying alterations in the genotypes of plant species, is also used for the betterment of drug-yield by tissue culturing. DNA of interest can be stored as germplasm, which is then used for future cultivation.

DNA fingerprinting of herbal drugs, though still in its early years, seems to be a promising tool for the authentication of medicinal plant species and for ensuring better quality herbs and nutraceuticals.

These markers have most effective utility in quality control of commercially important medicinal herbs which are adulterated. And these tools are also utilised in any form of the drug i.e., processed or unprocessed.

REFERENCES:

1. Inamdar, N., S. Edalat, V.B. Kotwal and S. Pawar, Herbal Drugs in Milieu of Modern Drugs,Int. J. Green Pharm, 2008, Vol 2 (1),2-8.
2. Kalpana Joshi, Preeti Chavan, Dnyaneshwar Warude and Bhushan Patwardhan, Molecular markers in herbal drug technology, Current science, 2004, 87:2,159.
3. Quality control methods for medicinal plant material, World health organization guidelines, 1998.
4. Shinde V.M et.al, Application of quality control principles to herbal drugs, International Journal of Phytomedicine, 2009, 1, pp 4-8.
5. Breithaupt, H. (2003) Back to the roots EMBO Rep 4(1): 10-12.
6. Mihalov, J. J., Marderosian, A. D., and Pierce, J. C. (2000) DNA identification of commercial ginseng samples J Agric Food Chem 48(8)3744-3752.
7. Yadav et.al; Development of Randomly Amplified Polymorphic DNA-Sequence Characterized Amplified Region Marker for Authentication of Swertia chirayita, An Endangered Himalayan Plant, Journal of Plant Molecular Biology & Biotechnology, 2012,3:1, p 27.
8. Kanika Bamhania et.al; Genetic Variability Analysis Using ISSR Markers in Withania somnifera L. Dunal Genotypes from Different Regions, Journal of Herbs, Spices & Medicinal Plants, 2013,19:1.
9. Ghosh,S. et.al, Species-specific AFLP markers for identification of Zingiber officinale, Z. montanum and Z. zerumbet (Zingiberaceae), Genetics and Molecular Research, 2100 , 10:1, pp 218-229.
10. Kurane,J., Shinde,V., Harsulkar,A., Application of ISSR marker in pharmacognosy: Current update, Phcog Rev, 2009, 3, pp216-228.
11. Kumar L.D., Kathirvel M., Rao G.V and Nagaraju J., DNA profiling of disputed chilli samples (Capsicus annum) using ISSR-PCR and FISSR-PCR marker assays. Forensic Sci. Int., 2001,116,p.no 63–68.
12. Joshi S.P, Ranjekar PK and Gupta V S, Molecular markers in plant genome analysis, curr.sci, vol-77, 1999; pp230-240.

RECENT ADVANCES ON PACKAGING TECHNOLOGY

Pawan Jalwal*, Balvinder Singh, Arun Kumar

Faculty of Pharmaceutical Sciences, Baba Mastnath University, Email: pawan_jalwal@rediffmail.com

ABSTRACT

Packaging may be defined as the science, art and technology of protecting the pharmaceutical products for storage, presentation, identification, information, transportation, distribution, sales, uses and convenience to encourage compliance with a course of therapy etc. It may also define as the process of design, evaluation and production of packages. Packaging plays an important role in pharmaceutical sector by protecting the dosage forms from environmental factors like physical as well as chemical stimuli. Packaging acts as a barrier for the physical and chemical stability of the medicine from light, moisture, oxygen, bacteria, and volatiles. It is also prevent the product from mechanical trauma.

The commonly used packaging materials are container, closure, carton or outer and box. The containers may be made up of glass, plastic, metal or paper. The material for closure may include cork, glass, plastic, metal and rubber. There are various tests for determination of quality, integrity and compatibility of packaging materials. The specification and requirement of quality testing depends on type of pharmaceutical materials used. containers are tested by many methods of which commonly used test for glass are crushed glass test, whole container test, chemical resistance of test, water attack test etc. closure materials are tested by transparency test penetrability fragmentation test self seal ability test, extractive test etc. the requirement of packaging material testing is set according to specification of regulatory agencies like WHO, GMP, USFDA and ICH guidelines.

KEYWORDS: container, packaging, stability, carton etc.

DEFINITION

Packaging of drug or dosage form is an integral part of any pharmaceutical industry. Packaging affects the quality, stability and identification of drug product. Packaging provide an adequate degree of protection, minimize the loss of constituents and should not interact physically or chemically with the contents in a way that will alter their quality to an extent beyond the limits given in the individual monograph, or present a risk of toxicity. The package must ensure adequate stability of the product throughout the shelf life.

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Types of packaging

There are three types of packaging named primary, secondary and tertiary packaging used in the packaging technology.

Primary packaging

It is the material that first envelops the product and holds it. This usually is the smallest unit of distribution or use and is the package which is in direct contact with the contents.

Secondary packaging

It is outside the primary packaging which used to group primary packages together.

Tertiary packaging

It is used for bulk handling, warehouse storage and transport shipping. The most common form is a palletized unit load that packs tightly into containers.

Objectives of Packaging

Physical Protection: The dosage forms enclosed in the package may require protection from many things like light, microorganisms, environmental impurities, mechanical shock, electrostatic discharge, compression, temperature and vibration etc.

Barrier Protection: A barrier from oxygen, water vapor, dust etc., is often required. Permeation is a critical factor in design. Some packages contain desiccants or oxygen absorbers to help extend shelf life. Keeping the contents clean, fresh, sterile, and safe for the intended shelf life is a primary function.

Containment or agglomeration: Small objects are typically grouped together in one package for reasons of efficiency. For example, a single strip contain 10 to 20 tablets requires less physical handling than 10 to 20 single tablets. Liquids, powders, and granular materials need containment.

Information transmission: Packages and labels communicate how to use, store, list of ingredients used in the dosage forms, manufacturing & expiry dates, transport, recycle or dispose of the package or product. With pharmaceuticals, food, medical and chemical products, some types of information are required by governments. Some packages and labels also are used for tract and trace purposes.

Marketing: The packaging and labels can be used by marketers to encourage potential buyers to purchase the product. Package graphic design and physical design have been important and constantly evolving phenomenon for several decades. Marketing communications and graphic design are applied to the surface of the package and the point of sale display.

Security: Packaging can play an important role in reducing the security risks of shipment.

Convenience: Packages can have features that add convenience in distribution, handling, stacking, display, sale, opening, reclosing, use, dispensing, reuse, recycling, and ease of disposal.

Portion control: Single dosage packaging has a precise amount of contents to control usage. Bulk commodities can be divided into packages that are a more suitable size for individual households. It is also aids the control of inventory: selling sealed one liter bottles of milk, rather than having people bring their own bottles to fill themselves.

Storage function: The materials used for packaging should be stored properly so as to preserve the quality of the material both before packaging and once the package contents have been used.

Properties of packaging materials

Mechanical properties

- The materials used should possess sufficient mechanical strength to withstand while handling, filling, closing and processing. Typical care is needed during transport, storage and also at the time of usage by the consumer especially in case of glass containers.

Physical properties

- The packaging must have a suitable size, thus, rubber may prevents problems if it perishes.

- The material must protect from light if necessary, that is, it must be ultraviolet absorbent.
- The container must not absorb substances from the products; e.g. absorption of water from creams in to cardboard box.

Chemical properties

- The product should not react with the container or closure, as might happen if alkaline substances are placed in aluminium containers.
- The container or closure must not yield substances to the product; for example, alkali from glass, plasticizers from plastics etc.

Biological properties

- The material of the container must be able to withstand attack by insects if this hazard is like to be encountered. The packing should not support mould growth.

Ideal requirements of good package

1. They should be able to hold the product without loss on account of leakage, spoilage or permeation.
2. They should afford protect against environmental conditions like light, air and moisture during storage.
3. They should not have any permeability for gases.
4. They should possess sufficient strength to withstand shocks of handling, transportation etc.
5. They should facilitate efficient safe and convenient use of contents.
6. The material must not interact with the content.
7. The containers should afford protection from moulds, bacterial etc.
8. The cost of material should be as low as possible without compromising the quality.
9. They should facilitate easy identification.
10. They should afford protection from moulds, bacteria etc.
11. The container should not absorb or adsorb any material containing.
12. The closure should provide air tight closing to the container.
13. The closure should be compatible with the preparation.

Table 3: Hazards encountered by the package

S. No.	Hazards	Details
1	Shock or impact damage	Damage due to shock is usually caused by rough handling, during transport etc. cushioning can be provided and a warning label may be useful. Restriction of movement and more careful handling should be made
2	Compression	Fragile items may be broken or collapsible articles crushed by compression, the usual procedure then being to protect with a rigid outer package. Top pressure or loading can distort inside.
3	Vibration	Vibration consists of two variables frequency and amplitude. Considerable vibration may occur during transport, especially with exported items. Damage may be external, such as the 'scuffing' of labels, but some products may be affected like the cracking of emulsions, abrasion of tablets, or segregation of mixed powders.
4	Environmental hazards	Environmental conditions encountered by the package are likely to vary considerably, especially in articles for export to the tropical areas.
5	Temperature	Extreme conditions may cause deterioration, low temperatures leading to aqueous solutions freezing and hence to fracture of

		containers.
6	Moisture	Moisture as liquid or water vapour may cause physical changes (e.g. color fading, softening, hardening etc.) or chemical changes (hydrolysis, oxidation, effervescence etc.).
7	Pressure	Decrease in pressure, as in mountainous regions or during flight in non-pressurized transport aircraft, may cause thin containers to burst or strip packs to inflate.
8	Biological hazards	The packaging materials must be reasonably clean initially and when put together to form a finished package and restrict any further contamination as much as possible. In the case of sterile products the package and its closure must maintain a 100% effective seal against microbiological contaminants like bacteria, moulds and yeasts.
9	Chemical Hazards	The main risk of chemical hazard is due to interaction or in compatibility between the product and package. These may be associated with interaction or contamination, covering migration, absorption, adsorption, extraction and corrosion etc.

Selection of suitable package

Different types of packages are used for packaging. A **transport package** or **distribution package** can be the shipping container used to ship, store, and handle the product or inner packages. Some identify a **consumer package** as one which is directed toward a consumer or household. Packaging may be described in relation to the type of product being packaged like medical device packaging, bulk chemical packaging, over the counter drug packaging, retail food packaging, military material packaging, pharmaceutical packaging etc. It is sometimes convenient to categorize packages by layer or function: "primary", "secondary", etc.

- Primary packaging is the material that first envelops the product and holds it. This usually is the smallest unit of distribution or use and is the package which is in direct contact with the contents.
- Secondary packaging is outside the primary packaging, perhaps used to group primary packages together.
- Tertiary packaging is used for bulk handling, warehouse storage and transport shipping.
- These broad categories can be somewhat arbitrary. For example, depending on the use, a shrink wrap can be primary packaging when applied directly to the product, secondary packaging when combining smaller packages, and tertiary packaging on some distribution packs.

Different types of closures

Closures may be defined as the devices and techniques used to close or seal a bottle, jug, jar, tube, can etc. closures can be a cap, cover lid. Plug etc. various types of containers such as boxes and drums may also have closures.

Purposes of closures

Many containers and packages require a means of closing. It can be a separate device or seal or sometimes an integral latch or lock. Depending on the contents and container, closures have several functions:

- Keep the container closed and the contents contained for the specified shelf life until the time of opening of container
- Provide a barrier to dirt, oxygen, moisture, etc. Control of permeation is critical to many types of products: foods, chemicals, etc.
- Keep the product secure from undesired premature opening

- Provide a means of reclosing or reusing the container
- Assist in dispensing and use of product
- Allow reasonable ease to open the container by the intended user. Difficult to open containers may cause wrap rage. The force or torque required to open a closure is an important consideration for packaging engineers.

Many types of packaging with their closures are regulated for strength, safety, security, communication, recycling, and environmental requirements.

Types of closures

Closures need a means of attaching to the container with sufficient security. Threads, lugs, hinges, locks, adhesives, etc. are used.

Many closures need to have the ability to adjust to slight manufacturing variation in the container and the closure structure. Some closures are made of flexible material such as cork, rubber or plastic foam. Often an o-ring or a closure liner (gasket made of pulp or foam cap liner) is used. Lineless closures often use a deformable plastic rim or structure to maintain the seal.

Secondary seals are common with sensitive products that may deteriorate or where extra security is needed. Foil or plastic inner seals are used on some bottles, Heat sealed lid ding films are used on some tubs. External shrink bands, labels, and tapes are sometimes used outside the primary closure structure.

Child resistant packaging

By law, manufacturers must package certain household items-including some medicines, cleaning products and gardening goods- in child resistant packaging.

The world health organization and UNICEF state that child resistant packaging is one of the best documented successes in preventing accidental poisoning of children.

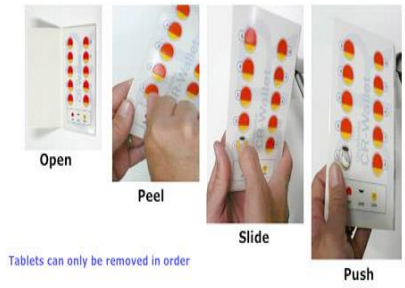
Child-resistant packaging or **C-R packaging** is special packaging used to reduce the risk of children ingesting dangerous items. This is often accomplished by the use of a special **safety cap**. It is required by regulation for prescription drugs, over-the-counter medications, pesticides, and household chemicals. In some jurisdictions, *unit packaging* such as blister packs is also regulated for child safety. The child-resistant locking closure for containers was invented in 1967 by Dr. Henri Breault.

A child resistant package usually requires a special trick to open it – something too complicated for most young children to work out. For example, users might have to push or squeeze a lid at the same time as turning it. It's also possible to make non-reclosable packs, such as blister packs, child resistant by using very strong material or covers that have to be peeled off.

Child resistant packaging standards: the details

The standards state that child resistant packaging should be tested with children and adults as follows:

- A group of children aged between 42 and 51 months are asked to open a pack. If they don't succeed after five minutes, they are shown how to open it, and then given five more minutes to try again
- A child resistant pack should be impossible to open for at least 85 per cent of children in the first five minutes, and for at least 80 per cent following the silent demonstration
- The pack is also tested with a panel of adults aged between 50 and 70. At least 90 per cent of this group must be able to open and reclose the pack or – for a non-reclosable pack – open it and remove one item. The test uses older adults as they are most likely to have difficulty opening and reclosing child resistant containers
- Group sizes vary because a sequential testing method is used. The child test, for example, usually involves between 30 and 60 children, but testers might need to use as many as 200 children to get a clear picture of a pack's child resistance




Tamper proof packaging

Tamper resistance is resistance to tampering (intentional malfunction or sabotage) by either the normal users of a product, package, or system or others with physical access to it. There are many reasons for employing tamper resistance.

Tamper resistance ranges from simple features like screws with special heads, more complex devices that render themselves inoperable or encrypt all data transmissions between individual chips, or use of materials needing special tools and knowledge. Tamper-resistant devices or features are common on packages to deter package or product tampering.

In some applications, devices are only tamper-evident rather than tamper-resistant.

Tampering

Tampering involves the deliberate altering or adulteration of a product, package, or system. Solutions may involve all phases of product production, packaging, distribution, logistics, sale, and use. No single solution can be considered as "tamper-proof". Often multiple levels of security need to be addressed to reduce the risk of tampering. Some considerations might include:

- Identify who a potential tamperer might be: average user, child, psychopath, misguided joker, saboteur, organized criminals, terrorists. What level of knowledge, materials, tools, etc. might they have?
- Identify all feasible methods of unauthorized access into a product, package, or system. In addition to the primary means of entry, also consider secondary or "back door" methods.
- Control or limit access to products or systems of interest.
- Improve the tamper resistance to make tampering more difficult, time-consuming, etc.
- Add tamper-evident features to help indicate the existence of tampering.
- Educate people to watch for evidence of tampering.
- Tamper means interfere with (something) without authority or so as to cause damage.


Cushioning design

Knowledge of the pharmaceutical and food product to be packaged is critical. Manufacturers sometimes spend for laboratory analysis that can help to quantify the fragility of their product. At times a product can be made more rugged or can be supported to make it less susceptible to breakage. This support can be made through **cushion packaging** for a cushion must deform under shock for it to function. Package cushion is used to help protect fragile items during shipment. It is common for a transport package to be dropped, kicked, and

impacted. Delicate products cannot escape from transportation vibration especially from conveyors, trucks, railroads, or aircraft which can also damage some items. These unfortunate events may produce potentially damaging shocks and vibrations but all these can be controlled by cushioning so that the chance of product damage is greatly reduced. Moreover, internal packaging materials are also used for functions other than cushioning. Some are used just to immobilize the products in the box and to block them in place. Others are just used to fill a void and do not have a cushioning function.

Additionally, the amount of shock transmitted by a particular cushioning material is largely dependent on the thickness of the cushion, the drop height, and the load-bearing area of the cushion. If a product is on a large load-bearing area, the cushion may not deform and will not cushion the shock. In addition, engineering judgment can also be an excellent starting point for understanding the science of cushion. Engineers use cushion curves to choose the best thickness and load-bearing area for a cushioning material. Often two to three inches of cushioning are needed to protect fragile items. On other hand, delicate products also need protection from vibration. The process for vibration protection involves similar considerations as that for shock. **Cushion packaging** can be thought of as performing like springs. Depending on cushion thickness and load-bearing area and on the forcing vibration frequency, the cushion may not have any influence on input vibration, amplify the input vibration or isolate the product from the vibration. With this, proper design is critical for cushion performance.

When designing packaging, the preference may depend on factors such as effective protection of product from shock and vibration, creep resistance, humidity and air pressure on cushioning, cleanliness of cushioning, effect on size of external shipping container, material and labor cost, whether cushioning is resilient and sensitivity of product to static electricity. Also, types of cushioning must be considered for this will guide the manufacturers on what is the best for their product. Some types include loose fill, paper, corrugated fiberboard pads, foam structures, molded pulp, inflated products and environmentally friendly.

Hence, proper performance of cushioning is dependent on its proper design and use. It is often best to invest with a well trained packaging consultant or an engineer to validate and verify the prototype designs. The design of a package and its cushioning is a process involving several designs, evaluation and redesigns. Consequently, the main goal of **cushion packaging** is to transport products safely and securely by minimizing the shipping damages and breakages cause by shock, vibration and sometimes from scratching by utilizing the correct form of packaging design of pharmaceutical and food products. Therefore, customer satisfaction will be realized for product standard and services are scientifically emphasized.

Design factors

When designing packaging, the choice of cushioning may depend on many factors:

- effective protection of product from shock and vibration
- cushioning is resilient (performs for multiple impacts)
- resistance to creep – cushion deformation under static load
- material costs
- labor costs, productivity
- effects of temperature, humidity, and air pressure on cushioning
- Cleanliness of cushioning (dust, insects, etc.)
- effect on size of external shipping container
- environmental and recycling issues
- Sensitivity of product to static electricity.

Common types of cushioning

Loose Fill - Some cushion products are flowable and are packed loosely around the items in the boxes. The box is closed to tighten the pack. This includes expanded polystyrene foam

pieces (Foam peanuts), similar pieces made of starch-based foams, and common popcorn. The amount of loose fill material required and the transmitted shock levels vary with the specific type of material.

Paper - Paper can be manually or mechanically wadded up and used as a cushioning material. Heavier grades of paper provide more weight-bearing ability than old newspapers. Creped cellulose wadding is also available. (Movers often wrap objects with several layers of Kraft paper or embossed pulp before putting them into boxes.)

Corrugated fiberboard **pads** - Multi-layer or cut-and-folded shapes of corrugated board can be used as cushions. These structures are designed to crush and deform under shock stress and provide some degree of cushioning. Paper board composite honey comb structures are also used for cushioning.

Foam **structures** - Several types of polymeric foams are used for cushioning. The most common are: Expanded Polystyrene (also Styrofoam), polypropylene, polyethylene, and polyurethane.

These can be molded engineered shapes or sheets which are cut and glued into cushion structures. Some degradable foam is also available. Foam-in-place is another method of using polyurethane foams.


REFERENCES

1. Garrett, E.R.: J. Pharm. Sci., 51:35, 1962.
2. Levy, G.: J. Pharm, Sci., 50:429, 1961.
3. Dimbley, V.: J. Pharm. And Pharmacol., 5.:969, 1953.
4. Boddapati, S., and Butler, D. L., Im,S,and DeLuca, P.P.: J.Pharm. Sci., 69:608, 1980.
5. Sarrut, S., and Nezelof, C .: Presse Med., 68:375, 1960.
6. Vanga, S. V.: J. Parent. Drug. Assoc., 33:61, 1979.
7. Moretti, C.: Boll Chem. Farm. 103:69, 1964.
8. Lipper, R.A., and Nevola, M.M..: Influence of Amber Glass on the Decomposition of Thiomerosol in Aqueous solution. In press.
9. Enever, R.P., and LiWanPo, A., and Schotton, E.: J.Pharm.Sci., 66:1087,1977.
10. Kassem, M.A.,Kassem, A.,A., and Ammar , H.O.: Pharm.Acta Helv., 44:611, 1969.
11. European Pharmacopoeia 5.0, 3.2.1. Glass containers for pharmaceutical use 01,6th edition
12. United State Pharmacopoeia XIX. Mack Publishing, Eastron, PA 1975
13. D. E. McDonald, Am. J. Hosp. Pharm. 29: 223-228 (1972).
14. J. W. DIETRICH, Am. J. Hosp. Pharm. 30: 805-808 (1973).
15. S. TURCO and R. E. KING, *Sterile Dosage Forms, Their Preparation and Clinical Application*, Lea & Febiger, Philadelphia, 1974.
16. P. MOORHATCH and W. L. CHIOU, Am. J. Hosp. Pharm. 31: 72-78 (1974).
17. P. MOORHATCH and W. L. CHIOU, Am. J. Hosp. Pharm. 31: 149-152 (1974).
18. J. AUTIAN and C. N. DHORDA, Am. J. Hosp. Pharm. 16: 176-179 (1959).
19. L. D. BIGHLEY, J. WILLE and J. L. LACK, Am. J. Hosp. Pharm. 31: 736-739 (1974).
20. W. L. CHIOU and P. MOORHATCH, J. Am. Med. Assoc. 224: 1298-1299 (1973).
21. E. MARCUS, H. K. KIM and J. AUTIAN, J. Pharm. Sci. 48: 457-462 (1959).

ANTIBIOTIC RESISTANCE

Rajiv Kumar Arora, Monika Kaushik, Balvinder Singh

Faculty of Pharm. Sciences, Baba Mastnath University, Asthal Bohar, Rohtak

INTRODUCTION

An increase in the frequency of antibiotic resistance in bacteria since the 1950s has been observed for all major classes of antibiotics used to treat a wide variety of respiratory illnesses, skin disorders, and sexually transmitted diseases. Is this resistance the result of bacteria evolving new genes in response to the presence of antibiotics, or are antibiotic-resistant bacteria selected for in the environment by possessing antibiotic resistance genes beforehand? To answer these questions a discussion of several factors involved in antibiotic resistance will show that resistance is a designed feature of pre-existing genes enabling bacteria to compete with the antibiotic producers in their environment.

A brief look at an example of penicillin resistance reveals the increase in the frequency of antibiotic-resistant organisms since the time when antibiotic use became common. Penicillin is an antibiotic produced by the common bread mold Penicillium that was discovered accidentally in 1929 by the British microbiologist, Alexander Fleming. By the 1940s, penicillin was available for medical use and was successfully used to treat infections in soldiers during World War II. Since then, penicillin has been commonly used to treat a wide range of infections. In 1967 the first penicillin-resistant Streptococcus pneumoniae was observed in Australia, and seven years later in the U.S. another case of penicillin-resistant S. pneumoniae was observed in a patient with pneumococcal meningitis. In 1980 it was estimated that 3-5% of S. pneumoniae were penicillin-resistant and by 1998, 34% of the S. pneumoniae sampled were resistant to penicillin.1 Antibiotic resistance by other organisms reflects the same trend observed between S. pneumoniae and penicillin. Tetracycline resistance by normal human intestinal flora has exploded from 2% in the 1950s to 80% in the 1990s. Kanamycin, an antibiotic used in the 1950s, has become clinically useless as a result of the prevalence of kanamycin-resistant bacteria. The increase in resistance among these organisms clearly indicates a change in the frequency of antibiotic resistance genes.

Since World War II many more antibiotics isolated from fungi (molds) and bacteria have been used to treat a wide range of human and animal infections. One group of bacteria, the Streptomyces, produces most of the medically important antibiotics. Streptomyces release antibiotics into the soil in a sort of "biochemical warfare" scenario to eliminate competing organisms from their environment. These antibiotics are small molecules that attack different parts of an organism's cellular machinery. Streptomyces-produced quinolone and coumarin antibiotics, such as novobiocin, interfere with a protein called gyrase that assists in the normal separation of double-stranded DNA during replication of DNA or transcription of messenger RNA. Failure of DNA to properly separate during these processes results in a bacterium not being able to divide normally or produce functional proteins. Ribosomes, the structures where protein synthesis is catalyzed, are the targets of many other Streptomyces antibiotics such as spectinomycin, tetracycline, and streptomycin. Spectinomycin and tetracycline prevent proteins from being assembled by the cell and streptomycin induces the assembly of the wrong amino acids into the translated protein. Without proteins, which are necessary for normal cell function, the cell dies. The slight differences between human ribosomes which are not bound by these antibiotics and bacterial ribosomes make this type of antibiotic ideal for treating many illnesses. Other antibiotics, such as penicillin, block the assembly of the bacterial cell wall causing it to weaken and burst. Penicillin is an effective antibiotic for human diseases because it interferes with a biological component in bacteria

(cell wall) not found in human cells. The production of antibiotics by these organisms provides them with a competitive advantage over non-resistant bacteria in their environment. Just as large organisms such as plants and animals must compete for living space, food, and water, these microbes use antibiotics to eliminate competition with other microbes for these same resources.

However, not all bacteria are defenseless against the antibiotic producers. Many possess genes that encode proteins to neutralize the affects of antibiotics and prevent attacks on their cell machinery. Efflux pumps, located in the cell membrane, are one method of protection that many bacteria use against the influx of antibiotics. The offensive antibiotic is pumped out of a cell that possesses these pumps before the antibiotic can cause harm to the cellular machinery. Although many efflux pumps may be specific for the substrate they pump out of the cell, they are not uncommon.

Ribosomal protection proteins (RPP)

Ribosomal protection proteins (RPP) are another source of resistance bacteria use to protect themselves from antibiotics. These proteins protect ribosomes by binding them and changing their shape or conformation. The change in the ribosome shape prevents an antibiotic from binding and interfering with protein synthesis. The RPP-bound ribosomes are able to function normally during protein synthesis, an important feature of this method of antibiotic resistance. Some bacteria produce enzymes that neutralize antibiotics by adding acetyl (COCH3) or phosphate (PO32-) groups to a specific site on the antibiotic. This modification reduces the ability of the antibiotic to bind to ribosomes, rendering it harmless to the cell. Interestingly, all three types of antibiotic-resistant genes that produce efflux pumps, ribosomal protection proteins, and modifying enzymes are found in Streptomyces species, the producers of many antibiotics. It appears this is the method Streptomyces uses to protect itself from its own antibiotics.

Transfer of resistance genes to other bacteria

A unique bacterial characteristic that has not been demonstrated in plant and animal cells is the ability to transfer genes from one bacterium to another, a process called lateral gene transfer. Genes located on a circular strand of DNA called an R-plasmid may contain several antibiotic-resistant genes. Through a process called conjugation an antibiotic-resistant bacterium can transfer the antibiotic resistance genes from an R-plasmid to a non-resistant bacterium. Ironically, several antibiotic resistance genes found in other pathogenic bacteria are very similar in DNA sequence to the genes found in Streptomyces species. The efflux pumps that Streptomyces use to pump out antibiotics to eliminate their competitors are likely the same pumps that other species of bacteria are now using to pump out the offensive antibiotic delivered from Streptomyces. The antibiotic-resistant bacteria likely have acquired the genes for these efflux pumps through lateral gene transfer. The presence of ribosomal protection proteins and antibiotic modifying enzymes in resistant bacteria has also likely originated from Streptomyces or some other antibiotic-producing microbe. Bacteria don't appear to be evolving new genes; they are acquiring previously existing antibiotic resistance genes through lateral gene transfer. This allows a species of bacteria to possess enough genetic variability to adapt to a changing environment and to compete with its neighbors. (This method of defense is very similar to the genetic variability of mammalian antibody-producing B lymphocytes—a topic for another Impact article.) The bacterium that acquires the antibiotic resistance genes still has the physical and metabolic qualities that distinguish it from other bacteria kinds and associates it with its own kind of bacteria. The observed increase in the frequency of antibiotic-resistant bacteria has resulted from the increased use of antibiotics in medicine and agriculture, resulting in the reduction of organisms that do not possess antibiotic resistance genes.

Through Mutation

Antibiotic resistance in bacteria can also be achieved when mutations in a ribosome or protein change the site where an antibiotic binds. For example, four of the antibiotics mentioned earlier, tetracycline, streptomycin, kanamycin, and spectinomycin, bind to a specific region of a ribosome and interfere with protein synthesis. Mutations may prevent an antibiotic from binding to the ribosome (kanamycin) or allow the ribosome to function even while the antibiotic is bound (streptomycin and spectinomycin). Although it appears these mutations are beneficial and provide an advantage to the bacterium possessing them, they all come with a cost. Ribosomal mutations, while providing antibiotic resistance for the organism, slow the process of protein synthesis, slow growth rates, and reduce the ability of the affected bacterium to compete in an environment devoid of a specific antibiotic. Furthermore, a mutation that confers resistance to one antibiotic may make the bacterium more susceptible to other antibiotics. These deleterious effects are what would be expected from a creationist model for mutations. The mutation may confer a benefit in a particular environment, but the overall fitness of the population of one kind of bacterium is decreased as a result of a reduced function of one of the components in its biological pathway. The accumulation of mutations doesn't lead to a new kind of bacterium—it leads to extinction.

RESEALED ERYTHROCYTES AS A NOVEL DRUG TARGETING CARRIERS. A REVIEW

Rishi Pal,

SBMNIPS&R , BMU, Asthal Bohar , Rohtak, Haryana (INDIA) Email I.D:- rishi2105@gmail.com

INTRODUCTION

Blood consists of different types of cells like RBC(erythrocytes), WBC (leucocytes), and platelets. The most abundant cells in the human body are erythrocytes having potential carrier capabilities to deliver a number of drugs. They are biodegradable, biocompatible, having circulating capability with a long life span, with a variety of biologically active compounds using various chemical and physical methods. This drug delivery system can be explored to maximize therapeutic performance, reducing undesirable side effects of drug as well as increase patient compliance. At present the development of drug delivery, to maximize the drug targeting with maximum therapeutic drug efficacy to specific organ or tissue with minimum side effects.

Such drug-loaded carrier erythrocytes are prepared simply by collecting blood samples from the organism of interest, separating erythrocytes from plasma, entrapping drug in the erythrocytes, and resealing the resultant cellular carriers, so called resealed erythrocytes. RBCs having life span of 120 days, deliver O_2 to the body, each circulate takes 20 sec. so max .efficacy can be attained in delivery of a drug to a specific site.

Resealed Erythrocytes:

The drug-loaded carrier erythrocytes are prepared simply by collecting blood samples from the organism of interest, separating erythrocytes from plasma, entrapping drug in the erythrocytes, and resealing the resultant cellular carriers(1). Hence, these carriers are called resealed erythrocytes. The overall process is based on the response of these cells under osmotic conditions. Upon reinjection, the drug-loaded erythrocytes serve as slow circulating depots and target the drug.. . Jain and Vyas have described a well-established protocol for the isolation of erythrocytes(2). In 1953, Gardos tried to load erythrocyte ghost using adenosine triphosphate (ATP)(3). In 1959, Marsden and Ostting reported the entrapment of dextran (molecular weight 10–250 kDa). In 1973, the loading of drugs in erythrocytes was reported separately by Ihler et al.(4) and Zimmermann(5). In 1979, the term carrier erythrocytes were coined to describe drug-loaded erythrocytes (6).

Erythrocytes can be used as carriers in two ways:-

1. Targeting particular tissue/organ.: As the drug-loaded erythrocytes serve as slow circulating depots and target the drugs to disease tissue or organ.(7)- For targeting, only the erythrocyte membrane is used. This is obtained by splitting the cell in hypotonic solution and after introducing the drug into the cells, allowing them to reseal into spheres. Such erythrocytes are called Red cell ghosts.

2. For continuous or prolonged release of drugsAlternatively, erythrocytes can be used as a continuous or prolonged release system, which provide prolonged drug action. There are different methods for encapsulation of drugs within erythrocytes. They remain in the circulation for prolonged periods of time (up to 120 days) and release the entrapped drug at a slow and steady rate.

Advantages of erythrocytes as drug carriers

- Biocompatible, particularly when autologous cells are used hence no possibility of triggered immune response.
- Biodegradability with no generation of toxic products.
- Considerable uniform size and shape of carrier.

- Relatively inert intracellular environment can be encapsulated in a small volume of cells.
- Isolation is easy and large amount of drug can be loaded.
- Prevention of degradation of the loaded drug from inactivation by endogenous chemical.
- Entrapment of wide variety of chemicals can be possible.
- Entrapment of drug can be possible without chemical modification of the substance to be entrapped.
- Possible to maintain steady-state plasma conc., decrease fluctuation in conc..
- Protection of the organism against toxic effect of drug.
- Targeting to the organ of the RES.
- Ideal zero-order drug release kinetic.
- Prolong the systemic activity of drug by residing for a longer time in the body.

Advantages of erythrocytes as drug carriers

- They have a limited potential as carrier to non-phagocyte target tissue.
- Possibility of clumping of cells and dose dumping may be there.(8,9)

Methods of drug loading

Several methods can be used to load drugs or other bioactive compounds in erythrocytes, including physical (e.g., electricalpulse method) osmosis-based systems, and chemical methods (e.g., chemical perturbation of the erythrocytes membrane).

Irrespective of the method used, the optimal characteristics for the successful entrapment of the compound requires the drug to have a considerable degree of water solubility, resistance against degradation within erythrocytes, lack of physical or chemical interaction with erythrocyte membrane, and well-defined pharmacokinetic and pharmacodynamic properties (10).

1. Hypotonic hemolysis.
2. .Use of red cell loader.
3. Hypotonic dilution.
4. Hypotonic preswelling:
5. Hypotonic dialysis..
6. Isotonic osmotic lysis.
7. Chemical perturbation of the membrane.
8. Electro-insertion or electroencapsulation..
9. Entrapment by endocytosis
10. Loading by electric cell fusion.
11. Loading by lipid fusion.

***In-vitro* characterization**

After loading of therapeutic agent on erythrocytes, the carrier cells are exposed to physical, cellular as well as biological evaluations. .*In-vitro* characterization forms an important part of studies involving such cellular carriers. The given table summarizes the various evaluation parameters and the techniques applied for their determination. The morphology of erythrocytes decides their life span after administration.

Sr.No.	Parameters	Method/instrument used
1	**Physical characterization**	
	Vesicle size and size distribution	Transmission electron microscopy (TEM), optical microscopy(OM).
	Shape and surface morphology	TEM, scanning electron microscopy (SEM), phase contrast microscopy, optical microscopy.
	Drug release	Diffusion cell , dialysis

	Drug contants	Deproteinization of cell membrane followed by assay of resealed drug, radiolabelling
	Surface pH	Zeta potential measurement
	Surface electric potential.	pH-sensitive probes
2	**Cellular characterization**	
	% Hb content	Deproteinization of cell membrane followed by hemoglobin assay
	Cell volume	Laser light scattering
	Cell recovery	Neubaur's chamber, hematological analyzer
	Osmotic fraglity	Stepwise incubation with isotonic to hypotonic saline solutions &determination of drug and hemoglobin assay
	Osmotic shock	Dilution with distilled water and estimation of drug and hemoglobin
	Turbulent shock	Passage of cell suspension through 30- gauge hypodermic needle at 10 mL/min flow rate and estimation of residual drug and hemoglobin, vigorous shaking followed by hemoglobin estimation
	Erythrocyte sedimentation rate	ESR methods
3	**Biological characterization**	
	Sterility	Sterility test
	Pyrogenicity	Rabbit method, LAL Test
	Animal toxicity	Toxicity tests

Application of resealed erythrocytes:

- Slow drug release.
- Drug targeting
- Targeting RES organ
- Targeting of liver
    i. Enzyme deficiency/replacement therapy
    ii. Treatment of hepatic tumors
    iii Treatment of parasitic diseases.
    iv. Removal of RES iron overload
    v. Removal of toxic agents
- Targeting organs other than those of RES.
- Delivery of antiviral agents.
- Enzyme therapy
- Improvement in oxygen delivery to tissues.

Novel approaches

Erythrosomes:.These are specially engineered vesicular systems that are chemically cross-linked to human erythrocytes' support upon which a lipid bilayer is coated. This process is achieved by modifying a reverse-phase evaporation technique. These vesicles have been proposed as useful encapsulation systems for macromolecular drugs (11,12,13).

Nanoerythrosomes.:These are prepared by extrusion of erythrocyte ghosts to produce small vesicles with an average diameter of 100 nm. Daunorubicin was covalently conjugated to nanoerythrosomes using gluteraldehyde spacer. This complex was more active than free daunorubicin alone, both in vitro and in vivo (14,15).

CONCLUSION:

The use of resealed erythrocytes looks promising for a safe and sure delivery of various drugs for passive and active targeting. However, the concept needs further optimization to become a routine drug delivery system. The same concept also can be extended to the delivery of biopharmaceuticals and much remains to be explored regarding the potential of resealed erythrocytes. It is very effective and safe delivery system for anti cancer drug with or without less toxicity.

REFERENCES:

1. Ropars C., Chassaigne M., and Nicoulau C. Advances in the Biosciences, Pergamon Press, Oxford, 1987: 67
2. S.P. Vyas and R.K. Khar, Resealed Erythrocytes in Targeted and Controlled Drug Delivery: Novel Carrier Systems (CBS Publishers and Distributors, India, 2002), pp. 87–416.
3. G. Gardos,"Akkumulation de Kalium Onen Durch Menschiche Blutkorperchen," Acta Physiol. Acad. Sci. Hung 6, 191–196 (1953).
4. G.M. Iher, R.M. Glew, and F.W. Schnure, "Enzyme Loading of Erythrocytes," Proc. Natl. Acad. Sci. USA 70, 2663–2666 (1973).
5. U. Zimmermann, Jahresbericht der Kernforschungsanlage Julich GmbH (Nuclear Research Center, Julich, 1973), pp. 55–58.
6. S. Jain and N.K. Jain, "Engineered Erythrocytes as a Drug Delivery System," Indian J. Pharm. Sci. 59, 275–281 (1997).
7. G.M. Ihler,"Erythrocyte Carriers,"Pharmacol. Ther.(1989).20, 151–169.
8. Carmen Gutierrez Millan, Maria Luisa Sayalero Marinero, Aranzazu Zarzuelo Castaneda and Jose M. Lanao. Drug, enzyme and peptide delivery using erythrocytes as carriers, Journal of Controlled Release, 2004; 95(1): 27-49.
9. Mehrdad Hamidi, Adbolhossein Zarrina, Mahshid Foroozesha and Soliman Mohammadi-Samania. Applications of carrier erythrocytes in delivery of biopharmaceuticals, Journal of Controlled Release, 2007; 118(2):145-160.
10. M. Hamidi and H. Tajerzadeh, "Carrier Erythrocytes: An Overview," Drug Delivery 10, 9–20 (2003).
11. .J. Cuppoletti et al., "Erythrosomes: Large Proteoliposomes Derived from Cross-Linked Human Erythrocyte Cytoskeletons and Exogenous Lipid," Proc. Natl. Acad. Sci. USA 78 (5), 2786–2790 (1981).
12. .C.Y. Jung, Methods in Enzymology (Academic Press, New York, NY, 1987), pp. 149–217.
13. S.P. Vyas and V.K. Dixit, Pharmaceutical Biotechnology 1 (CBS Publishers & Distributors, New Delhi, 1999), pp. 655.
14. M. Moorjani et al., "Nanoerythrosomes, A New Derivative of Erythrocyte Ghost II: Identification of the Mechanism of Action," Anticancer Res. 16 (5A), 2831–2836 (1996).
15. A. Lejeune et al., "Nanoerythrosomes, A New Derivative of Erythrocyte Ghost: III. Is Phagocytosis Involved in the Mechanism of Action?," Anticancer Res. 17, 5A (1997)

A REVIEW ON PREPARATION AND STABILITY OF EMULSION

Gaurav Khurana

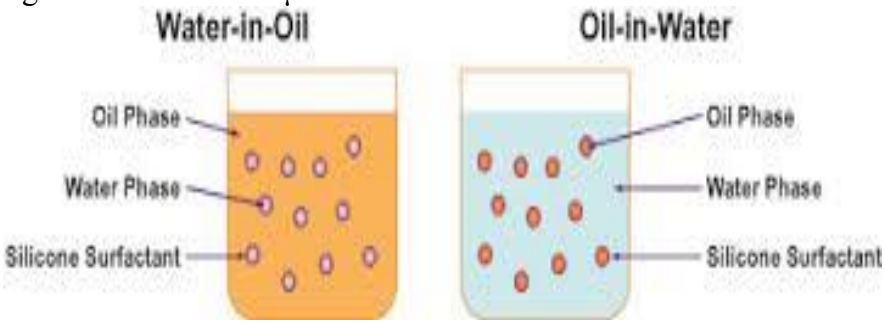
Department of Pharmaceutics SBMNIPSR, Rohtak**Email:** gaurav.khurana786@gmail.com

The pharmaceutical term "emulsion" is most time used to indicate preparations prepared for internal use. Emulsions for external use are always given a different title that it focus may indicate their use, e.g. lotion and cream

An emulsion may be defined as a biphasic system consisting of two immiscible liquids, one of which (the dispersed phase) is finely and uniformly dispersed as globules throughout the second phase (the continuous phase). Since emulsions are a thermodynamically unstable system, a third agent, the emulsifier is added to stabilize the system (Agarwal and Rajesh, 2007).

Emulsifier stabilizes the system by forming a thin film around the globules of dispersed phase (Javed et al., 2008). Either the dispersed phase or the continuous phase may vary in consistency from that of a mobile liquid to semisolid (Alfred, 2005). Thus, pharmaceutical emulsions range from lotions (low viscosity) to creams (high viscosity). The particle size of the dispersed phase

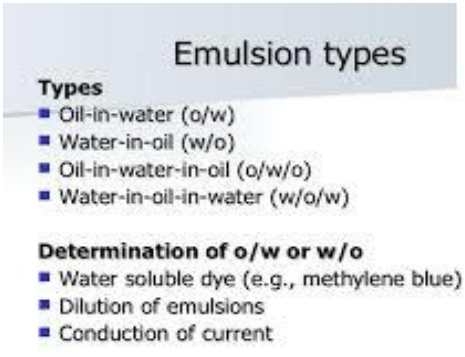
commonly ranges from 0.1 to 100 μm.



Types of emulsion

The word "emulsion" comes from the Latin word for "to milk", as milk is an emulsion of fat and water, among other components.

Two liquids can form different types of emulsions. As an example, oil and water can form, first, an oil-in-water emulsion, wherein the oil is the dispersed phase, and water is the dispersion medium. Second, they can form a water-in-oil emulsion, wherein water is the dispersed phase and oil is the external phase. Multiple emulsions are also possible, including a "water-in-oil-in-water" emulsion and an "oil-in-water-in-oil" emulsion.



Multiple emulsions

Multiple emulsions are complex systems, termed "emulsions of emulsions", i.e. the droplets of the dispersed phase contain even smaller dispersed droplets themselves. Each dispersed globule in the double emulsion forms a vesicular structure with single or multiple aqueous compartments separated from the aqueous phase by a layer of oil phase compartments.

Microemulsion

Microemulsion is defined as microemulsion are clear, transparent, thermodynamically stable dispersions of oil and water, stabilized by an interfacial film of surfactant frequently in combination with a co-surfactant.

Microemulsions are easily prepared and require no energy contribution during preparation this is due to better thermodynamic stability.

The dispersed phase, lipophilic or hydrophilic (O/W, or W/O microemulsions) can act as a potential reservoir of lipophilic or hydrophilic drugs, respectively.

Emulsifiers

An **emulsifier** is a substance that stabilizes an emulsion by increasing its kinetic stability. One class of emulsifiers is known as "surface active agents", or surfactants.

Examples of food emulsifiers are:

- Egg yolk – in which the main emulsifying agent is lecithin. In fact, *lecithos* is the Greek word for egg yolk.
- Mustard – where a variety of chemicals in the mucilage surrounding the seed hull act as emulsifiers
- Soy lecithin is another emulsifier and thickener
- Pickering stabilization – uses particles under certain circumstances

Water-in-water emulsion

Water-in-water emulsion is a system that consists of droplets of water-solvated molecules in another continuous aqueous solution; both the droplet and continuous phases contain different molecules that are entirely water-soluble. As such, when two entirely aqueous solutions containing different water-soluble molecules are mixed, water droplets containing predominantly one component are dispersed in water solution containing another component. Recently, such a water-in-water emulsion was demonstrated to exist and be stable from coalescence by the separation of different types of non-amphiphilic, but water-soluble molecular interactions. These molecular interactions include hydrogen bonding, pi stacking, and salt bridging. This w/w emulsion was generated when the different water-solvated molecular functional groups get segregated in an aqueous mixture consisting of polymer and liquid crystal molecules

PHARMACOVIGILANCE: A CONTEMPORARY VIEW

Jyoti dahiya[1], Priti Mehndiratta[1]

[1]Department of Pharmaceutical Sciences, Baba Mastnath University, Rohtak(124001), India. Email: jyotidahiya.2710@gmail.com, pritimehndiratta88@gmail.com

INTRODUCTION

Pharmacovigilance is the branch of science deals with the safety of medicines and error free medical services. According to WHO, duty of Pharmacovigilance includes detection, assessment, understanding and prevention of any obnoxious adverse drug reactions at therapeutic concentration on animal and human beings[1,2]. It plays a foremost role in ensuring that patient receive safe drugs. Further, pharmacovigilance may be considered as the phase IV(Post marketing surveillance) of clinical trials which involves the analysis and management of safety information from many sources. Pharmacovigilance programme identifies anticipated as well as unanticipated toxic effects of approved medicines, and helps to prevent further harm to the patients. In fact, when there is a safety data exchange agreement is there in between two companies, safety information is to be exchanged by either companies which is not possible without a well established pharmacovigilance system. Hence, call for a well established pharmacovigilance system is increasing day by day all over the world[3-6].

HISTORY

In 1961 initially an international efforts were made to consider drug safety issues after Thalidomide disaster. At that time number of phocomelic (congenitally deformed) babies were born in different countries to pregnant ladies who had consumed thalidomide, an approved medicine for morning sickness. These events gave birth to concept of PHarmacovigilance. Hence viewing the enormity of task the Central Drugs Standard Control Organization(CDSCO) has started a well structured National Pharmacovigilance Programme(NPP). There is need of a vivacious Pharmacovigilance system in india to protect the society from harms of new drugs approved for marketing in india. National Pharmacovigilance Programme of India was inaugurated by Health minister on 23rd November 2004 at New Delhi. All over the world including 80 countries these NPP are supported by an International Drug Monitoring Programmes established by WHO whose centre is located in Uppsala Sweden, hence also known as Uppsala Monitoring Centre[4,5,28].

OBJECTIVES OF PHARMACOVIGILANCE PROGRAMME

The particular aims and objectives of the Pharmacovigilance Programme are[6-9]:

- To minimize the stretch of unexpected adverse effects.
- To improve patients complete safety.
- To assess the benefits:risks ratio of medicines.
- To prevent prospective hazards by withdrawing unsafe medicines from market.
- Encouraging safe, rational and more effective use of drugs.
- To promote understanding, education and clinical training in Pharmacovigilance.
- Effectively communicate to the public.

SCOPE OF PHARMACOVIGILANCE

Within last few years, there has been an enormous growth about concept of Pharmacovigilance. That's why, recently the safety concerns have been extended to Herbal and natural products, ayurvedic medicines, blood products and cosmetics in addition to allopathic medicines. Advancements in communication technology all over the world have resulted in new safety concerns such as[10,11]:

- Illegal sales of medicine.
- Self medication practices.

- Sale of counterfeit medicines.
- Medical errors due to negligence of health professionals.

CURRENT APPROACHES IN PHARMACOVIGILANCE

- **Pharmacoenvironmentology:** Scientists and environmentologists are focusing these days on impact of drugs on environment and surroundings. Hence 'Pharmacoenvironmentology' deals with environmental crash of drugs given to humans and animals at therapeutic doses. Upcoming research in ecology enlighten the adverse effects of many drugs on environment. We are living in an environment polluted by heavy metals, pesticides, emissions from gasoline and also with pharmaceutical chemicals. In 1976 the first study was carried out at Big Blue River sewage treatment plant in Kansas that detected presence of drugs in sewage[12,13]. Number of regulatory bodies like FDA and European union set a cut-off limit for environmental concentrations of drugs, still there is no testing has been carried out after a drug is marketed regarding that cut-off limit. Hence this concept of Pharmacoenvironmentology must concern with drugs and their concentration in environment. Many ecofriendly techniques such as bioremediation and phytoremediation also help in minimizing concentration of environment pollutant in aquatic environment. Finally it is concluded that we need to monitor the effects of drugs not only as good medical practice, but also to safeguard our environment[14-16,28].
- **Pharmcovigilance in ayurvedic medicines:** Ayurveda describes diagnosis and therapy of disease as well as ways to maintain positive health. Although the term Pharmacovigilance does not fit in ayurvedic text. The objective behind pharmacovigilance is to improve patient care and safety to drug use, and similarly promote rational drug use are themes of ayurvedic pharmacology and therapeutics. The use of ayurvedic medicines is popular in india since years- and recently become accepted in other countries too. In accordance with this increasing use, safety concerns of ayurvedic medicines also lead forward. So, there is need for pharmacovigilance of ayurvedic medicines growing day by day[17-19].
- **Pharmacovigilance in herbal medicines:** Herbal medicines are broadly used these days in developed as well as developing countries. The widespread interest in phytotherapy over last decade has engrossed the attention of pharmaceutical companies and also the quality control, efficacy and safety of herbal drugs. Therefore, pharmacovigilance and regulations for herbal medicines are also important these days. Herbal pharmacovigilance should be implemented and authorities should record various aspects of herbal formulation and single herbs. Improvements in safety monitoring of herbal medicines include modifications in methodology already existed, patient reporting and pharmacogenetics consideration. Pharmacovigilance in herbal medicines is perhaps an unthought- of concept as yet; however we do not need an " Herbal Thalidomide" to wake the pharmacovigilance community to the need of the hour[20-23].
- **Cosmetovigilance:** Women, across the world use a gigantic range of cosmetics daily but this excessive use is leading to a chemical lay up in their bodies. Cosmetic preparation are the substances meant for application on external parts of body like hair system, epidermis, nails, lips, teeth and mucous membrane of oral cavity. These preparations are basically used to fulfill the purpose of cleaning, perfuming, changing the appearance and keeping them in good condition. Although cosmetic industries perform various analytical testing for these preparartions on their own to ascertain the safety, efficacy and tolerability of cosmetics but nevertheless cosmetics can induce adverse reactions. Main objectives of cosmetovigilance system should be standard

reporting of adverse reactions, routine quality controls, public health concern and regular evaluatin of ADR to cosmetics[24].

FUTURE PROSPECTS

On a regulatory level, progress has been made during the past few years. However, the results of these changes have yet to become apparent and, therefore, it has not yet been proven if these developments have contributed to better pharmacovigilance conduct[24,25]. In order to further prove

pharmacovigilance as a science, it is essential that academia develops new methods which can strengthen the current system. Active surveillance is necessary to receive information about the safety of a drug at an early stage. Pharmacogenetics could play a role in identifying individual risk factors for the occurrence of certain ADRs. The field of pharmacovigilance has made a remarkable journey since it was recognized in the early 1960s after the thalidomide disaster. Recent events, such as the withdrawal of aprotinin and the questioning of the safety of rosiglitazone, show that it is a topic that lies close to people's hearts[28].

CONCLUSION

Pharmacovigilance play an important role in meeting the challenges offered by the increased range and potency of medicines. When adverse effects and toxicity do appear especially when previously unknown it is essential that these are reported, analysed and their significance communicated effectively to an audience that has the knowledge to interpret the information. Moreover, providing the regulators with the necessary information to amend the recommendations on the use of the medicines; improving communication between the health professionals and the public; and educating the health professionals to understand the effectiveness and risk of medicines they prescribe, is the need of the moment.

REFERENCES

1. WHO Policy Perspectives on Medicines. Geneva: WHO; 2004. Geneva: World Health Organization. Looking at the Pharmacovigilance: ensuring the safe use of medicines.
2. Klepper MJ. The periodic safety update report as a pharmacovigilance tool. Drug Saf 2004;27:569-78.
3. Livio F, Renard D, Buclin T. Pharmacovigilance. Rev Med Suisse 2012;8:116-9
4. Pharmacovigilance. Mann RD, Andrews EB, eds. John Wiley & Sons Ltd, Chichester, 2002.
5. Harmark L, van Grootheest AC. Pharmacovigilance: Methods, recent developments and future perspectives. Eur J Clin Pharmacol 2008;64:743-52.
6. Biswas P, Biswas A. Setting standards for proactive pharmacovigilance in India: The way forward. Ind J Pharmacol 2007;39:124-8.
7. Arnott J, Hesselgreaves H, Nunn AJ, Peak M, Pirmohamed M, Smyth RL, et al. What can we learn from parents about enhancing participation in pharmacovigilance?. Br J Clin Pharmacol 2012; in press.
8. Gerritsen R, Faddegon H, Dijkers F, van Grootheest K, van Puijenbroek E. Effectiveness of pharmacovigilance training of general practitioners: a retrospective cohort study in the Netherlands comparing two methods. Drug Saf 2011;34:755-62.
9. Kshirsagar N, Ferner R, Figueroa BA, Ghalib H, Lazdin J. Pharmacovigilance methods in public health programmes: the example of miltefosine and visceral leishmaniasis. Trans R Soc Trop Med Hyg 2011;105:61-7.
10. The importance of pharmacovigilance. Geneva: World Health Organization; 2002.
11. Joshi SR, Sapatnekar SM. Pharmacovigilance in India: how safe are the new drugs? How sure are we?. J Assoc Physicians Ind 2008;56:933-4.
12. Olsson S. Pharmacovigilance training with focus on India. Ind J Pharmacol 2008;40: S28-S30.
13. WHO, Pharmacovigilance: ensuring the safe use of medicines, Geneva: WHO 2004.
14. The Importance of Pharmacovigilance, WHO 2002.
15. WHO Medicines Strategy: Framework for Action in Essential Drugs and Medicines Policy 2000-2003.
16. Olsson S. The role of the WHO Programme for International Drug Monitoring in coordinating worldwide drug safety efforts. Drug Saf 1998;19:1-10.
17. Coulter DM. The New Zealand intensive medicines monitoring programme in proactive safety surveillance. Pharmacoepidemiol Drug Saf 2000;9:273-80.
18. Folb PI, ten Ham M. Drug monitoring in developing countries: A drug regulator's perspective. Drug Inf J 1995;29:303-5.

19. Talbot JC, Nilsson BS. Pharmacovigilance in the pharmaceutical industry. Br J Clin Pharmacol 1998;45:427-31.
20. Moore N. The role of the clinical pharmacologist in the management of ADRs. Drug Saf 2001;24:1-7.
21. Hall M, McCormack P, Arthurs N, Feely J. The spontaneous reporting of ADRs by nurses. Br J Clin Pharmacol 1995;40:173-5.
22. Hornbuckle K, Wu HH, Fung MC. Evaluation of spontaneous adverse event reports by primary reporter: A 15-year review (1983 to 1997) Drug Inf J 1999;33:1117-24.
23. Jeetu G, Anusha G. Pharmacovigilance: a worldwide master key for drug safety monitoring. J Young Pharm 2010;2:315-20.
24. WHO. Safety monitoring of medicines. Guidelines for setting up and running a pharmacovigilance centre. Geneva, World Health Organization, 2000.
25. Department of CDSCO, Ministry of health & family welfare, Government of India: Protocol for national pharmacovigilance programme, November 2004.
26. WHO. The World Health Organization Collaborating Centre for International Drug Monitoring [web site]. The World Health Organization Collaborating Centre for International Drug Monitoring, Uppsala.
27. Biswas P. Pharmacovigilance: Safety matters. Ind J Phamacol 2008;40:1-2.
28. Dikshit RK. Challenges in pharmacovigilance. Ind J Pharmacol 2010;42:333

AN OVERVIEW ON DRY EYE DISEASE

Shilpi Arora, Upma,

Baba Mastnath University, Faculty of Pharmacy, Rohtak, Haryana.

ABSTRACT

Dry eye is a disease of the ocular surface attributable to different disturbances of the natural function and protective mechanisms of the external eye, leading to an unstable tear film during the open eye state. Managing the eye inflammation

proved helpful to patients with dry eye disease and current treatment is based on the use of topically applied artificial tear products/lubricants, tear retention management, stimulation of tear secretion and using anti-inflammatory drugs.

INTRODUCTION

Dry eye disease is the outcome of many factors resulting in inflammation of the cornea and conjunctiva. The dysfunction of the tear secretory glands leads to changes in tear composition such as hyper-osmolarity which stimulates the production of inflammatory mediators on the ocular surface. This inflammation can be initiated either by chronic irritative stress like contact lens wearing or a systemic inflammatory autoimmune disease like rheumatoid arthritis. Anti-inflammatory drugs are widely used for the treatment of the inflammation produced by the disease with the topical corticosteroid drops being the most common therapy. Corticosteroids can rapidly and effectively relieve the symptoms and signs of moderate or severe dry eye. Steroids on the other hand produce severe side effects after prolonged use. The effects include risk of bacterial or fungal infection, elevated intraocular pressure and cataract formation, therefore steroids are typically used only for one to two weeks in dry eye patients. As a consequence, non-steroidal antiinflammatorydrugs (NSAID) are increasingly used as dryeye treatment instead of steroids because of their non-severe side effects and because steroids locally suppress the immuneresponse in patients with an already compromised ocular surface. The NSAIDs acutely decrease the eye discomfort

due to its analgesic effect and furthermore is reducing the inflammation. In 2002 U.S. Food and Drug Administration approvedthe drug RESTASIS_ of the company Allergan as thefirst prescription medicine helping to increase tear productionreduced by inflammation due to chronic dry eye disease.

CLASSIFICATION

The major classes of DE, as identified by the International Dry Eye Workshop (DEWS) report are aqueous deficient dry eye (ADDE) and evaporative dry eye (EDE). Although both ADDE and EDE present with similar signs of reduced stability and increased tear film osmolarity, ADDE chiefly refers to a failure of lacrimal secretion and EDE is due to excessive water loss from the exposed ocular surface in the presence of normal lacrimal secretory function. It is also important to recognize that ADDE and EDE may coexist.

PREVALENCE OF DRY EYE-PATIENT FACTORS

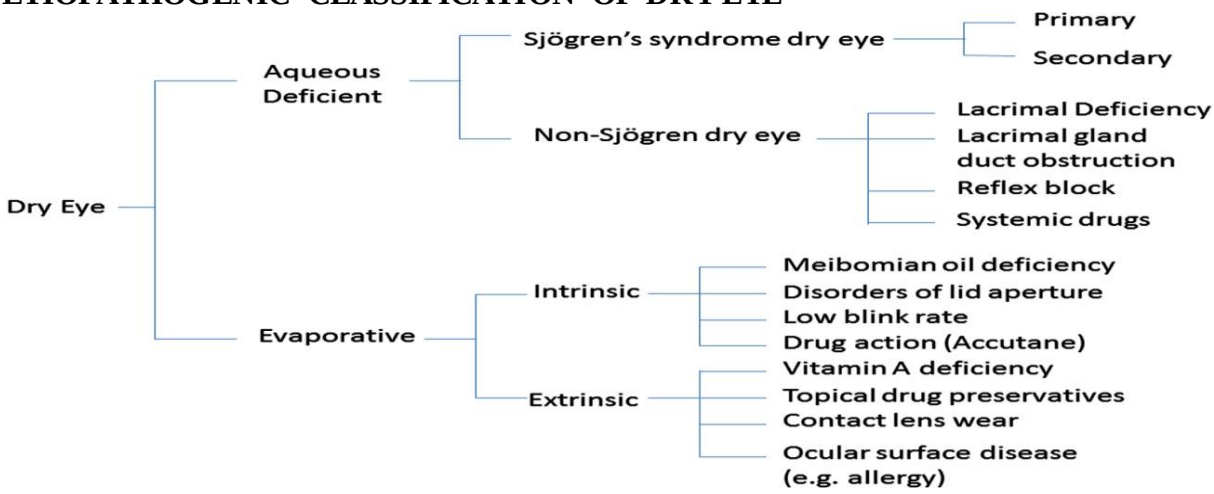
- Older age
- Female gender
- Post-menopausal
- Tobacco smoking
- Contact lens wear
- Prolonged staring (e.g. computer work)

PREVALENCE OF DRY EYE-ENVIRONMENTAL FACTORS

- Air Pollution
- Artificial, forced air

- Allergens
- Low humidity

ETIOPATHOGENIC CLASSIFICATION OF DRY EYE



SYMPTOMS OF DRY EYE

- Stinging or burning
- A sandy or gritty feeling, as if something is in the eye
- Stringy discharge
- Pain and redness
- Periods of blurred vision;
- Heavy eyelids
- Inability to cry
- Uncomfortable contact lenses
- Difficulty reading, working on the computer, or any activity that requires focused vision
- Eye fatigue

DIAGNOSIS OF DRY EYE

- ❑ **Subjective evaluation**: validated questionnaires in recording symptomatic information
- ❑ **Objective evaluation**

- Corneal staining
- Tear film assessment
- Conjunctival staining
- Schirmer test

TREATMENT RECOMMENDATION BY SEVERITY LEVEL

Level 1:
Education and environmental/dietary modifications
Elimination of offending systemic medications
Artificial tear substitutes, gels/ointments
Eye lid therapy

Level 2:
If Level 1 treatments are inadequate, add:
Anti-inflammatories
Tetracyclines (for meibomianitis, rosacea)
Punctal plugs
Secretogogues
Moisture chamber spectacles

Level 3:
If Level 2 treatments are inadequate, add:
Serum
Contact lenses
Permanent punctal occlusion

Level 4:
If Level 3 treatments are inadequate, add:
Systemic anti-inflammatory agents
Surgery (lid surgery, tarsorrhaphy; mucus membrane, salivary gland, amniotic membrane transplantation)

OTC: Artifical Tear Solution, Gels and Ointments

Product	Lubricating Agents	Preservatives
Akwa Tears Ointment	Lanolin 2%, mineral oil 15%, white petrolatum 83%	No preservative
Bausch & Lomb Advanced Eye Relief Dry Eye Environmental Lubricant Eye Drops	Glycerin 1%	Benzalkonium chloride 0.01%, EDTA
Bausch & Lomb Advanced Eye Relief Night Time Lubricant Eye Ointment	Mineral oil 20%, white petrolatum 80%	No preservative
Bion Tears, Single-Use Vials	Dextran 70 0.1%, HPMC 0.3%	No preservative
GenTeal Gel Drops	CMC 0.25%, HPMC 0.3%	GenAqua[a]
GenTeal Mild	HPMC 0.2%	GenAqua[a]
GenTeal PM	Mineral oil 15%, white petrolatum 85%	No preservative
Hypotears	PVA 1%, polyethylene glycol 400 1%	Benzalkonium chloride
Murine Tears Dry Eyes	PVA 0.5%, povidone 0.6%	Benzalkonium chloride, EDTA
Refresh Celluvisc	CMC 1%	No preservative
Refresh Classic	PVA 1.4%, povidone 0.6%	No preservative
Refresh Dry Eye Therapy Sensitive	Glycerin 1%, polysorbate 80 1%	No preservative
Refresh Liquigel	CMC 1%	Purite[b]
Refresh Optive	CMC 0.5%, glycerin 0.9%	Purite[b,c]
Refresh PM	Mineral oil 42.5%, white petrolatum 57.3%	No preservative
Soothe XP	Light mineral oil 1.0%, mineral oil 4.5%	EDTA, octoxynol-40, PHMB
Systane	Polyethylene glycol 400 0.4%, propylene glycol 0.3%	Polyquad[c,d]
Systane Nighttime Ointment	Mineral oil 3%, white petrolatum 94%	Anhydrous liquid lanolin 3%
Tears Naturale Forte	HPMC 0.3%, glycerin 0.2%, dextran 70 0.1%	Polyquad[d] 0.001%
Tears Naturale PM	Mineral oil 3%, white petrolatum 93%	No preservative
Tears Naturale II Polyquad	HPMC 0.3%, dextran 70 0.1%	Polyquad[c,d] 0.001%
Tears Renewed	HPMC 0.3%, dextran 70 0.1%	Benzalkonium chloride, EDTA
TheraTears (bottle)	CMC 0.25%	Sodium perborate[c]
Visine Tears Dry Eye Relief	Glycerin 0.2%, HPMC 0.2%, polyethylene glycol 400 1%	Benzalkonium chloride[c]
Visine Tears Long Lasting Dry Eye Relief	Glycerin 0.2%, HPMC 0.36%, polyethylene glycol 400 1%	Benzalkonium chloride

[a] GenAqua: sodium perbonate. [b] Purite: sodium chlorite. [c] Also available preservative free. [d] Polyquad: polyquaternium. CMC: carboxymethylcellulose; EDTA: ethylenediaminetetraacetic acid; HPMC: hydroxypropyl methylcellulose; PHMB: polyhexamethylene biguanide; PVA: polyvinyl alcohol.

REFERENCES

1. Research in dry eye: report of the Research Subcommittee of the International Dry WorkShop (2007). Ocul Surf 2007;5(2):179–93.
2. The epidemiology of dry eye disease: report of the Epidemiology Subcommittee of the International Dry Eye WorkShop (2007). Ocul Surf 2007;5(2):93–107.
3. Behrens A, Doyle JJ, Stern L, et al. Dysfunctional tear syndrome: a Delphi approach to reatment recommendations. Cornea 2006;25(8):900–7.
4. McGinnigle S, Naroo SA, Eperjesi F. Evaluation of dry eye. Surv Ophthalmol 2012;57(4):293–316.

5. Serin D, Karsloglu S, Kyan A, Alagoz G. A simple approach to themrepeatability of the Schirmer test without anesthesia: eyes open or closed? Cornea 2007;26(8):903–6.
6. Lemp MA, Bron AJ, Baudouin C, et al. Tear osmolarity in the diagnosis and managementof dry eye disease. Am J Ophthalmol 2011;151(5):792–8, e1.
7. Gilbard JP. Tear film osmolarity and keratoconjunctivitis sicca. CLAO J 1985;11(3):243–50.
8. Dana R, Chauhan S, Therapeutic compositions for treatment of ocular inflammatory disorders, WO/2009/0890362009.
9. Rouvier E, Luciani MF,Golstein P. Fas involvement inCa(2+)-independentT cell-mediated cytotoxicity. J ExpMed 1993;177(1):195–200.
10. Miossec P, Kolls JK. Targeting IL-17 and TH17 cells in chronic inflammation. Nat Rev Drug Dis 2012;11(10):763–76.

LC-MS METHOD FOR QUANTITATIVE BIOANALYSIS

Priti mehndiratta, Jyoti Dahiya

Department of Pharmaceutical Sciences & Research, BMU

Bioanalytical methods employed for the quantitative determination of drugs and their metabolites in biological matrix (plasma, urine, saliva, serum etc.) play a significant role in evaluation and interpretation of bioavailability, bioequivalence and pharmacokinetic data. Chromatographic methods such as Gas Chromatography (GC), Liquid Chromatography Mass Spectrometry (LC-MS) etc. are commonly used in laboratories for the qualitative and quantitative analysis of drug substances and biological samples throughout all phases of method development of a drug in research and quality control. Further, method validation is carried out to ensure that the method developed is accurate, specific, reproducible and rugged over the specified range in which an analyte is analyzed. As per Bioanalytical Method Validation (BMV) guidelines for industry, these guidelines are applied to bioanalytical methods that are used for the quantitative determination of drugs and their metabolites in biological matrices such as plasma, urine and preclinical studies

CONCEPTS OF BIOAVAILABILITY AND BIOEQUIVALENCE

Bioavailability and bioequivalence studies are required by regulatory bodies to ensure therapeutic equivalence between a pharmaceutically equivalent test product and a reference product. Both bioavailability and bioequivalence focus on the release of drug substance from its dosage form and subsequent absorption into the systemic circulation. For this reason, similar approaches to measure bioavailability should generally be followed in demonstrating bioequivalence.

SAMPLE COLLECTION AND SAMPLE PREPARATION

The biological media that contain the analyte are usually blood, plasma, urine, serum etc. Blood is usually collected from human subjects by vein puncture with a hypodermic syringe up to 5 to 7 ml (depending on the assay sensitivity and the total number of samples taken for a study being performed). The venous blood is withdrawn into tubes with an anticoagulant, e.g. EDTA, heparin etc. Plasma is obtained by centrifugation at 4000 rpm for 15 min. About 30 to 50% of the original volume is collected.

PROTEIN PRECIPITATION OR DENATURATION

Precipitation is most widely used in processing of biomolecules or biological products such as proteins. Precipitation is usually induced by addition of a neutral salt such as ammonium sulfate which compresses the salvation layer and increases the protein-protein interaction or by changing the pH to alter the nature of the solution. In this process, precipitation of proteins is done by trichloroacetic acid, perchloric acid, methanol, acetonitrile, or by proteolysis enzymes etc.

LIQUID-LIQUID EXTRACTION

It is based on the principles of differential solubility and partitioning equilibrium of analyte molecules between aqueous (the original sample) and the organic phases. Liquid-liquid extraction generally involves the extraction of a substance from one liquid phase to another liquid phase.

SOLID PHASE EXTRACTION

Solid-phase extraction (SPE) is used to isolate analytes of interest from a wide variety of matrices, including urine, blood, water samples, beverages, soil, animal tissue, and consumerproducts. The separation of desired analyte depends upon its affinity towards stationary phase. Analyte which is retained on the stationary phase can then be eluted from the solid phase extraction cartridge with the appropriate solvent.Types of solid phase

extraction cartridges are, hlbcartridge,mcxcartridge,maxcartridge,wcxcartridgewax,cartridgebond and elute plexa.

CHOOSING THE PROPER PACKING TYPE FOR SOLID PHASE EXTRACTION

There are generally three types of sorbents used for solid phase extraction **Reversed Phase:** Reversed phase packings such as C18, C8 are the most popular and most widely used for reversed phase.

Normal Phase

Normal phase are used to retain polar compounds from non-polar matrices. Silica, Amino and Alumina are commonly used for reversed phase. Normal phase packing generally requires conditioning with a non-polar solvent and elution is carried with polar solvents.

Ion Exchange

Ion-exchange retains charged compounds or removes ionic interferences. Anions and cations are retained on the resin by exchanging the Anion/Cation in the sample with the Anion/Cation on the resin.

Benefits of Using SPE

SPE gives greater reproducibility as compared to other techniques.

• Cleaner extracts are obtained using Oasis SPE products.
• High recovery of the analyte.
• No emulsion is formed during sample preparation by using SPE.
• SPE gives increased productivity as it is easy to use.
• Sample processing of about 40-50 matrix samples in a batch can be processed with an inexpensive vacuum manifold.

LC-MS/MS Method Development

A large number of laboratories practice analytical chemistry in many diverse ways. Method of analysis are being routinely developed, improved, validated, collaboratively studied and applied. Required chromatographic separations are mainly dependent on the samples to be analyzed. The knowledge of chromatographic procedure is a must for the systematic approach to LC-MS/MS method development.

Procedure for Method Development

• Collection of information from the literature for the physicochemical properties of drug molecules.
• Determine solubility profile.
• MS scanning and optimization.
• Mobile phase selection.
• Selection of extraction method and optimization.
• Selection of chromatographic method (based on solubility study, retention of compound).

Reversed Phase Chromatography

Reversed phase packings such as C18, C8 are the most popular and most widely used for reversed phase.

Normal Phase Chromatography

Normal phase packings include silica, amino and alumina. Normal phase packing generally requires conditioning with a non-polar solvent and elution is carried with polar solvents.

LC-MS/MS Method Development

Proper knowledge about the sample is necessary for an effective method development. Some information regarding the analyte is necessary like:

• Number of compounds present
• Molecular weights of compound
• Sample Solubility
• Drug Stability
• Concentration range of compounds in samples of interest

Approaches for Method Development

Following criteria for the development of an efficient LC-MS/MS method may be opted.

• Reverse phase chromatography is normally tried first.

• Normal phase is tried if reverse phase fails.

Method Optimization

During the optimization stage, the initial sets of conditions that were evolved during the method development are improved and maximized in terms of resolution and peak shape, plate counts asymmetry, capacity, elution time, detection limits, limit of quantitation, and overall ability to quantify the specific analyte of interest. Optimization of a method can follow either of two general approaches such as manual or computer driven.

Mode of Separation Technique

Since most of the pharmaceutical compounds are polar in nature so reverse phase chromatography is normally tried first in which a non-polar stationary phase is used. The mobile phase consists of water or buffer and organic phase (acetonitrile or methanol). Hence polar compounds get eluted first and non-polar compounds are retained for a longer time. The stationary phases used in reverse phase chromatography are n-octadecyl (RP-18), n-octyl (RP-8), ethyl (RP-2), phenyl, cyano, diol and hydrophobic polymers.

Selection of Stationary Phase/Column

Prior to selection of column it is necessary to understand the properties of column packing material. Silica tends to dissolve above pH 8 and cross-linked polymeric particles, for example, polystyrene or poly methacrylates are used for separation of bases, which can withstand strongly basic mobile phase.The most commonly used non-polar bonded phases (for reversed phase chromatography) are C18 and C8 with C18.

Selection of Mobile Phase

The main criterion in selection and optimization of mobile phase is to achieve optimum separation of all the individual impurities and degradrants from each other and from the analyte peak.

METHOD VALIDATION:

Bio-Analytical Method Validation Parameters

The method developed will be validated for system suitability, specificity / selectivity, linearity, precision and accuracy (intra-day or within run/inter-day or within batch), recovery, reinjection reproducibility, ruggedness and stability studies (freeze and thaw stability, bench top stability etc.).

Mass Spectrometric Detection and Data System

Liquid chromatography/mass spectrometry (LC-MS) is fast becoming the preferred tool of liquid chromatography.

Mass Spectrometry

Mass spectrometers are divided into three fundamental parts like ionization source, analyzer and detector

Sample Introduction

The samples can be inserted directly into the ionization source or can also undergo some type of chromatography to the ionization source. This method usually involves the lc-ms technique in which mass spectrometer is coupled directly to (HPLC) or (GC).

Methods of Sample Ionization

Many ionization methods are available each having its own advantages and disadvantages. The ionization method used depends on the type of sample under investigation and the mass spectrometer available. Ionization methods are of many types and include the following:

a) Atmospheric pressure chemical ionization (APCI)

b) Electro spray ionization (ESI)

c) Fast atom bombardment (FAB) and,

d) Matrix assisted laser desorption ionization (MALDI**)**

REFERENCE:

1. Wehr T. LCGC North America. 2002; 20(1): 41-47.
2. Mazzeo J R, Neue U D, kele M, Plumb R S. Anal. Chem. 2005; 77: 460-467 A.
3. Hartmann C, Symeres V J, Massart D L, MCdowall R D. Validation of bioanalytical chromatographic methods Pharma Biomed Anal. 1998; 17: 193-218.
4. Kimanani E K. Bioanalytical calibration curves: proposal for statistical criteria. J Pharma Biomed Anal. 1998; 16:1117-24.

DIABETES

Monika Kaushik , Rajiv Kumar Arora, Minakshi Gupta

Faculty of Pharm. Sciences, Baba Mastnath University, Asthal Bohar, Rohtak

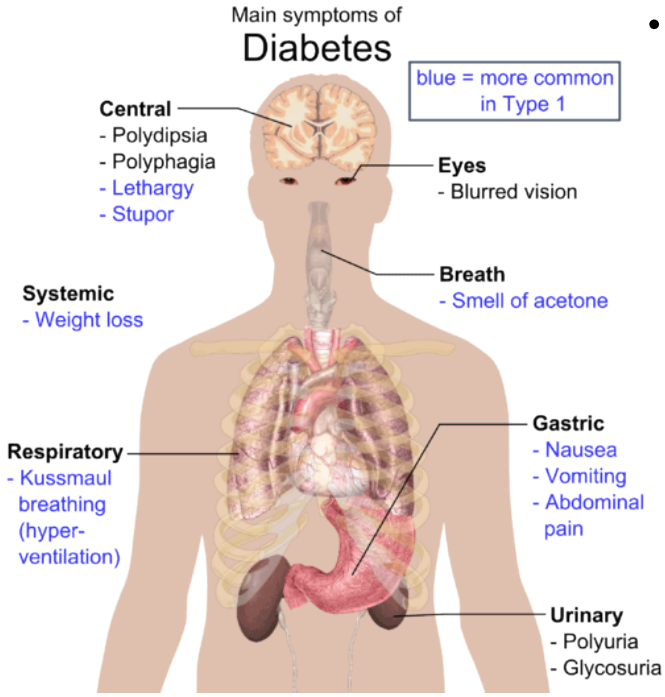
Diabetes is a chronic disease in which the body cannot regulate the amount of sugar in the blood.

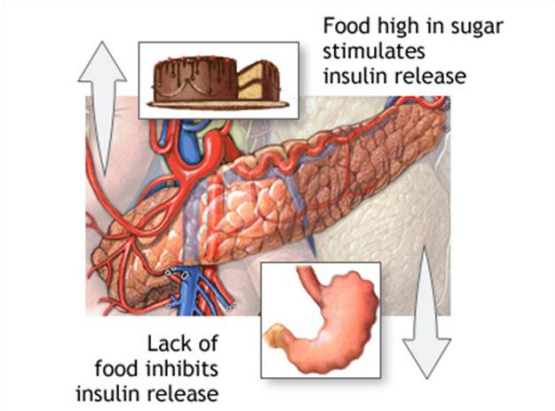
CAUSES

Insulin is a hormone produced by the pancreas to control blood sugar. Diabetes can be caused by too little insulin, resistance to insulin, or both.

To understand diabetes, it is important to first understand the normal process by which food is broken down and used by the body for energy. Several things happen when food is digested:

- A sugar called glucose enters the bloodstream. Glucose is a source of fuel for the body.An organ called the pancreas makes insulin. The role of insulin is to move glucose from the bloodstream into muscle, fat, and liver cells, where it can be stored or used as fuel.



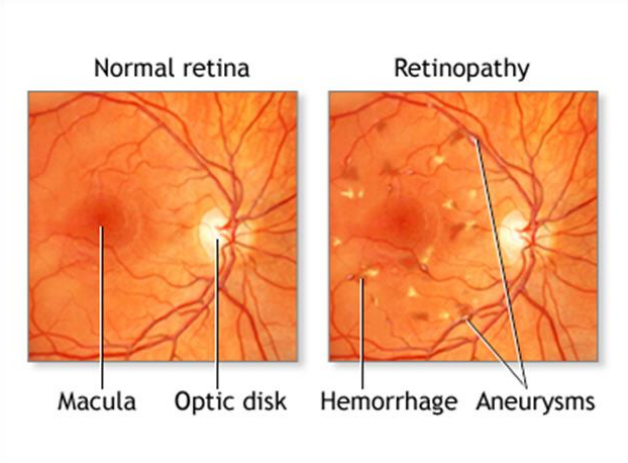


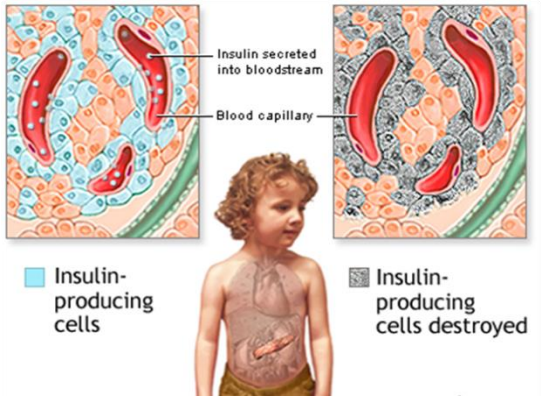
People with diabetes have high blood sugar because their body cannot move sugar from the blood into muscle and fat cells to be burned or stored for energy, and because their liver makes too much glucose and releases it into the blood. This is because either:

- Their pancreas does not make enough insulin
- Their cells do not respond to insulin normally
- Both of the above

There are two major types of diabetes. The causes and risk factors are different for each type:

- Type 1 diabetes can occur at any age, but it is most often diagnosed in children, teens, or young adults. In this disease, the body makes little or no insulin. This is because the pancreas cells that make insulin stop working. Daily injections of insulin are needed. The exact cause is unknown.





- Type 2 diabetes is much more common. It most often occurs in adulthood, but because of high obesity rates, teens and young adults are now being diagnosed with this disease. Some people with type 2 diabetes do not know they have it.
- There are other causes of diabetes, and some patients cannot be classified as type 1 or type 2. Gestational diabetes is high blood sugar that develops at any time during pregnancy in a woman who does not have diabetes.

If your parent, brother, or sister has diabetes, you may be more likely to develop the disease.

SYMPTOMS

A high blood sugar level can cause several symptoms, including:

- Blurry vision
- Excess thirst
- Fatigue
- Frequent urination
- Hunger
- Weight loss

Because type 2 diabetes develops slowly, some people with high blood sugar have no symptoms.

Symptoms of type 1 diabetes develop over a short period. People may be very sick by the time they are diagnosed.

After many years, diabetes can lead to other serious problems. These problems are known as diabetes complications, and include:

- Eye problems, including trouble seeing (especially at night), light sensitivity, and blindness
- Sores and infections of the leg or foot, which untreated can lead to amputation of the leg or foot
- Damage to nerves in the body, causing pain, tingling, a loss of feeling, problems digesting food, and erectile dysfunction
- Kidney problems, which can lead to kidney failure
- Weakened immune system, which can lead to more frequent infections
- Increased chance of having a heart attack or stroke

EXAMS AND TESTS

A urine analysis may show high blood sugar. But a urine test alone does not diagnose diabetes.

Your health care provider may suspect that you have diabetes if your blood sugar level is higher than 200 mg/dL. To confirm the diagnosis, one or more of the following tests must be done.

BLOOD TESTS:

- Fasting blood glucose level -- diabetes is diagnosed if the fasting glucose level is higher than 126 mg/dL on two different tests. Levels between 100 and 126 mg/dL are called impaired fasting glucose or pre-diabetes. These levels are risk factors for type 2 diabetes.
- Hemoglobin A1c (A1C) test --
    - Normal: Less than 5.7%
    - Pre-diabetes: 5.7% - 6.4%
    - Diabetes: 6.5% or higher
- Oral glucose tolerance test -- diabetes is diagnosed if the glucose level is higher than 200 mg/dL 2 hours after drinking a sugar drink. (This test is used more often for type 2 diabetes.)
- Screening for type 2 diabetes in people who have no symptoms is recommended for:
- Overweight children who have other risk factors for diabetes, starting at age 10 and repeated every 3 years
- Overweight adults (BMI of 25 or higher) who have other risk factors
- Adults over age 45, repeated every 3 years

TREATMENT

Type 2 diabetes may be reversed with lifestyle changes, especially losing weight with exercise and by eating healthier foods. Some cases of type 2 diabetes can also be improved with weight-loss surgery.

There is no cure for type 1 diabetes.

Treating either type 1 diabetes or type 2 diabetes involves medicines, diet, and exercise to control blood sugar level.

Everyone with diabetes should receive proper education and support about the best ways to manage their diabetes. Ask your healthcare provider about seeing a diabetes educator.

Getting better control over your blood sugar, cholesterol, and blood pressure levels helps reduce the risk of kidney disease, eye disease, nervous system disease, heart attack, and stroke.

To prevent diabetes complications, visit your health care provider at least two to four times a year. Talk about any problems you are having. Follow your health care provider's instructions on managing your diabetes.

Support Groups

Many resources can help you understand more about diabetes. If you have diabetes, you can also learn ways to manage your condition and prevent diabetes complications.

PREVENTION

Keeping an ideal body weight and an active lifestyle may prevent or delay the start of type 2 diabetes.

Type 1 diabetes cannot be prevented.

BENZOTRIAZOLES: AN OVERVIEW

Savitri Kumari[*], Minakshi Gupta[1], Anusha[1]

Assistant Professor, Faculty of Pharmaceutical Sciences, Baba Mastnath University, Asthal Bohar, Rohtak (124001)

INTRODUCTION

Benzotriazoles have high stabilities both at high temperatures and in presence of ultra violet (UV) light (USEPA, 1977). They have a wide range of industrial uses. Since they complex strongly with some metals, they are commonly used as corrosion inhibitors in glycol-based aircraft deicing fluids (ADFs). The majority of simple benzotriazoles currently produced go into such anticorrosion applications. The structure and ring numbering of 1H-benzotriazole is as shown in Figure 1. Common derivatives include methyl groups at the 4 or 5 position, chlorine at the 5 position, or replacement of the H by OH at the 1 position. Typical applications include the protection of copper containing parts (for which benzotriazole excels), by inclusion of benzotriazoles in automobile antifreeze solutions, in recirculating water systems such as power plant and commercial air-condi-tioning cooling systems, and in coatings for protection of copper alloys in architectural and decorative applications. More complex nonpolar derivatives (generally at the 2N position) are also used widely to stabilize plastics and similar materials against the decomposition that would otherwise take place upon exposure of these materials to UV radiation. Metal chelation ability (to silver) accounts for their use in photography, mainly as an antifogging constituent of films (USEPA, 1977). Recently, 1hydroxybenzotriazole has been proposed to serve as a mediator of laccase action in pulp bleaching to decrease the amount of corrosive bleaching chemicals in the paper-making process (Call and Mucke, 1997).

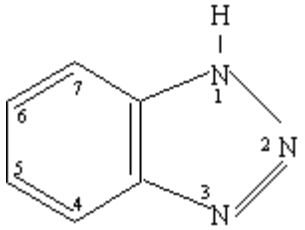


Figure 1. Structure of benzotriazole, showing ring numbering

In modern agro-ecosystems where large quantities of N fertilizers are used continuously, the efficiency of their use may be low and may result in environmental pollution (Puttanna et al., 2001a,b). Much fertilizer N applied to soils is in the form of ammonium or ammonium-producing compounds such as urea. Ammonium is usually oxidized quite rapidly to nitrate by nitrifying microorganisms in soil. The nitrate thus produced is susceptible to loss by leaching and denitrification, and there is international concern about pollution of ground and surface waters by fertilizer-derived nitrate. This concern has stimulated research to find compounds that will effectively inhibit nitrification of fertilizer N when

ANTICANCER BENZOTRIAZOLES

A variety of anticancer drugs such as alkylating agents, platinum complexes, porphyrin drugs and azole agents have been successfully developed and clinically used to treat various cancers. However, most of the clinical anticancer drugs are often toxic to normal tissues, thus causing numerous side effects, which, in turn, limit the treatment efficacy (Figure 1). Long term effectiveness is also limited by dose-related cumulative cardiotoxicity as well as drug resistance . Therefore, an increasing number of researches have been directing towards the design and development of new therapeutic agents for the treatment of cancers. Several benzotriazole derivatives have been found to possess potent anticancer activity, for example,

the antineoplastic agent vorozole that is in clinical trial, and 4,5,6,7-tetrabromobenzotriazole (TBB) (compound 1a) is a commercial available anticancer drug with high selective inhibition against protein kinase CK2. The successful exploration of TBB stimulates the continuous effort towards the development of novel benzotriazole-based anticancer agents targeting various kinases or receptors. Moreover, an increasing number of new structural benzotriazole derivatives as well as benzotriazole-containing metal complexes have displayed considerable potentiality to overcome the diverse drawbacks of currently available clinical drugs. The inhibition of kinases is one of the most important pathways to treat cancers attributing to the significant roles of kinases in cell multiplication. The special structure of benzotriazole derivatives could readily bind with different kinases *via* multiple non-covalent forces such as hydrogen bonds, coordination, ion-dipole, cation-π, π-π stacking, hydrophobic effect and van der Waals force, thus effectively inhibiting the activity of various kinases including protein kinases CK2 and CHK1, histone deacetylases and focal adhesion kinase and so on.

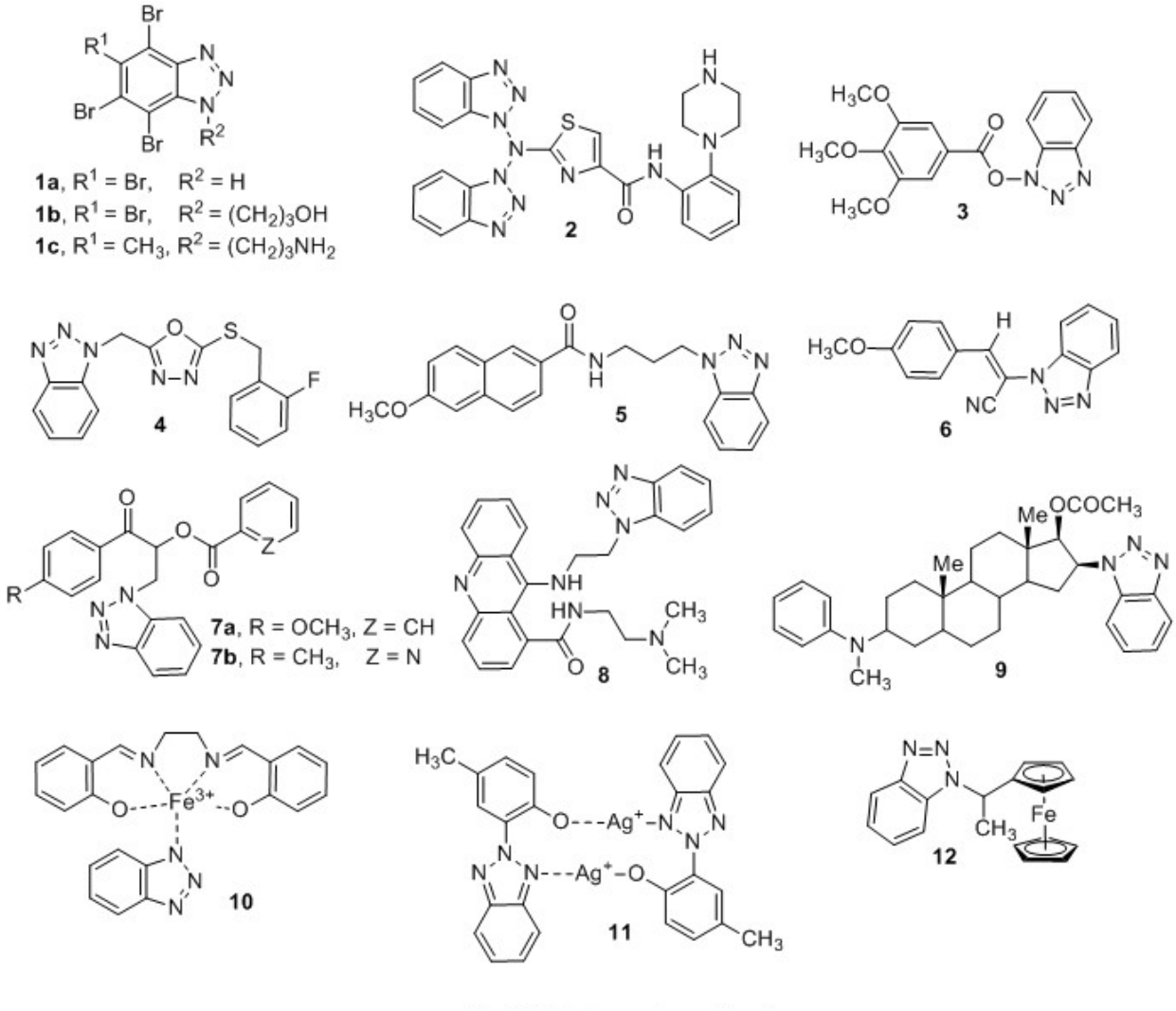


Figure 1: Anticancer Benzotriazoles

ANTIFUNGAL BENZOTRIAZOLES

Fungal infections are a kind of quite prevalent diseases. Among different kinds of antifungal agents, azole compounds have been rapidly developed as the mainstream for fungal infection treatment and are widely used in clinic. Benzotriazole with a benzene ring endures a larger conjugated system than triazole or imidazole as well as a three-nitrogen containing structure could more readily bind with the receptors in organisms with less toxicity. A lot of researches

and exploitations have been devoted to benzotriazoles due to their potentiality as novel antifungal agents.

STRUCTURAL MODIFICATION OF CLINICAL ANTIFUNGAL DRUGS BY BENZOTRIAZOLE RING

A variety of antifungal azoles representing as an important class of nitrogen-containing heterocycles with desirable electron-rich properties, have been early discovered and successfully used to develop clinical agents. With the growing emergence of the intrinsicand acquired antifungal resistance caused by the abuse of available drugs, especially the multidrug-resistant fungi (Figure 2), it is urgent to develop novel structural agents with more efficiency, less toxicity, better lipophilicity and stronger antifungal ability. Notably, the structural modification of clinical azole antifungal drugs like fluconazole and clotrimazole is regarded as a helpful strategy to improve their physicochemical property and binding affinity, overcome their
shortcomings, and effectively broaden their antifungal spectrum

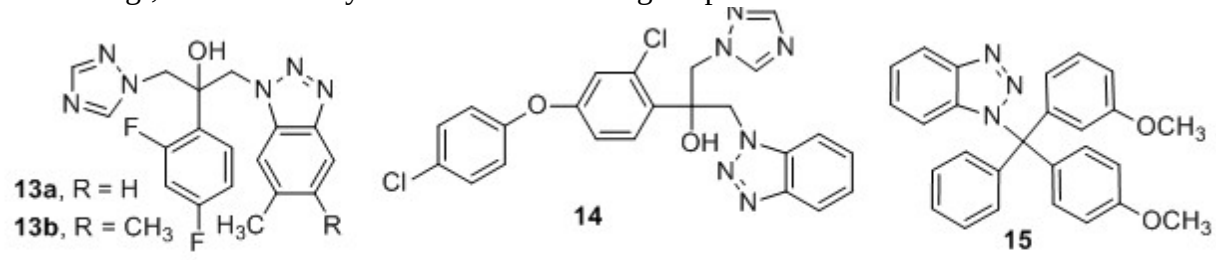


Fig. (2). Structural Modification of Clinical Antifungal Drugs by Benzotriazole Ring

NEW STRUCTURAL BENZOTRIAZOLES AS ANTIFUNGAL AGENTS

The combination of multiple functional groups with different action modes into one molecule could produce new antifungal agents. Heterocyclic molecules usually containing N, O or S heteroatom in their cyclic structures as one of the most active classes of compounds possess a wide spectrum of biological activities, and have showed large potentiality in pharmaceutical science (Figure 3). The introduction of benzotriazole ring into other heterocyclic scaffolds to form some new structural compounds with improved antifungal ability has attracted increasing attention in medicinal chemistry, and this field is worthy of further investigations

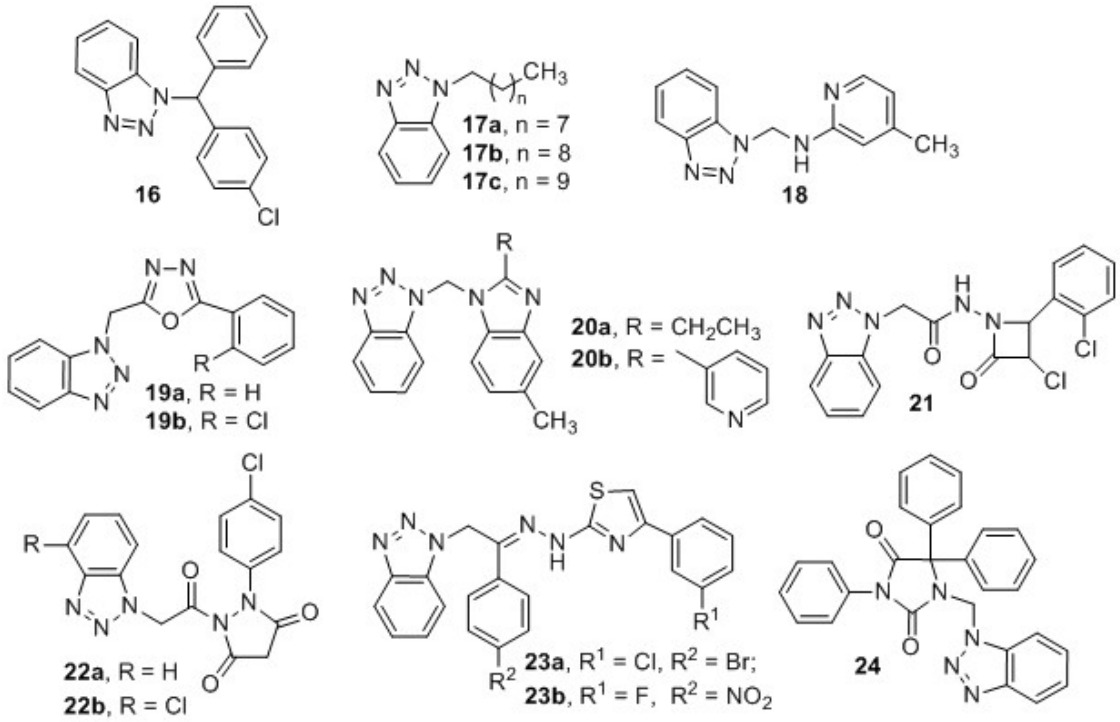


Fig. (3). New Structural Benzotriazoles as Antifungal Agents

ANTIBACTERIAL BENZOTRIAZOLES

Bacterial infections are frequently occurring infective diseases all around the world, particularly in Indian subcontinent, portions of South America and tropical fraction of Africa. The morbidity and mortality caused by food poisoning, rheumatic, salmonellosis of diarrhea from bacterial infection are the major healthy problems [89,90]. Despite a lot of antibiotics and chemotherapeutics like *beta*-lactams, tetracyclines, aminoglycosides, macrolides, polyenes etc. and synthetic drugs such as sulfonamides, quinolones, oxazolidones, allylamines and so on are

available for clinical use, the treatment of bacterial infectious diseases still remains an important and challenging problem due to a series of factors such as emerging infectious diseases, severely adverse effects, narrow antibacterial spectrum as well as single dosage form [91,92].

More importantly, an increasing number of multidrug resistant microbial pathogens like methicillin-resistant *Staphylococcus aureus* (MRSA) and carbapenems-resistant *Enterobacteriaceae* force a real need to develop new compounds acting through distinct mechanisms from the well-known classes of antibacterial agents. The development of benzotriazole derivatives as antibacterial drugs has become a rapidly developing field with considerable breakthroughs .

STRUCTURAL MODIFICATION OF CLINICAL ANTIBACTERIAL DRUGS BY BENZOTRIAZOLE RING

Structural modification of clinical antibacterial drugs to broaden their antimicrobial spectrum and increase therapeutic indexes has provoked special interest in the realm of medicinal chemistry. Some researches have manifested that the incorporation of benzotriazole ring

into clinical drugs could evidently improve their antibacterial efficiency and reduce cytotoxicity (Figure 5).

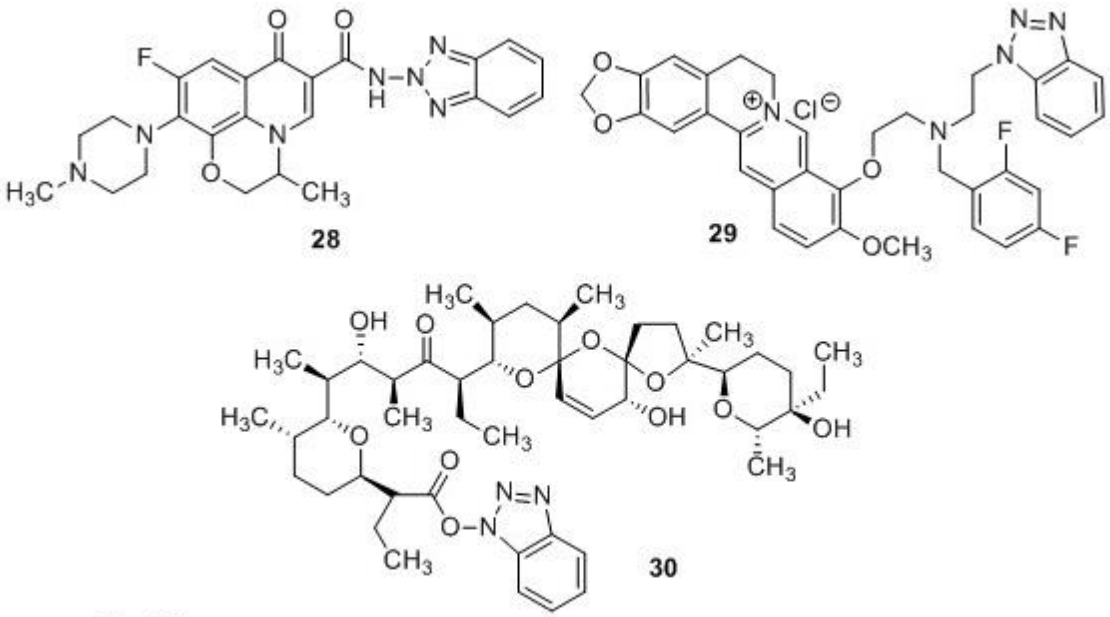


Fig. (5). Structural Modification of Clinical Antibacterial drugs by Benzotriazole Ring

NANOPARTICLES: AN INTRODUCTION TO LIPOSOMES, SOLID LIQUID NANOPARTICLES AND POLYMERIC NANOPARTICLES

Sneh Lata*, Savitri, Ruchi Goyal

Faculty of Pharmaceutical Sciences, Baba Mastnath University, Asthal Bohar, Rohtak-124001

ABSTRACT

Controlled drug delivery systems have many advantages as compared to the traditional forms of drugs. A drug is transported to the site of action and thus its influence on vital tissues and undesirable side effects can be minimized. Accumulation of therapeutic compounds in the target site increases and consequently, the required doses of drugs are lower. This modern form of therapy is especially important when there is a discrepancy between the dose or the concentration of a drug and its therapeutic results or toxic effects. Cell-specific targeting can be accomplished by attaching drugs to specially designed carriers. Various nanostructures, including liposomes, polymers, dendrimers, silicon or carbon materials and magnetic nanoparticles, have been tested as carriers in drug delivery systems. In this article special attention is paid to the functionalization of magnetic nanoparticles as carriers in drug delivery systems.

Key words:

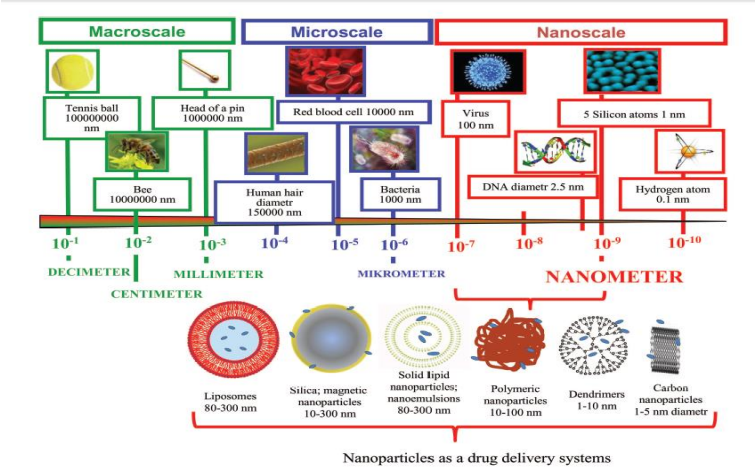
Drug delivery system, nanocarriers, nanoparticles, magnetic nanoparticles.

INTRODUCTION

In controlled drug delivery systems (DDS) the drug is transported to the place of action, thus, its influence on vital tissues and undesirable side effects can be minimized. In addition, DDS protects the drug from rapid degradation or clearance and enhances drug concentration in target tissues, therefore, lower doses of drug are required. This modern form of therapy is especially important when there is a discrepancy between a dose or concentration of a drug and its therapeutic results or toxic effects. Cell-specific targeting can be achieved by attaching drugs to individually designed carriers. Recent developments in nanotechnology have shown that nanoparticles (structures smaller than 100 nm in at least one dimension) have a great potential as drug carriers. Due to their small sizes, the nanostructures exhibit unique physicochemical and biological properties (e.g., an enhanced reactive area as well as an ability to cross cell and tissue barriers) that make them a favorable material for biomedical applications.

NANOCARRIERS

The prefix "nano" is commonly used for particles that are up to several hundred nanometers in size. According to the definition from NNI (*National Nanotechnology Initiative*), nanoparticles are structures of sizes ranging from 1 to

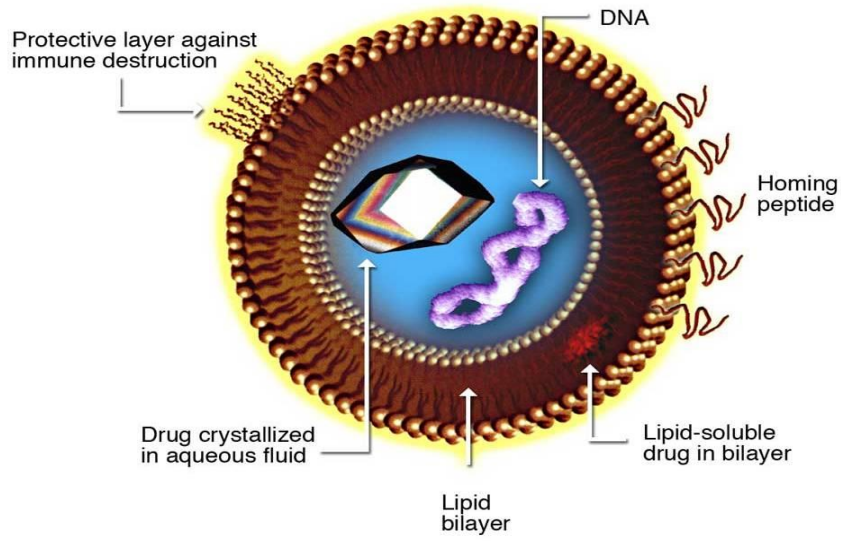


100nm in at least one dimension. Liposomes, solid lipids nanoparticles, dendrimers, polymers, silicon or carbon materials and magnetic nanoparticles are the examples of

nanocarriers that have been tested as drug delivery systems. The way of conjugating the drug to the nanocarrier and the strategy of its targeting is highly important for a targeted therapy. A drug may be adsorbed or covalently attached to the nanocarriers surface or else it can be encapsulated into it. Covalent linking has the advantage over other ways of attaching as it enables to control the number of drug molecules connected to the nanocarrier, i.e., a precise control of the amount of therapeutic compound delivered. Cell-specific targeting with nanocarriers may be accomplished by using active or passive mechanisms. The first strategy relies on the attraction of a drug – the nanocarriers conjugate to the affected site by using recognition ligands, attached to the surface of conjugates antibodies, low molecular ligands, e.g., folic acids, peptides, etc. The active strategy can be also achieved through a manipulation of physical stimuli (e.g., temperature, pH, magnetism).

LIPOSOMES

Liposomes have been the first to be investigated as drug carriers. They are nano/micro-particular or colloidal carriers, usually with 80–300 nm size range. They are spherical vesicles composed of phospholipids and steroids (e.g., cholesterol), bilayers, or other surfactants and form spontaneously when certain lipids are dispersed in aqueous media where liposomes can be prepared, e.g., by sonication. Liposomes have been reported to increase the solubility of drugs and improve their pharmacokinetic properties, such as the therapeutic index of chemotherapeutic agents, rapid metabolism, reduce harmful side effects and increase of *in vitro* and *in vivo* anticancer activity. A drug is incorporated in liposomes by the encapsulation process. The release of a drug from liposomes depends on the liposome composition, pH, osmotic gradient, and the surrounding environment. Additionally, a prolonged residence time increases the duration of action of such particles, but decreases their number. Interactions of liposomes with cells can be realized by: adsorption, fusion, endocytosis, and lipid transfer. There are a lot of drug examples in liposomal formulations, such as anticancer drugs, neurotransmitters (serotonin) [2], antibiotics, anti-inflammatory and antirheumatic drugs.



Recent studies have reported the clinical outcomes and side effects of photodynamic therapy (PDT) by means of intense pulsed light (IPL) and spray (liposome encapsulated 0.5% 5-aminolevulinic acid) which was used for the treatment of inflammatory facial acne. Modified liposomes are an interesting type of such lipid structures. The multifunctional liposomes, containing the specific proteins, antigens, or other biological substances, can be used to design drugs which act selectively on a particular tissue. It is a promising approach for targeted delivery of therapeutics. Biswas et al. presented hydrazine-functionalized poly-(ethylene glycol)-phosphatidylethanolamine (PEG-PE)- based amphiphilic polymer which can conjugate a variety of ligands. The researchers investigated the reversible model ligands monoclonal antinucleosome antibody 2C5 and antimyosin

antibody 2G4, as well as glycoproteins concanavalin A (Con-A). The re-versible attachment of homing devices is useful especially in modified liposomal systems, whereafter they successfully perform the function of targeting at the specific site.

NANOPARTICLES BASED ON SOLID LIPIDS

SLN (solid lipid nanoparticles), NLC (nanostructured lipid carriers) and LDC (lipid drug conjugates) are types of carrier systems based on solid lipid matrix, i.e., lipids solid at the body temperature . They have been exploited for the dermal [1], peroral, parenteral, ocular [13], plumonary, and rectal delivery. SLN are particles made of solid lipids, e.g., highly purified triglycerides, complex glyceride mixtures or waxes stabilized by various surfactants. The main characteristics of SLN include a good physical stability, protection of incorporated drugs from degradation, controlled drug release, and good tolerability. Additionally, some disadvantages have been observed, such as low loading capacity (limited by the solubility of drug in the lipid and the structure and polymorphic state of the lipid matrix), drug expulsion after crystallization, and relatively high water content of the dispersions. NLC and LDC are modifications of lipid based nanoparticles that have been developed to overcome limitations of conventional SLN. NLC are produced by mixing solid lipids with liquid lipids, which leads to special nanostructure with increased payload and prevented drug expulsion. Three types of NLC have been introduced: imperfect type NLC (general imperfections in the matrix nanostructure form free spaces for the accommodation of the guest molecules), multiple type NLC (drugs are solved in oils and protected from degradation by the surrounding solid lipid) and amorphous type NLC (the crystallization that causes drug expulsion is avoided).

POLYMERIC NANOPARTICLES

Polymeric nanoparticles (PNPs) are structures with a diameter ranging from 10 to 100 nm. The PNPs are obtained from synthetic polymers, such as poly-*e*-caprolactone [20], polyacrylamide [14] and polyacrylate, or natural polymers, e.g., albumin, DNA, chitosan and gelatin.

Based on *in vivo* behavior, PNPs may be classified as biodegradable, i.e., poly(L-lactide) (PLA), polyglycolide (PGA), and non-biodegradable, e.g.,polyurethane. PNPs are usually coated with nonionic surfactants in order to reduce immunological interactions (e.g., opsonization or presentation PNPs to CD8 T-lymphocytes) as well as intermolecular interactions between the surface chemical groups of PNPs (e.g., van der Waals forces, hydrophobic interaction or hydrogen bonding). Drugs can be immobilized on PNPs surface after a polymerization reaction or can be encapsulated on PNP structure during a polymerization step. Moreover, drugs may be released by desorption, diffusion, or nanoparticle erosion in target tissue. Among the aforementioned applications the one that is particularly interesting is the immobilization of retinyl acetate (RA) on ethyl cellulose (EC), which improves aqueous stability and photostability of a drug. In *ex vivo* tests on a skin tissue of mice, a 100% absorption of RA after 24 h has been demonstrated [8]. It is also worth to point out the biodegradeable thermo-responsive chitosan-*g*-poly(*N*-vinylcaprolactam)-biopolymer used for the delivery of 5-fluorouracil to cancer cells. The hypothesized mechanism of 5-FU controlled release from this polymeric nanocarrier is swelling followed by conformational changes during a LCST (lower critical solution temperature) transition. The *in vitro* drug release showed a significant release above LCST. The high toxicity to cancer cells, comparatively lower to the normal ones, was observed.

CONCLUSION

Nanocarriers as drug delivery systems are designed to improve the pharmacological and therapeutic properties of conventional drugs. The incorporation of drug molecules into nanocarrier can protect a drug against degradation as well as offers possibilities of targeting and controlled release. Due to small dimensions, nanocarriers are able to cross the blood-brain-barrier (BBB) and operate on cellular level. In comparison with the traditional form of

drugs, nanocarrier-drug conjugates are more effective and selective. They can reduce the toxicity and other adverse side effects in normal tissues by accumulating drugs in target sites. In consequence, the required doses of drugs are lower. Although there are several nanoparticle-based therapeutic agents, being developed and are under preclinical evaluation, only a handful of nanoparticle drugs are available on the pharmaceutical market, e.g., liposomal conjugates: Doxil® (doxorubicin) or DaunoXome® (daunorubicin). It is due to the fact that nanoparticle based drug delivery systems do have a lot of drawbacks and limitations. Some of them arise from scaling up problems. For instance, small size and large surface area of nanoparticle-based targeting system can lead to an aggregation, making physical handling difficult. Despite all the limitations and shortcomings, nanoparticle DDS which respond to slight changes in the local cellular environment have a potential to resolve many of the current drug delivery problems. The key applications of nanoparticles in medicine are as diagnostic agents and in target therapy, however, their wider use is still the future.

REFERENCES:

1. Abdel-Mottaleb MMA, Neumann D, Lamprecht A: Lipid nanocapsules for dermal application: A comparative study of lipid-based versus polymer-based nanocarriers. Eur J Pharm Biopharm, 2011, 79, 36–42.
2. Afergan E, Epstein H, Dahan R, Koroukhov N, Rohekar K, Danenberg HD, Golomb G: Delivery of serotonin to the brain by monocytes following phagocytosis of liposomes. J Control Release, 2008, 132, 84–90.
3. Agnihotri SA, Aminabhavi TM: Novel interpenetrating network chitosan-poly(ethylene oxide-*g*-acrylamide)hydrogel microspheres for the controlled release of capecitabine. Int J Pharm, 2006, 324, 103–115.
4. Ahola MS, Säilynoja ES, Raitavuo MH, Vaahtio MM, Salonen JI, Yli-Urpo AU: In vitro release of heparin from silica xerogels. Biomaterials, 2001, 22, 2163–2170.
5. Ai J, Biazar E, Montazeri M, Majdi A, Aminifard S, Safari M, Akbari HR: Nanotoxicology and nanoparticle safety in biomedical designs. Int J Nanomedicine, 2011, 6, 1117–1127.
6. Ajima K, Murakami T, Mizoguchi Y, Tsuchida K, Ichihashi T, Iijima S, Yudasaka M: Enhancement of in vivo anticancer effects of cisplatin by incorporation inside singlewall carbon nanohorns. ACS Nano, 2008, 2, 2057–2064.
7. Amato G: Silica-encapsulated efficient and stable si quantum dots with high biocompatibility. Nanoscale Res Lett, 2010, 5, 1156–1160.
8. Arayachukeat S, Wanichwecharungruang SP, Tree-Udom T: Retinyl acetate-loaded nanoparticles: Dermal penetration and release of the retinyl acetale. Int J Pharm, 2011, 404, 281–288.
9. Arias JL, López-Viota M, Delgado AV, Ruiz MA: Iron/ethylcellulose (core/shell) nanoplatform loaded with 5 fluorouracil for cancer targeting. Colloids Surf B Biointerfaces, 2010, 77, 111–116.
10. Arruebo M, Fernández-Pacheco R, Ibarra, MR, Santamaría J: Magnetic nanoparticles for drug delivery. Nano Today, 2007, 2, 22–32.
11. Arsawang U, Saengsawang O, Rungrotmongkol T, Sornmee P, Wittayanarakul K, Remsungnen T, Hannongbua S: How do carbon nanotubes serve as carriers for gemcitabine transport in a drug delivery system? J Mol Graph Model, 2011, 29, 591–596.
12. Asmatulu R, Zalich MA, Claus RO, Riffle J: Synthesis, characterization and targeting of biodegradable magnetic nanocomposite particles by external magnetic fields. J Magn Magn Mater, 2005, 292, 108–119.
13. Attama AA, Schicke BC, Paepenmüller T, Müller- Goymann CC: Solid lipid nanodispersions containing mixed lipid core and a polar heterolipid: characterization. Eur J Pharm Biopharm, 2007, 67, 48–57.
14. Bai J, Li Y, Du J, Wang S, Zheng J, Yang O, Chen X: One-pot synthesis of polyacrylamide-gold nanocomposite. Mater Chem Phys, 2007, 106, 412–415.
15. Bajpai AK, Gupta R: Magnetically mediated release of ciprofloxacin from polyvinyl alcohol based superparamagnetic nanocomposites. J Mater Sci Mater Med, 2011, 22, 357–369.
16. Balogh L, Swanson DR, Tomalia DA, Hagnauer GL, McManus AT: Dendrimer-silver complexes and nanocomposites as antimicrobial agents. Nano Lett, 2001, 1, 18–21.
17. Beg S, Rizwan M, Sheikh AM, Hasnain MS, Anwer K, Kohli K: Advancement in carbon nanotubes: basics, biomedical applications and toxicity. J Pharm Pharmacol, 2011, 63, 141–163.
18. Bhirde AA, Patel S, Sousa AA, Patel V, Molinolo AA, Ji Y, Leapman RD et al.: Distribution and clearance of PEG-single-walled carbon nanotube cancer drug delivery vehicles in mice. Nanomedicine, 2010, 5, 1535–1546.

19. Bhirde AA, Patel V, Gavard J, Zhang G, Sousa AA, Masedunskas A, Leapman RD et al.: Targeted killing of cancer cells in vivo and in vitro with EGF-directed carbon nanotube-based drug delivery. ACS Nano, 2009, 3, 307–316.
20. Bilensoy E, Sarisozen C, Esendagl G, Dogan LA, Aktas Y, Sen M, Mangan AN: Intravesical cationic nanoparticles of chitosan and polycaprolactone for the delivery of Mitomycin C to bladder tumors. Int J Pharm, 2009, 371, 170–176.

GUM ARABIC: GREEN BIOPOLYMER USED IN PHARMACEUTICALS

Bijender Sagar[1*]

Assistant Professor,[1]Shri Baba Mastnath Institute of Pharmaceutical Sciences and Research, Asthal Bohar, Rohtak-124001 (Haryana) E-mail: bijendersagar@rediffmail.com

1. INTRODUCTION

Gum Arabic (GA) or Acacia gum is an edible biopolymer obtained as exudates of mature trees of Acacia Senegal and Acacia seyal which grow principally in the African region of Sahe in Sudan (Williams & Phillips, 2000) .

Chemically, GA is a complex mixture of macromolecules of different size and composition (mainly carbohydrates and proteins) Today, the properties and features of GA have been widely explored and developed and it is being used in a wide range of industrial sectors such as textiles, ceramics, lithography, cosmetics and pharmaceuticals, encapsulation, food etc. Regarding food industry, it is used as a stabilizer, a thickener and/or an emulsifier agent (e.g., soft drink syrup, gummy candies and creams) (Verbeken et al., 2003). In the pharmaceutical industry, GA is used in pharmaceutical preparations and as a carrier of drugs since it is considered a physiologically harmless substance. Additionally, recent studies have highlighted GA antioxidant properties (Hinson et al., 2004), its role in the metabolism of lipids (Evans et al., 1992), its positive results when being used in treatments for several degenerative diseases such as kidney fail, cardiovascular and gastrointestinal (Rehman et al., 2003).

2. CHEMICAL COMPOSITION

The chemical composition of GA is complex and consists of a group of macromolecules characterized by a high proportion of carbohydrates (~97%), which are predominantly composed of D-galactose and L-arabinose units and a low proportion of proteins (<3%) (Islam et al., 1997).

3. PHYSICOCHEMICAL PROPERTIES

GA functional properties are closely related to its structure, which determines, for example, solubility, viscosity, degree of interaction with water and oil in an emulsion, microencapsulation ability, among others.

a) Solubility and viscosity

GA has high water solubility and a relatively low viscosity compared with other gums. Most gums cannot dissolve in water in concentrations above 5% due to their high viscosity. Instead, GA can get dissolved in water in a concentration of 50% w/v, forming a fluid solution with acidic properties (pH ~ 4.5)(Williams et al., 1990).

b) Emulsifying properties

Randall *et al.*,1988, reported that the AGP complex is the main component responsible for GA ability to stabilize emulsions, by the association of the AGP amphiphilic protein component with the surface of oil droplets.

c) Molecular association

Al-Assaf *et al.*, 2007, showed that molecular associations in GA can lead to an increase in molecular weight in the solid state by maturation under controlled heat and humidity. The process does not involves change in the basic structural components and, while the maturation takes place, the level of association increases giving way to AGP with higher molecular weight and protein content.

4. PHARMACOLOGICAL ACTION

Recent reports have confirmed that GA has some biological properties as an antioxidant (Hinson *et al.,* 2004) on the metabolism of lipids (Evans *et al.*,1992), positive contribution in treating kidney, cardiovascular and gastrointestinal diseases (Rehman *et al.,* 2003).

5. APPLICATIONS

GA is being widely used for industrial purposes such as a stabilizer, a thickener, an emulsifier and an encapsulating in the food industry, and to a lesser extent in textiles, ceramics, lithography, cosmetic, and pharmaceutical industry (Verbeken *et al.,* 2003). In the food industry, GA is primarily used in confectionery, bakery, dairy, beverage, and as a microencapsulating agent.

6.CONCLUSION

GA is a natural biopolymer with wide industrial use as a stabilizer, a thickener, an emulsifier and in additive encapsulation not only in food industry but also in textiles, ceramics, lithography, cosmetic and pharmaceutical industry

7. REFERENCES

1. Al-Assaf S., Phillips G.O., Aoki H. & Sasaki Y. Characterization and properties of Acacia senegal (L.) Willd. var. senegal with enhanced properties :Part 1-Controlled maturation of Acacia senegal var. senegal to increase viscoelasticity, produce a hydrogel form and convert a poor into a good emulsifier *Food Hydrocolloids*.200721(3):319-328
2. Evans A.J., Hood R.L., Oakenfull D.G. & Sidhu G.S. Relationship between structure and function of dietary fibre: a comparative study of the effects of three galactomannans on cholesterol metabolism in the rat. *British Journal of Nutrition*.1992;68(1):217-229
3. Hinson J.A., Reid A.B., McCullough S.S. & James L.P. Acetaminophen-induced hepatotoxicity: role of metabolic activation, reactive oxygen/nitrogen species, and mitochondrial permeability transition. *Drug Metabolism Reviews*.2004;36(3-4):805–822
4. Islam A.M., Phillips G.O., Sljivo A., Snowden M.J. & Williams P. A review of recent developments on the regulatory, structural and functional aspects of gum arabic. *Food Hydrocolloids*.1997;11(4)493-505
5. Rehman K.U., Wingertzahn M.A., Teichberg S., Harper R.G. & Wapnir R.A. Gum arabic. (GA) modifies paracellular water and electrolyte transport in the small intestine. *Digestive Diseases and Sciences*.2003;48(4):755–760
6. Randall R.C., Phillips G.O. & Williams P.A. The role of the proteinaceous component on the emulsifying properties of gum arabic. *Food Hydrocolloids*.1988;2(2)131–140
7. Verbeken D.,Dierckx S. & Dewettinck. Exudate gums: Occurrence, production and applications. *Applied Microbiology and Biotechnology*.2003;63(1):10–21
8. Williams P.A., Phillips G.O. In Handbook of Hydrocolloids; Williams, P. A., Phillips, G.O., Eds.; CRC Press: Cambridge, 2000; p 155-168.

METHODS OF PREPARATION, CAHARACTERIZATION AND THERAPEUTIC APPLICATIONS OF NANOPARTICLES

Anusha[1], Minakshi Gupta[1]

Assistant Professor, Shri Baba Mastnath University of Pharmaceutical Sciences And Research, Asthal Bohar, Rohtak (124001)

Nanomedicine is a growing research field dealing with the creation and manipulation of materials at a nanometer scale for the better treatment, diagnosis and imaging of diseases. In cancer medicine, the use of nanoparticles as drug delivery systems has advanced the bioavailability, *in vivo* stability, intestinal absorption, solubility, sustained and targeted delivery, and therapeutic effectiveness of several anticancer agents. Cell-specific targeting can be achieved by attaching drugs to individually designed carriers. Recent developments in nanotechnology have shown that nanoparticles (structures smaller than 100 nm in at least one dimension) have a great potential as drug carriers. Due to their small sizes, the nanostructures exhibit unique physicochemical and biological properties (e.g., an enhanced reactive area as well as an ability to cross cell and tissue barriers) that make them a favourable material for biomedical applications. Nanoparticles could also improve the bioavailability of water-insoluble drugs, carry large payloads, protect the therapeutic agents from physiological barriers, as well as enable the development of novel classes of bioactive macromolecules (e.g., DNA and siRNA). Additionally, the incorporation of imaging contrast agents within nanoparticles can allow us to visualize the site of drug delivery or monitor the in vivo effi cacy of the therapeutic agent. Thus far, over two-dozen nanotechnology products have been approved by the US Food and Drug Administration (FDA) for clinical use, and many are under clinic and preclinic development. Interestingly, the majority of these clinically approved, first-generation nanotechnology products are comprised of liposomal drugs and polymer–drug conjugates, which are relatively simple and generally lack active targeting or controlled drug release components. Nanoparticles are defined as particulate dispersions or solid particles with a size in the range of 10-1000nm. The drug is dissolved, entrapped, encapsulated or attached to a nanoparticle matrix. Depending upon the method of preparation, nanoparticles, nanospheres or nanocapsules can be obtained. Nanocapsules are systems in which the drug is confined to a cavity surrounded by a unique polymer membrane, while nanospheres are matrix systems in which the drug is physically and uniformly dispersed. In recent years, biodegradable polymeric nanoparticles, particularly those coated with hydrophilic polymer such as poly (ethylene glycol) (PEG) known as long-circulating particles, have been used as potential drug delivery devices because of their ability to circulate for a prolonged period time target a particular organ, as carriers of DNA in gene therapy, and their ability to deliver proteins, peptides and genes.

NANOPARTICLE CLASSIFICATION

CATEGORY	EXAMPLES
Nanotubes	carbon, (fullerenes)
Nanowires	metals, semiconductors, oxides, sulfides, nitrides
Nanocrystals	quantum dots insulators, semiconductors, metals, magnetic materials
Other nanoparticles	ceramic oxides, metals
Nanobots	biochip,nubots

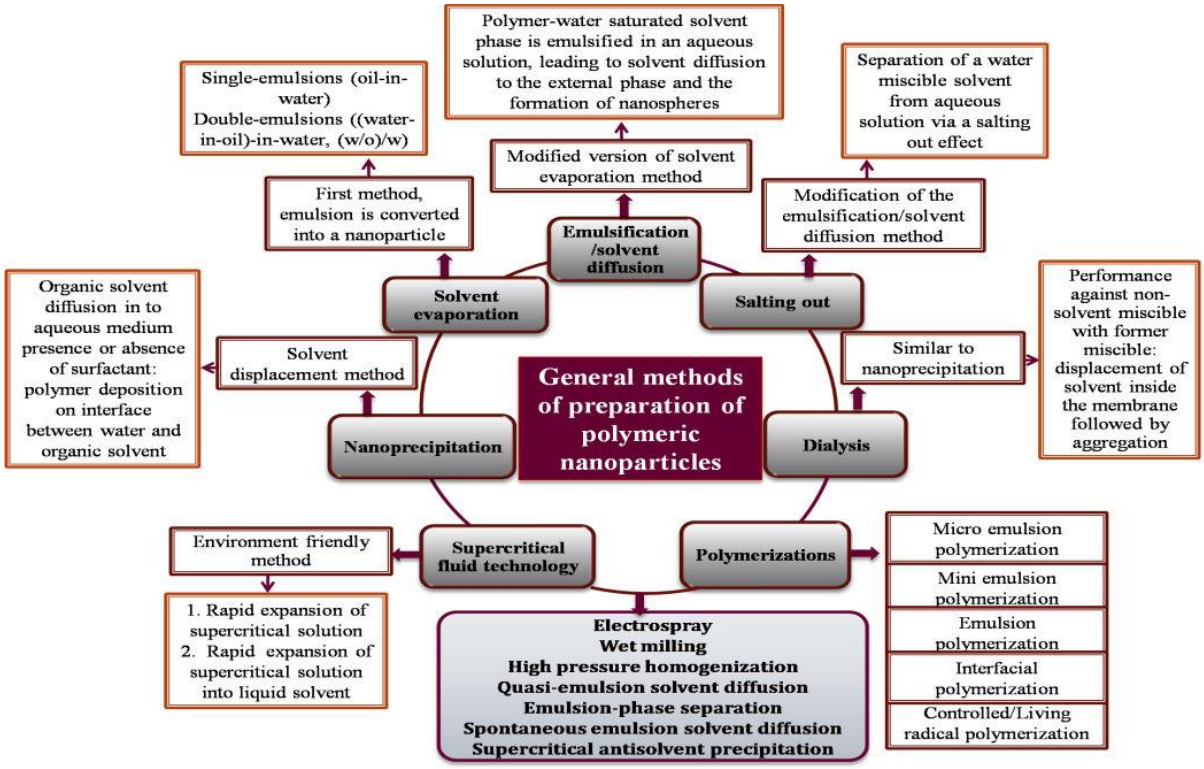


Figure 1. General methods of preparation of polymeric nanoparticles and their principle involved in the mechanisms

ADVANTAGES:

The advantages of using nanoparticles as a drug delivery system include the following:

1. Particle size and surface characteristics of nanoparticles can be easily manipulated to achieve both passive and active drug targeting after parenteral administration.
2. They control and sustain release of the drug during the transportation and at the site of localization, altering organ distribution of the drug and subsequent clearance of the drug so as to achieve increase in drug therapeutic efficacy and reduction in side effects.
3. Controlled release and particle degradation characteristics can be readily modulated by the choice of matrix constituents. Drug loading is relatively high and drugs can be incorporated into the systems without any chemical reaction; this is an important factor for preserving the drug activity.
4. Site-specific targeting can be achieved by attaching targeting ligands to surface of particles or use of magnetic guidance.
5. The system can be used for various routes of administration including oral, nasal, parenteral, intra-ocular etc. In spite of these advantages, nanoparticles do have limitations. For example, their small size and large surface area can lead to particleparticle aggregation, making physical handling of nanoparticles difficult in liquid and dry forms. In addition, small particles size and large surface area readily result in limited drug loading and burst release. These practical problems have to be overcome before nanoparticles can be used clinically or made commercially available. The present review details the latest development of nanoparticulate drug delivery systems, surface modification issues, drug loading strategies, release control and potential applications of nanoparticles.

Table 1: Various parameters and characterization methods for nanoparticles

Parameters	**Characterization methods**
Particle size and distribution	Photon correlation spectroscopy(PCS)

	Laser defractometry Transmission electron microscopy Scanning electron microscopy Atomic force microscopy
Surface hydrophobicity	Water contact angle measurement Rose Bengal(dye) binding X-ray photoelectron spectroscopy
Charge determination	Laser Doppler Anemometry Zeta potentiometer
Nanoparticle dispersion stability	Critical flocculation temperature(CFT)
Chemical analysis of surface	Static secondary ion mass spectrometry Sorptometer
Carrier-drug interaction	Diffential scanning calorimetry
Drug stability	Bioassay of drug extacted from Nanoparticles Chemical analysis of drug
Release profile	In vitro release characteristics under physiologic and sink conditions

Table 2: Therapeutic applications of nanoparticles

Material	**Application**	**Purpose**
Alginate nanoparticles, poly (D, L-lactic acid) nanoparticles	Oligonucleotides delivery	Enhanced delivery of oligonucleotides
Polyesters with adsorbed poly ethylene glycols or pluronics	Prolonged systemic circulation	Prolonged systemic effect, avoid by the uptake of reticuloendothelial system
Poly (alkyl cyanoacrylate) nanoparticles with anticancer agents, oligonucleotides	Cancer therapy	Targeting, reducing toxicity, enhanced uptake of antitumour agents, improved *in vitro* and *in vivo* stability
DNA-gelatine nanoparticles, DNA-chitosan nanoparticles, PDNA-poly (D, L-lactide-co-glycolide) nanoparticles	DNA delivery	Enhanced delivery and significantly higher expression levels
Poly (methyl methacrylate) nanoparticles with proteins and therapeutic agents	Peroral absorption	Enhanced bioavailability, protection from GIT enzymes
Poly (alkyl cyanoacrylate) polyesters nanoparticles with anti-parasitic or antiviral agents	Intracellular targeting	Targeting reticuloendothelial intercellular infections
Poly (alkyl cyanoacrylate) nanoparticles with steroids, anti-inflammatory agents, anti-bacterial agents for glaucoma	Ocular delivery	Improved retention of drug/reduced wash-out

Poly (alkyl cyanoacrylate) nanoparticles with peptides Poly (alkyl cyanoacrylate) nanoparticles Nanoparticles with adsorbed enzymes	Other applications	Crosses blood‐brain barrier Improved absorption and permeation for transdermal application Enzyme immunoassays

Some of the examples of nanoparticles which are being used currently are:

- The Magnetic Resonance Imagining scan can be enhanced by Iron oxide nanoparticle.
- Breakdown of volatile organic compounds in air can be catalyst by gold embedded nanoparticle in porous manganese oxide.
- Coating the nanoparticles with protein is used as drug deliver to damage regions of arteries in cardiovascular disease treatment.
- Cancer cell can be removed before establishing new tumours from blood cell using Magnetic nanoparticles.
- Palladium nanoparticles are used for detection of hydrogen.
- Presences of cancer cell are located by Quantum Dots (crystalline nanoparticles).
- NOMFET (Nanoparticle Organic Memory Field-Effect Transistor) Gold nanoparticle are combined with organic molecules used for delivery of chemotherapy drugs directly into tumour cells.
- For the removal of pollutant like carbon tetrachloride from water Iron nanoparticles can be used.
- To increase power of battery and reducing it recharge time anodes of lithium-ion batteries are coated with Silicon nanoparticles.

Table 3: Marketed formulations of nanoparticles

Trade name	Indication	Composition	Administration	Company
Pegasys	Hepatitis B, Hepatitis C	PEG–a-interferon 2a	Subcutaneous	Nektar, Hoffmann-La Roche
Tricor	Anti-hyperlipidemic	Nanocrystalline fenofibrate	Oral	Elan, Abbott
Estrasorb	Menopausal therapy	Micellular estradiol	Topical	Novavax
Emend	Antiemetic	Nanocrystalline aprepitant	Oral	Elan, Merck
Rapamune	Pharmaceuticals Immunosuppressant	Nanocrystalline sirolimus	Oral	Elan, Wyeth
Berna Biotech	Hepatitis A	Liposomal IRIV vaccine	Intramuscular	Berna Biotech
Abelect	Fungal infection	Liposomal amphotericin B	Intravenous	Enzon
Renagel	End-stage renal disease	Poly(allylamine hydrochloride)	Oral	Genzyme

CONCLUSION:

Nanotechnology is an exciting novel field with hopes for improvements in wide variety of uses in drug delivery in pharmaceutical research. Polymeric nanoparticles offer a new avenue to achieve drug delivery and drug targeting with newly discovered disease site‐specific drugs and existing poorly soluble drugs. The present pharmaceutics is often characterized by poor bio-availability which far too often results in higher patient costs and inefficient treatment but also, more importantly, increased risks of toxicity or even death. Nanotechnology focuses on the very small and it is uniquely suited to creating systems that can better deliver drugs to

tiny areas within the body. The payoff for doctors and patients from nanotechnology-enabled drug delivery should be lower drug toxicity, reduced cost of treatments, improved bioavailability and an extension of the economic life of proprietary drugs. Nanoparticles have been showed great promise for the development of drug delivery system.

REFERENCE:

1. Allemann E, Gurny R, Doelker E. Drug-Loaded Nanoparticles: Preparation Methods and Drug Targeting Issues. European Journal of Pharmaceutics and Biopharmaceutics 1993;39(5) 173-191.
2. Kawashima Y. Nanoparticulate Systems for Improved Drug Delivery. Advanced Drug Delivery Reviews 2001;47(1) 1-2.
3. Soppimath KS, Aminabhavi TM, Kulkarni AR, Rudzinski WE. Biodegradable Polymeric Nanoparticles as Drug Delivery Devices. Journal of Controlled Release 2001;70(1-2) 1-20.
4. Panyam J, Labhasetwar V. Biodegradable Nanoparticle from Drug and Gene Delivery to Cells and Tissue. Advanced Drug Delivery Reviews 2003;55(3) 329-347.
5. Brigger I, Dubernet C, Couvreur P. Nanoparticles in Cancer Therapy and Diagnosis. Advanced Drug Delivery Reviews 2002;54(5) 631-651.
6. Cohen H, Levy RJ, Gao J, Fishbein I, Kousaev V, Sosnowski S, Slomkowski S, Golomb G. Sustained Delivery and Expression of DNA Encapsulated in Polymeric Nanoparticles. Gene Therapy 2000;(22) 71896-1905.
7. Cao X, Lai S, Lee LJ. Design of a Self-Regulated Drug Delivery Device. Biomedical Micro devices 2001;3(2)109-118.
8. Kasagana VN, Karumuri SS, Thirumal. M. Recent Advances in Smart Drug Delivery Systems. International Journal of Pharmacy and Biotechnology 2011;1(3) 201-207.
9. Kreuter, J. Nanoparticles and Nanocapsules-New Dosage Forms in the Nanometer Range. Pharmaceutica Acta Helvetiae 1978;53(2) 33-39.
10. Kreuter J. Nanoparticles. In: Swarbrick J, Boylan JC. (ed.) Encyclopedia of Pharmaceutical Technology. Marcel Dekker Inc: New York, NY, 1994. p165-190.
11. Birrenbach G, Speiser PP. Polymerized Micelles and their use as Adjuvants in Immunology. Journal of Pharmaceutical Sciences 1976;65(12) 1763-1766.
12. Rollot JM, Couvreur P, Roblot-Treubel L, Puisieux F. Physicochemical and Morphological Characterization of Polyisobutyl Cyanoacrylate Nanocapsules. Journal of Pharmaceutical Sciences 1986;75(4) 361-364.
13. Legrand P, Barratt G, Mosqueira V, Fessi H, Devissaguet JP. Polymeric Nanocapsules as Drug Delivery Systems. A review. STP Pharma Sciences 1999;9(5) 411-418.
14. Barratt GM. Therapeutic Applications of Colloidal Drug Carriers. Pharmaceutical Science and Technology Today 2000;3(5) 163-171.
15. Fresta M, Cavallaro G, Giammona G, Wehrli E, Puglisi G. Preparation and Characterization of Polyethyl-2-Cyanoacrylate Nanocapsules containing Antiepileptic Drugs. Biomaterials 1996;17(8) 751-758.
16. Schroeder U, Sommerfeld P, Ulrich S, Sabel BA. Nanoparticle Technology for Delivery of Drugs across the Blood-Brain Barrier. Journal of Pharmaceutical Sciences 1998;87(11) 1305-1307

LYOPHILIZATION: A NOVEL APPROACH TO STABLE FORMULATIONS

Jyoti dahiya[1], Priti Mehndiratta[1]

[1]Department of Pharmaceutical Sciences, Baba Mastnath University, Rohtak(124001), India. Email: jyotidahiya.2710@gmail.com, pritimehndiratta88@gmail.com

INTRODUCTION

Lyophilization has been used in a number of applications for many years, regularly in the food and pharmaceutical industries. This process is also known as freeze drying process. On the other hand there are, many other uses for the process including heat-sensitive sample preparation, plant material research, the stabilization of living materials such as microbial cultures, long term storage of HPLC samples, preservation of whole animal specimens for museum display and restoration of books and other items damaged by water[1-3]. Lyophilization is a process which extracts the water from foods and other products so that the foods or products remain stable and are easier to store at room temperature (ambiant air temperature). Lyophilization is based on a simple principle of physics named as sublimation. Sublimation is defined as a process by which solid state is directly converted into vapour state without passing through an intermediate liquid state. However, in this process, the material does not go through the liquid phase, and it allows the preparation of a stable product that is easy to use and artistic in manifestation. Various steps involved in lyophilization process are as following[1,2,3-7]:

1. Initial step is to freeze the food so that the water in the food become ice;
2. Secondly under a vacuum, sublimating the ice directly into water vapour;
3. Extracting off the water vapour;
4. Once the sublimation is completed, the foods are freeze-dried and can be removed from the machine.

TIME HONOURED LYOPHILIZATION TECHNOLOGY

Traditional lyophilization is a complex process that requires an alert evaluation of product, equipment, and processing techniques. For nearly 30 years, lyophilization has been used to stabilize many types of chemical components. In their liquid form, many such biochemicals and chemical reagents are unstable, temperature sensitive, and chemically reactive with one another. Because of these qualities, the chemicals may have a very short shelf life, may need to be refrigerated, or may degrade unless stabilized[7-11]. When performed accurately, the process of lyophilization solves these problems by putting reagents into a state of suspended activity. The process gives a product excellent solubility quality, allowing for hasty reconstitution. Heat- and moisture-sensitive compounds retain their feasibility[11-16].

PRINCIPLES OF FREEZE DRYING

Three stages are basically involved in freeze drying: freezing, primary drying and secondary drying.[3,17,18]

Freezing: The freezing process is realized by contacting a liquid sample with or placing it in a cold bath. The frozen sample is then placed in a freeze dryer to remove the frozen solvent by sublimation. During the freeze drying process, the frozen sample should be kept below the glass transition temperature or melting point and the frozen solvent is removed under vacuum.[17,18]

Primary drying: Several factors can affect the ability to freeze dry a frozen suspension. After freezing the product, conditions must be established in which ice can be removed from the frozen product via sublimation, resulting in a dry, structurally integral product. This requires very careful control of the two parameters, temperature and pressure, involved in the

freeze drying system. The rate of sublimation of ice from a frozen product depends upon the difference in vapor pressure of the product compared to the vapor pressure of the ice collector. No matter what type of freeze drying system is used, conditions must be created to encourage the free flow of water molecules from the product[19-24].

Secondary drying: After primary freeze drying is complete, and all ice has sublimed, bound moisture is still present in the product. The product appears dry, but the remaining moisture content may be as high as 7-8%[25-27]. Sustained drying is necessary at the warmer temperature to reduce the remaining moisture content to most favorable values. This process is called isothermal desorption as the bound water is desorbed from the product. Secondary drying is usually carried out for approximately 1/3 to 1/2 the time required for primary drying[28].

LYOPHILIZATION EQUIPMENT

A lyophilizer consists of a vacuum chamber that contains product shelves capable of cooling and heating containers and their contents. A vacuum pump, a refrigeration unit, and associated controls are connected to the vacuum chamber. Chemicals are generally placed in containers such as glass vials that are placed on the shelves within the vacuum chamber. Cooling elements within the shelves freeze the product. Once the product is frozen, the vacuum pump evacuates the chamber and the product is heated. Heat is transferred by thermal conduction from the shelf, through the vial, and ultimately into the product[11,16,29-32].

PROCEDURE OF LYOPHILIZATION

Sample preparation (day before)

1. Samples must be fully frozen (preferably overnight), ideally on one or more of the stainless steel freeze dryer trays (so that the tray is also cold and samples don't melt as easily on the way to the freeze dryer).
2. Samples must be open to the air (for vapor to escape) but mostly covered (to prevent cross-contamination and particulates entering the pump). You can accomplish this by either poking a hole in the sample container or removing the lid and covering the container with a lab wipe or paper towel, rubber-banded on.

Setting up freeze dryer (15 minutes)

1. If chamber is assembled but freeze dryer is turned off, remove transparent cover of chamber and set it upside-down on the bench next to the freeze-dryer. If necessary, lift out the metal tray frame and set it aside. Lift the chamber off its base and set it on the lid (to keep the greased gaskets from getting crud on them).
2. Remove metal elbow pipe from freezer chamber outlet and plug the hole with a rubber stopper.
3. Make sure drain cock is closed.
4. Open ballast (switch on front of pump).
5. Turn on vacuum pump and run for 15 minutes to warm up the pump and purge any water from the pump oil.
6. In the meantime, also turn on the freezer to cool down the condensation chamber. The temperature should be down to about –40° C before you start.
7. Turn off vacuum pump. Remove rubber stopper from outlet pipe and replace with metal elbow.
8. Move immediately to setting up samples.

Setting up manifold and samples (15 minutes)

1. Make sure the gaskets on the top and bottom of the chamber are free of dirt and particles and lightly coated with silicone vacuum grease. (Grease should only be applied occasionally).
2. Place the manifold on the base, making sure that it is centered on the steel and not hanging over the edge.

3. Check that all ports on the manifold are set to the VENT position. (This sounds backward but "VENT" and "VAC" in this case refer to the vessels that can be attached to the ports.)
4. Put trays onto the tray rack and lower the rack into the freeze dryer (this is easier than lowering the chamber over the tray rack). Work quickly and don't let the samples melt.
5. Turn on vacuum pump. Immediately observe the top gasket through the clear manifold cover. You should see the black line (where the gasket meets up with the cover) getting wider as the chamber is evacuated and the cover presses down on the chamber. If the line is not getting wider, you have a leak. Find it and fix it right away!
6. Let the pump run with the ballast open for 10-15 minutes.
7. Close the ballast.
8. Run the freeze dryer overnight or until the vacuum gauge reads about 10-1 mbar.

Note: if the samples are very wet, or if there are a lot* of them, you may need to freeze dry them for several days. It is much more effective to defrost the chamber once daily rather than leaving the dryer to run on for multiple days without defrosting.

The efficiency of water removal decreases once the inside of the condenser chamber becomes coated with ice (after 12-24 hours).

Shutting down freeze dryer (20 minutes)

1. Turn off vacuum pump and freezer.
2. Release vacuum slowly by turning one of the red caps on the manifold ports about ¼ turn until you hear it hiss. If you release the vacuum too fast, things will fly around in there.
3. Remove the cover and set it upside down next to the dryer. Lift out the tray rack and set it aside. Lift off the manifold chamber and set it on its upside-down lid.
4. Remove elbow pipe and replace with rubber stopper.
5. Open ballast.
6. Turn pump on and run for 15 minutes.
7. In the meantime, defrost the condenser chamber. Fill the condenser chamber about halfway with hot tap water. Allow the ice to melt, then dip out the water using a plastic beaker (the drain is completely plugged these days)[28,30,31].

ADVANTAGES

Lyophilization has many advantages compared to other drying and preserving techniques[5,9,30-32].

1. Lyophilization maintains food/ biochemical and chemical reagent quality because they remains at a temperature that is below the freezing-point during the process of sublimation; The use of lyophilization is particularly important when processing lactic bacteria, because these products are easily affected by heat.
2. Food/biochemicals and chemical reagents which are lyophilized can usually be stored without refrigeration, which results in a significant reduction of storage and transportation costs.
3. Lyophilization greatly reduces weight, and this makes the products easier to transport. For example, many foods contain as much as 90% water. These foods are 10 times lighter after lyophilization.
4. Because they are porous, most freeze-dried products can be easily rehydrated. Lyophilization does not significantly reduce volume, therefore water quickly regains its place in the molecular structure of the food/ biochemicals and chemical reagents.

CONCLUSION

The processes of freezing and freeze drying have been widely used in materials synthesis and applications. The low surface tension involved during the freeze drying process can maintain the pore structure (particularly mesopores and micropores) which would otherwise collapse due to the high capillary force or surface tension during a normal drying process. For this reason, freeze drying has been employed frequently in the preparation of porous materials.

The distinct advantage of the freeze drying method may be the use in water-based systems which are essential for biological and environmental applications. Freeze drying processes may avoid the use of toxic organic solvents and the low temperature of the process helps maintain the activity of bio macromolecules (such as proteins, enzymes) and pharmaceuticals.

REFERENCES

1. Barbaree, J.M. and A. Sanchez. 1982. Cross-contamination during lyophilization. *Cryobiology* 19:443-447.
2. Barbaree, J.M., A. Sanchez and G.N. Sanden. 1985. Problems in freeze-drying: I. Stability in glass-sealed rubber stoppered vials. *Developments in Industrial Microbiology* 26:397-405.
3. Barbaree, J.M., A. Sanchez and G.N. Sanden. 1985. Problems in freeze-drying: II. Cross-contamination during lyophilization. *Developments in Industrial Microbiology* 26:407-409.
4. Flink, J.M. and Knudsen, H. 1983. *An Introduction to Freeze Drying*. Strandberg Bogtryk/Offset, Denmark.
5. Flosdorf, E.W. 1949. *Freeze-Drying*. Reinhold Publishing Corporation, New York.
6. Greiff, D. 1971. Protein structure and freeze-drying: the effects of residual moisture and gases. *Cryobiology* 8:145- 152.
7. Greiff, D. and W.A. Rightsel. 1965. An accelerated storage test for predicting the stability of suspensions of measles virus dried by sublimation in vacuum. *Journal of Immunology* 94:395-400.
8. Greaves, R.I.N., J. Nagington, and T.D. Kellaway. 1963. Preservation of living cells by freezing and by drying. *Federation Proceedings* 22:90-93.
9. Harris, R.J.C., Ed. 1954. *Biological Applications of Freezing and Drying*. Academic Press, New York.
10. Heckly, R.J. 1961. Preservation of bacteria by lyophilization. *Advances in Applied Microbiology* 3:1-76.
11. Heckly, R.J. 1985. Principles of preserving bacteria by freeze-drying. *Developments in Industrial Microbiology* 26:379-395.
12. Jennings, Thomas. 2002. *Lyophilization: Introduction and Basic Principles*. CRC Press, Florida.
13. King, C.J. 1971. *Freeze-Drying of Foods*. CRC Press, Cleveland.
14. May, M.C. E. Grim, R.M. Wheeler and J. West. 1982. Determination of residual moisture in freeze-dried viral vaccines: Karl Fischer, gravimetric, and thermogravimetric methodologies. *Journal of Biological Standardization* 10:249-259.
15. Mellor, J.D. 1978. *Fundamentals of Freeze-Drying*. Academic Press, London.
16. Nail, S.L. 1980. The effect of chamber pressure on heat transfer in the freeze-drying of parental solutions. *Journal of the Parental Drug Association* 34:358-368.
17. Nicholson, A.E. 1977. Predicting stability of lyophilized products. *Developments in Biological Standardization* 36:69-75.
18. Parkes, A.S., and A.U. Smith, Eds. 1960. *Recent Research in Freezing and Freeze-Drying*. Charles C. Thomas, Springfield.
19. Rey, L.R., Ed. 1960. *Traite de Lyophilization*. Hermann, Paris.
20. Rey, L.R., Ed. 1964. *Aspects Theorique et Industriels de la Lyophilisation*. Hermann, Paris.
21. Rowe, T.W.G. 1970. Freeze-drying of biological materials: some physical and engineering aspects. *Current Trends in Cryobiology:* 61-138.
22. Seligman, E.B. and J.F. Farber. 1971. Freeze-drying and residual moisture. *Cryobiology* 8:138-144.
23. H. Fessi, J.P.Devissaguet, F. Puisieux, Process for the preparation of dispersible colloidal systems of a substance in the form of nanoparticles. US patent 5, 118, 528, 2 Jun 1992.
24. S. Galindo-Rodriguez, E. Allémann, H. Fessi, E. Doelker, Physicochemical parameters associated with nanoparticle formation in the salting-out, emulsification-diffusion, and nanoprecipitation methods, Pharm. Res. 21 (2004) 1428–1439.
25. R. Oppenheim, Solid colloidal drug delivery systems: nanoparticles, Int. J. Pharm. 8 (1981) 217–234.
26. M.J. Alonso, Nanoparticulate drug carrier technology, in: S. Cohen, H. Bernstein (Eds.), Microparticulate Systems for the Delivery of Proteins and Vaccines, Marcel Dekker, New York, 1996, pp. 203–242.
27. I. Brigger, C. Dubernet, P. Couvreur, Nanoparticles in cancer therapy and diagnosis, Adv. Drug Deliv. Rev. 54 (2002) 631–651.
28. G. Storm, Sheila O. Belliot, T. Daemen, D. Lasic, Surface modification of nanoparticles to oppose uptake by the mononuclear phagocyte system, Adv. Drug Deliv. Rev. 17 (1) (1995),31–48.
29. J. Panyam, V. Labhasetwar, Biodegradable nanoparticles for drug and gene delivery to cells and tissue, Adv. Drug Deliv. Rev. 55 (3) (2003) 329–347 (24).
30. M.P. Desai, V. Labhasetwar, E.Walter, R.J. Levy, G.L. Amidon, The mechanism of uptake of biodegradable microparticles in Caco-2 cells is size dependent, Pharm. Res. 14 (1997).

31. M.P. Desai, V. Labhasetwar, G.L. Amidon, R.J. Levy, Gastrointestinal uptake of biodegradable microparticles: effect of particle size, Pharm. Res. 13 (1996) 1838–1845.
32. S.R. Schaffazick, A.R. Pohlmann, T. Dalla-Costa, S.S. Guterres, Freeze-drying polymeric colloidal suspensions: nanacapsules, nanospheres and nanodispersion. A comparative study, Eur. J. Pharm. Biopharm. 56 (2003) 501–506.

HERB-DRUG INTERACTIONS

Shilpi Arora, Upma,

Baba Mastnath University, rohtak, Harayana

ABSTRACT

Millions of people today use herbal therapies along with prescription and nonprescription medications. Although considered natural, many of these herbal therapies can interact with other medications, causing either potentially dangerous
side effects and / or reduced benefits from the medications. Currently, there is very little information published on herb-drug interactions whilst the use of herbs is progressively growing across the world. Ayurvedicsyrup shankhapushpi. Anthranoid-containing plants (including senna [*Cassia senna*] and cascara[*Rhamnus purshiana*]) and soluble fibres (including guar gum and psyllium) can decrease the absorption of drugs. Many reports of herb-drug interactions are sketchy and lack laboratory analysis of suspect preparations. Health-care.

INTRODUCTION

Millions of people today use herbal therapies along with prescription and nonprescription medications. Although considered natural, many of these herbal therapies can interact with other medications, causing either potentially dangerous
side effects and / or reduced benefits from the medications. Currently, there is very little information published on herb-drug interactions whilst the use of herbs is progressively growing across the world.

Many medicinal herbs and pharmaceutical drugs are therapeutic at one dose and toxic at another.Interactions between herbs and drugs may increase or decrease the pharmacological or toxicological effects of either component. Synergistic therapeutic effects may complicate the dosing of long-term medications—eg, herbs traditionally used to decrease glucose concentrations in diabetes1 could theoretically precipitate hypoglycemia if taken in combination with conventional drugs.

DETERMINATION OF HERBAL-DRUG INTERACTIONS

Recent research estimates 50% of adult Americans use at least one prescription medication and 7% of adult Americans take 5 or more prescription drugs.

Among prescription drug users, 16% also take herbal supplements however the prevalence of clinically significant interactions between herbals and medications is unknown. A factor that may account for the lack of data on the true prevalence of herbaldrugn interactions is that information needed to determine whether an interaction has occurred is often unavailable owing to the following:

1. A lack of information concerning the “contents” of the herbal product
2. Incomplete or inaccurate product information
3. Multiple ingredients

Additionally, patients may not inform health care providers of suspected interactions, or they do notattribute the reaction to the natural product.

MECHANISMS

Interactions between herbals and medications can be caused by either Pharmacodynamic or pharmacokinetic mechanisms.

Pharmacokinetic interactions:

Here herbal changes the absorption, distribution, metabolism, protein binding, or excretion of a drug that results in altered levels of the drug or its metabolites. Most of the current evidence of pharmacokinetic drug interactions involves metabolizing enzymes and drug transporters. Although drug interactions can involve enzymes such as glutathione S-transferases and uridine diphosphoglucuronyl transfereases (UGTs), most herbal-drug

interactions are related to oxidative metabolism by the cytochrome P-450 system (CYP) or by the effect of a herbal on the efflux drug transporter P-glycoprotein.

Pharmacodynamic interactions

Pharmacodynamic interactions are related to thepharmacologic activity of the interacting agents and can affect organ systems, receptor sites, or enzymes.

A Pharmacodynamic interaction may occur when herbals that possess antiplatelet activity are administered with antiplatelet/anticoagulant drugs, thus increasing the risk for bleeding. Other examples are when herbals that depress the central nervous system (CNS), such as kava, are administered with CNS depressant drugs or when herbals that may lower blood glucose are given with antidiabetic drugs. An example of an antagonistic interaction is when annherbal with high caffeine content, such as guarana, is administered with a sedative-hypnotic. In addition, herbals with the potential to cause organ toxicity may cause further risk of toxicity when drugs with similar toxicity are administered concurrently, such as when the hepatotoxic herbal comfrey is given with large andprolonged doses of acetaminophen.

Evaluation of the probability of herbal–drug interactions

- Adequate patient history (age, sex, relevant medical conditions) is reported
- Concurrent diseases, conditions, or other medications associated with adverse event (including dosing)
- Concomitant medications are documented (including dosing)
- Description of interactors is adequate
- Obvious alternate explanations have been excluded
- Chronology is complete
- Time sequence of drug administration to adverse event is reasonable
- Adverse event is adequately described
- Event ceases upon stopping drug
- Event recurs upon rechalleng

ROLE OF PHARMACIST IN PREVENTING HERB DRUG INTERACTION-:

Pharmacist can play a vital role in preventing drug herb interaction to occur by appropriately dispensing medicine and taking due care of patient's history and medication profile. In order to ensure that the drugs that he is dispensing to the patient are safe and will not cause any interaction, he should ask following questions:

· Are you taking an herbal product, herbal supplement or other "natural remedy?"
· If so, are you taking any prescription or nonprescription medications for the same purpose as the herbal product?
· Have you used this herbal product before?
· Are you allergic to any plant products?
· Are you pregnant or breast-feeding?

Make sure your pharmacist and your doctor(s) know about every drug you are taking, including prescription and nonprescription drugs, herbal products, and any dietary supplements, including vitamins and minerals; Only take medication that has been specifically prescribed for you by your physician; Medication must be taken properly to ensure its safety and effectiveness; Unless otherwise instructed, take medicine on an empty stomach to achieve a faster onset of action; When taking medicine with food or around a meal time is not recommended, take medicine one hour before meal/food or two hours after meals or eating food; Take your medicine with a full glass of water. Avoid concurrent use of alcohol with medicine. Avoid consuming excessive quantities of chocolate and beverages containing caffeine coffee,tea,colas and If you have any questions or concerns about your medicine or you believe you are having an adverse drug reaction or drug interaction, consult

your pharmacist or physician immediately. If there is a problem, your pharmacist can contact your physician, who can prescribe other medication to avoid the risk of drug related Problems.

REFERENCES

1. Mathur A. Who owns traditional knowledge? Working Paper No. 96, Indian Council for Research on International Economic Relations; 2003. p. 1-33.
2. Agrawal OP, Raju PS. *In*: Abdin MZ, Abroi YP, editors. Global market of herbal products: Opportunities for India. Traditional system of medicine. Narosa Publishing House: New Delhi, India; 2006. p. 5-10.
3. Wakdikar S. Global health care challenge: Indian experiences and new prescriptions. Electronic J Biotechnol 2004;7:214-20.
4. Eisenberg DM, Davis RB, Ettner SL, Appel S, Wilkey S, Van Rompay M, *et al*.
5. Trends in alternative medicine use in the United States, 1990-1997: Results of a follow-up national survey. JAMA 1998;28:1569-75.
6. Ramakrishnappa K. Biodiversity and the Ecosystem Approach in Agriculture, Forestry and Fisheries. New York, USA: 2003. Impact of Cultivation and Gathering of Medicinal Plants on Biodiversity: Case studies from India. United Nations Publications: p. 210-5.
7. WHO Traditional Medicine Strategy 2002-2005, Document WHO/EDM/TRM/2002.1, WHO: Geneva; 2002. p. 11.
8. Fugh-Berman A. Herb-drug interactions. Lancet

MICROPARTICULATE DRUG DELIVERY SYSTEM: A REVIEW

PritiMehndiratta, JyotiDahiya

Department of Pharmaceutical Sciences & Research, BMU

INTRODUCTION

Oral drug administration is by far the most preferable route for taking medications. However, their short circulating half-life and restricted absorption via a defined segment of intestine limits the therapeutic potential of many drugs. Such a pharmacokinetic limitation leads in many cases to frequent dosing of medication to achieve therapeutic effect. This results in pill burden and consequently, patient complains. Rational approach to enhance bioavailability and improve pharmacokinetic and pharmacodynamic profile is to release the drug in a controlled manner and site specific manner.

Microparticles are a type of drug delivery systems where the particle size ranges from one micron (one thousandth of a mm) to few mm. This microencapsulation technology allows protection of drug from the environment, stabilization of sensitive drug substances, elimination of incompatibilities, or masking of unpleasant taste. Hence, they play an important role as drug delivery systems aiming at improved bioavailability of conventional drugs and minimizing side effects.

MORPHOLOGY OF MICROPARITCLE

Microencapsulation is a technology used to entrap solids, liquids, or gases inside a polymeric matrix or shell. Microparticles are particulate dispersions or solid particles. Two general micromorphologies of microparticles can be distinguished- microcapsules and microspheres.

Microcapsuleis a system in which drug containing core is completely surrounded by a polymer shell. The core can be solid, liquid or gas; the shell is a continuous, porous or non-porous polymeric layer.

Microsphere is a system in which the drug substance is either homogenously dissolved or dispersed in a polymeric matrix. Microspheres show different release properties compared to true microcapsuless

APPLICATIONS

Microparticulate drug delivery offers several applications for drugs having poor bioavailability. A number of pharmaceutical encapsulated products are currently on the market, Such as aspirin, theophylline and its derivatives, vitamins, antihypertensive, potassium chloride, progesterone and contraceptive hormone combinations.

RELEASE MECHANISM

Different release mechanisms of encapsulated material provide controlled, sustained or targeted release of core material. Generally there are three different mechanisms by which the core material is released from a microcapsule-mechanical rupture of the capsule wall dissolution or melting of the wall and diffusion through the wall less common release mechanisms include ablation (slow erosion of shell) and biodegradation. Drug release from the microsphere occurs by general mechanism including diffusion, polymer degradation, and hydrolysis/erosion.

CARRIERS USED IN PREPARATION OF MICROPARTICLES

Polymers remain the most versatile class of biomaterials being extensively applied in medicine and biotechnology, as well as in food and cosmetic industries. A broad range of polymers can be used to form microcapsules. The polymer is judiciously combined with the drug/other active ingredient in such a way that the active agent is released from the material in a predetermined fashion and released the drug at a constant rate for a desired time period.

Hence, encapsulating the drug in a polymer matrix provides ideal pharmacokinetic profile for drugs, where the drug concentration reaches therapeutic levels without exceeding the max tolerable dose and maintains the concentrations for extended periods of time till the desired therapeutic effect is reached. The five key advantages that polymeric drug delivery products can offer are localized delivery of drug, sustained delivery of drugs, stabilization of drugs, release rate which is less dependent of the drug properties and steadier release rate with time.

TECHNIQUES OF MICROENCAPSULATION:

A variety of techniques are employed for the entrapment of solids or liquids within polymer coatings or matrices, the choice of preparation method essentially depends on the raw material intended to be used and on the solubility characteristics of the active compound to be associated with the particles. The preparation method can be broadly divided into 2 categories: Chemical methods and physical methods. Chemical Methods This method uses monomers/prepolymers as starting materials. These methods involve chemical reactions along with microsphere formation. These include suspension polymerization, emulsion polymerization, dispersion and interfacial methods. Among them emulsion polymerization method is widely used in drug delivery

FACTORS INFLUENCING ENCAPSULATION EFFICIENCY

1. Solubility of polymer in the organic solvent
2. Solubility of organic solvent in water
3. Concentration of the polymer
4. Ratio of dispersed phase to continuous phase (DP/ CP ratio)
5. Rate of solvent removal
6. Interaction between drug and polymer
7. Solubility of drug in continuous phase
8. Molecular weight of the polymer

EVALUATION OF MICROPARTICLES:

1. Microsphere recovery/yield:
2. Drug Entrapment Efficiency:
3. Surface Morphology:
4. Particle Size Analysis:
5. In vitro Release Studies:
6. Differential Scanning Calorimetry (DSC) Analysis:

REFERENCE:

1. Genç L, Demirel M, Güler E and MegazyN,Microencapsulation of ketorolac tromethamineby means of a coacervation phase separationtechnique induced by the addition of nonsolvent.J. Microencap, 1998; 15: 45–54.
2. Ueda M, Nakamura Y, Makita H andKawashima Y, Preparation of microcapsulesmasking the bitter taste of enoxacin by usingone continuous process technique ofagglomeration and microencapsulation. J.Microencap., 1993; 10: 461–473.
3. Yu W P, Wong J P, and Chang T M,Preparation of polylactic acid microcapsulescontaining ciprofloxacin. J. Microencap.,1998; 15: 515–523.
4. Tirkkonen S, Turokka L and Paronen P.Microencapsulation of indomethacin bygelatin–acacia complex coacervation in thepresence of surfactants. J. Microencap., 1994;11: 615–626.

ATYPICAL ANTIPSYCHOTICS

Monika Kaushik , Rajiv Kumar Arora

Faculty of Pharm. Sciences, Baba Mastnath University, Asthal Bohar, Rohtak

INTRODUCTION

The **atypical antipsychotics** (**AAP**; also known as **second generation antipsychotics** (**SGAs**)) are a group of antipsychotic drugs used to treat psychiatric conditions. The drugs in this class of antipsychotics act on many receptor types including dopamine and serotonin, but they are be more selective for dopamine receptors.

Atypical antipsychotics are newer antipsychotic agents that have a pharmacological profile different from older or typical antipsychotic drugs. They cause less extrapyramidal side effects compared to the older typical antipsychotic drugs. They are more effective in treatment-resistant patients and have a greater efficacy to treat negative symptoms, compared to the typical antipsychotics.

The second generation antipsychotics are usually the first choice for the treatment of schizophrenia. although they may not be officially approved for these other uses, they are sometimes used in the treatment of mood and anxiety disorders, such as bipolar, posttraumatic stress and obsessive-compulsive disorders.

List of Second-Generation / Atypical Antipsychotic Agents:

Medications available in this class include :

Risperidone (Risperdal)
Quetiapine (Seroquel)
Olanzapine (Zyprexa)
Ziprasidone (Zeldox)
Paliperidone (Invega)
Aripiprazole (abilify)
Clozapine (clozaril)

Atypical antipsychotics (serotonin-dopamine antagonists) are antagonists of D2 and serotonin 2A receptors, but they can affect many other types of receptors . They are much more efficient in treatment of negative symptoms of schizophrenia in comparison with conventional antipsychotics. Atypical antipsychotics have lower side effects (lower EPS or tardive dyskinesis)

Receptor systems affected by atypical antipsychotics

Risperidone	**D2, 5-HT$_{2A}$**, 5-HT$_7$, a$_1$, a$_2$
Sertindole	**D2**, **5-HT$_{2A}$**, 5-HT$_{2C}$, 5-HT$_6$, 5-HT$_7$, D3, a$_1$
Ziprasidone	**D2**, **5-HT$_{2A}$**, 5-HT$_{1A}$, 5-HT$_{1D}$, 5-HT$_{2C}$, 5-HT$_7$, D3, a$_1$, NRI, SRI
Loxapine	**D2**, **5-HT$_{2A}$**, 5-HT$_6$, 5-HT$_7$, D1, D4, a$_1$, M$_1$, H$_1$, NRI
Zotepine	**D2**, **5-HT$_{2A}$**, 5-HT$_{2C}$, 5-HT$_6$, 5-HT$_7$, D1, D3, D4, a$_1$, H$_1$, NRI
Clozapine	**D2**, **5-HT$_{2A}$**, 5-HT$_{1A}$, 5-HT$_{2C}$, 5-HT$_3$, 5-HT$_6$, 5-HT$_7$, D1, D3, D4, a$_1$, a$_2$, M$_1$, H$_1$
Olanzapine	**D2**, **5-HT$_{2A}$**, 5-HT$_{2C}$, 5-HT$_3$, 5-HT$_6$, D1, D3, D4, D5, a$_1$, M$_{1-5}$, H$_1$
Quetiapine	**D2**, **5-HT$_{2A}$**, 5-HT$_6$, 5-HT$_7$, a$_1$, a$_2$, H$_1$

NRI - norepinephrine reuptake inhibitor, SRI – serotonin reuptake inhibitor

MECHANISMS OF ACTION OF ATYPICAL ANTIPSYCHOTICS:

- D2 receptor blockade of postsynaptic in the mesolimbic pathway to reduce positive symptoms.
- Enhanced dopamine release and 5-HT$_{2A}$ receptor blockade in the mesocortical pathway to reduce negative symptoms.

- Other receptor-binding properties may contribute to efficacy in treating cognitive symptoms, aggressive symptoms and depression in schizophrenia.

MEDICAL USES:

- **Schizophrenia**
- **Schizoaffective disorder** most commonly in conjunction with either an antidepressant (in the case of the depressive subtype) or a mood stabiliser (in the case of the bipolar subtype).
- Bipolar disorder (acute mania and mixed episodes) may be treated with either typical or atypical antipsychotics, although atypical antipsychotics are usually preferred because they tend to have more favourable adverse effect profiles and, according to a recent meta-analysis, they tend to have a lower liability for causing conversion from mania to depression.
- Psychotic depression. In this indication it is a common practice for the psychiatrist to prescribe a combination of an atypical antipsychotic and an antidepressant as this practice is best supported by the evidence.
- Treatment-resistant (and not necessarily psychotic) major depression as an adjunct to standard antidepressant therapy.

They are not recommended for dementia or insomnia unless other treatments have not worked. They are not recommended in children unless other treatments are not effective or unless the child has psychosis.

Side-effects of second generation antipsychotics

- Dry mouth, dizziness, blurred vision and, rarely, seizures.
- Weight gain, diabetes
- movement effects (e.g., tremor, stiffness, agitation)
- Sedation (e.g., sleepiness, low energy)
- Decreased sex drive and function, missed periods, discharge from breasts

Drugs most likely to least likely to have these effects

- clozapine > olanzapine > quetiapine > risperidone > ziprasidone, aripiprazole
- risperidone > olanzapine, quetiapine, ziprasidone, aripiprazole > clozapine
- clozapine, olanzapine and quetiapine > risperidone, ziprasidone, aripiprazole
- risperidone > olanzapine, quetiapine > clozapine, ziprasidone

CONCLUSION

Atypical antipsychotics are a heterogeneous group of compounds with complex, multimodal mechanisms of action. Although they all share certain general characteristics, such as a lower propensity for causing EPS and other adverse effects compared with conventional antipsychotics, they differ from conventional agents by virtue of their actions at multiple receptor sites, including dopaminergic, serotonergic, adrenergic, cholinergic, and histaminic receptors within the CNS. Each atypical antipsychotic possesses its own unique receptor profile, imparting a distinct set of pharmacodynamic and clinical characteristics. It has been shown unequivocally that the dose of atypical antipsychotic affects the degree of receptor binding, which, in turn, profoundly influences both desired and undesired clinical effects, such as sedation, weight gain, and metabolic problems. Patients with schizophrenia or

schizoaffective disorder treated with the appropriate dose of an atypical antipsychotic would be expected to exhibit improvements in positive and negative symptoms, cognition, and mood, with low discontinuation rates due to adverse events. As clinical and investigative experience with the atypical agents grows, we are beginning to discover subtle yet clinically meaningful differences between agents in antipsychotic efficacy and tolerability. Further research is required to more definitively delineate these differences.

CLASSIFICATION, DRUGS &THERAPEUTIC AGENT OF SULFONAMIDES

Savitri Kumari[*], Minakshi Gupta[1], Anusha[1]

Assistant Professor, Faculty of Pharmaceutical Sciences, Baba Mastnath University, Asthal Bohar, Rohtak (124001)

Sulphonamides

Sulfonamides were the first class of antibiotics to be introduced in the 1930s. They remain important because they are effective, relatively safe and inexpensive, but adverse effects are relatively common. Up to 8% of hospitalised patients and 1–2% of those in the community are reported to suffer adverse effects from the combination of sulfamethoxazole with trimethoprim, although only about 3% of these are thought to represent hypersensitivity. The situation is markedly different in patients with HIV as up to 60% experience allergic adverse reactions. While most hypersensitivity reactions are relatively mild, sulfonamides account for a disproportionate number of cases of life-threatening Stevens-Johnson syndrome and toxic epidermal necrolysis. A sulfonamide grouping is derived from a sulfonic acid group by replacing its hydroxyl group with an amino group. Sulfonamides, also known as sulfa drugs, have a history that dates back to almost 70-80 years. A sulfonyl group plays a very important role as a key constituent of number of biologically active molecules. Sulfonamides occupy a unique position in the drug industry and exhibit a wide spectrum of biological activities. The first clinically used sulfonamide was named prontosil that showed protective action against streptococci in mice. Prontosil was active in vivo, but ineffective in vitro, which led to the conclusion that prontosil itself was not the active drug.

1.3 History of Sulfonamides:

In 1932, the German dye manufacturing company prepared a red azo dye, named prontosil for its dye properties. Remarkably, it was discovered that prontosil showed antibacterial action when it was used to dye wool. In 1935, Gerhard Domagk published the results of his research work indicating that prontosil was capable of curing staphylococcal infections in mice and rabbits. In 1939, Domagk earned Nobel Prize in medicine for this important discovery but an order from Hitler prevented Domagk from accepting the honour.

After sulfanilamide discovery, thousands of chemical variations were studied and the best therapeutic results were obtained from the compounds in which one hydrogen of the SO2NH2 group was replaced by heterocyclic ring. To date more than twenty thousand sulfanilamide derivatives, analogs and related compounds have been synthesized. These synthesis have resulted in the discovery of new compounds with varying pharmacological properties

1.4 Classification of Sulfonamides:

The general term ‘’sulfonamides’’ has been used for derivatives of *p*-aminobenzenesulfonamide (sulfanilamide), whereas specific compounds are described as N1 or N4-substituted sulfanilamides, depending on whether the substitution is on the sulfonamide amino group or aromatic amino group, respectively.

Most of the sulfonamides used currently are N1-derivatives.

Based on the structural variations, Johnson divided sulfonamides into three groups as follows:

- NI-acyl derivatives
- NI-heterocyclic derivatives containing six-membered rings (e.g. pyridine, pyrimidines, pyridazines and pyrazines).

- NI-heterocyclic derivatives containing five-membered rings (e.g. thiazole, oxazole, isoxazole, 1,3,4-thiadiazole and yrazole).

Another classification of sulfonamides is based on chemical structure, duration of action, spectrum of activity and therapeutic applications. The classification rate of absorption and half-life appears to be clinically relevant.

Based on this the sulfonamides are classified into three groups:

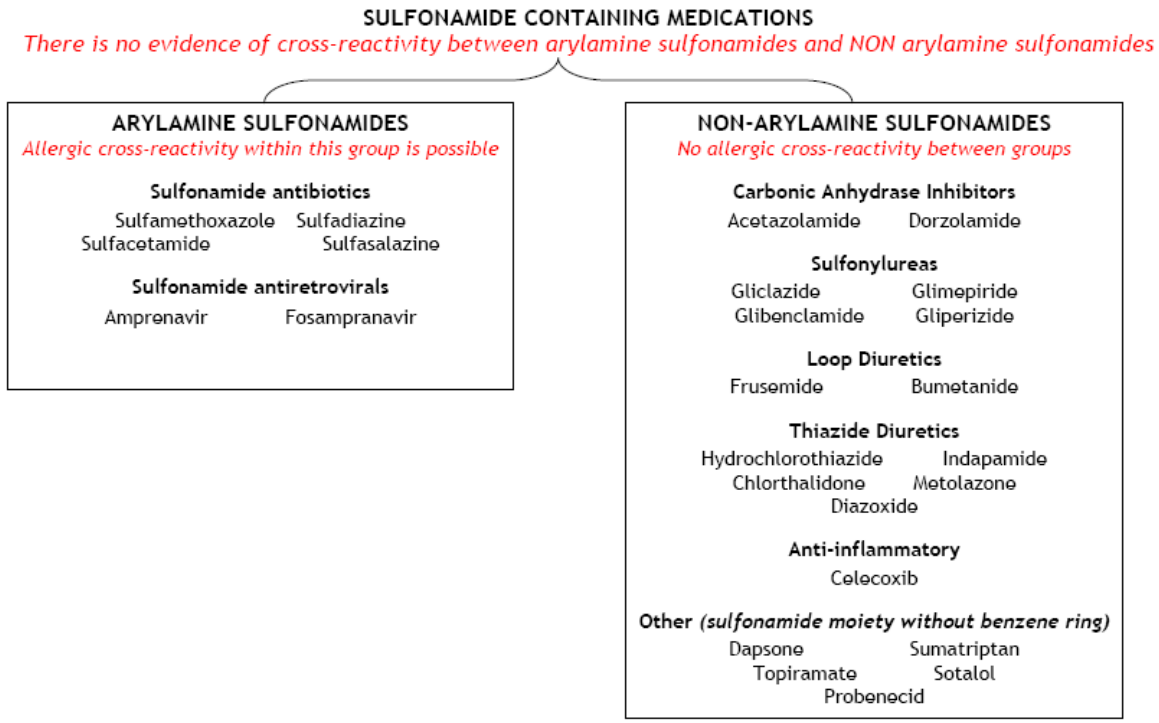
1. Short Acting
2. Intermediate Acting
3. Long Acting.

1. Short Acting: sulfonamides with a half-life less than 10 hours. (e.g. sulfamethazole, sulfisoxazole and sulfanilamide have been used for the treatment of urinary tract infections).

2. Intermediate Acting: Sulfonamides with a half-life between 10-24 hours. (e.g. sulfamethoxazole and sulfadiazine have been used for various infections especially active against invasive aspergillosis in AIDS patients).

3. Long Acting: Sulfonamides with a half-life longer than 24 hours. (e.g. Sulfadimethoxine and Sulfadioxine have been used for the treatment of ulceration colitis).

In addition to this, there are different types of sulfonamides which have been used in various types of infections such as mucous membrane, sulfabenzamide, superficial ocular infections, sulfaacetamide sodium, urinary infections, sulfadiazine and sulfamethazole, anticancer and others.



SULFONAMIDE CONTAINING DRUGS

Sulfur is found in drugs and preservatives in various forms including sulfites, sulfates, sulfhydryls and sulfonamides. Sulfonamide drugs are further differentiated into the those with an arylamine group (sulfonylarlamines) and those without. The sulfonylarlamines are almost exclusively sulfonamide antibiotics.

Sulfonamides antibiotics are among the most frequent causes of adverse drug reactions. The hypersensitivity is known to be associated with arylamine moiety. Therefore no cross-reactivity would be expected between sulfonamide antibiotics (sulfonylarlamines) and other

non-arylamine sulfonamides. While a small number of cases of possible cross-reactivity have been reported, these appear due to allergic susceptibility of the patient rather than a true cross-reactivity. Therefore non-antibiotic sulfonamide drugs are generally not considered as contraindicated in those with a history of hypersensitivity to antibiotic sulfonamides (sulfonylarlamines). Because of this, and the variety of sulfur containing drugs, use of the term "sulfur or sulfa allergy" is imprecise and should be avoided.

The need to precisely describe "allergies" is highlighted in the case of combined preparations. For example, a common antibiotic used in Australia is the combination of the sulfonamide antibiotic sulfamethoxazole and trimethroprim. When hypersensitivity to this preparation occurs, the sensitivity may to sulfamethoxazole or trimethroprim or both. In this case both sulfamethoxazole and trimethroprim should be documented.

SULFONAMIDES AS THERAPEUTIC AGENTS:

Antibacterial and antifungal:

Sulfonamides are a class of broad-spectrum synthetic bacteriostatic agents. They inhibit multiplication of bacteria but do not actively kill bacteria. They have been used against most gram-positive and many gram-negative organisms, some fungi and certain protozoa. A large number of substituted sulfonamide derivatives are used in pharmaceutical preparations as antibacterial and antifungal agents.

Sulfonamides as carbonic anhydrase inhibitors:

Carbonic anhydrase is an enzyme that helps to regulate the acid-base balance and *pH* in blood and other tissues. One of the functions of the enzyme is to interconvert carbon dioxide and bicarbonate. Carbonic anhydrase inhibitors are a class of pharmaceuticals that suppress the activity of carbonic anhydrases. The clinical use has been established as antiglucoma agents, diuretics, and antiepileptic in the management of mountain sickness, gastric and duodenal ulcers, neurological disorders or osteoporosis

Sulfonamides as anticancer agents:

Anticancer drugs are used to control the growth of cancerous cells. Cancer is commonly defined as the uncontrolled growth of cells, with loss of differentiation and commonly with metastasis, spread of the cancer to other tissues and organs. Cancers are malignant growths. In contrast, benign growths remain encapsulated and grow within a well-defined area. Benign tumors, if untreated may be fatal, due to pressure on essential organs, as in the case of a benign brain tumour. Surgery or radiations are the preferred method of treating growth, which has a well defined location. Drug therapy is used when the tumour is spread, or may spread to other areas of the body. Some of the sulfonamides exhibit anticancer activity.

REFERENCES:

1. Smith, W. B. (2008). "'Sulfur allergy' label is misleading." Aust Prescr 31(1): 8-10.
2. Dibbern, D. A., Jr., A. Montanaro, et al. (2008). "Allergies to sulfonamide antibiotics and sulfur-containing drugs." Annals of Allergy, Asthma, & Immunology 100(2): 91-100; quiz 100-3.
3. Ponka, D. and D. Ponka (2006). "Approach to managing patients with sulfa allergy: use of antibiotic and nonantibiotic sulfonamides." Canadian Family Physician 52(11): 1434-8.
4. Johnson K, Green D, Rife JP and Limon L. (2005) Sulfonamide Cross-Reactivity: Fact of Fiction? The Annuals of Pharmacotherapy 39(Feb):290-301.

DIFFERENT TYPES OF DOSAGE FORMS USED IN GLAUCOMA : A REVIEW

Sneh Lata*, Savitri[1]

Assistant Professor, Shri Baba Mastnath University of Pharmaceutical Sciences And Research, Asthal Bohar, Rohtak (124001)

Glaucoma is the second leading cause of blindness worldwide, disproportionately affecting women and Asians. There were 60.5 million people with OAG and ACG in 2010, increasing to 79.6 million by 2020, and of these, 74% will have OAG. Women comprise 55% of OAG, 70% of ACG, and 59% of all glaucoma in 2010(1). Glaucoma is a disease with a characteristic of a higher level of intraocular pressure (IOP) which might progressively hurt visibility. The average IOP of population is 15.5 ± 2.57 mmHg. The people, whose intraocular pressure is 20.5mmHg or more, could be suspected of having glaucoma, and IOP over 24 mmHg were definite case of glaucoma. The constraint of physiological factors, the ocular bioavailability of ophthalmic preparations is much lower than other dosing route due to factors such as lacrimal drainage, reflex tearing and drug spoilage onto the cheek (2). Glaucoma is a slow progressive degeneration of the retinal ganglion cells (RGCs) and the optic nerve axons, leading to irreversible blindness if left undiagnosed and untreated. Although increased intraocular pressure is a major risk factor of glaucoma, other factors include increased glutamate levels, alterations in nitric oxide (NO) metabolism, vascular alterations and oxidative damage caused by reactive oxygen species(3).

2. SYMPTOMS OF GLAUCOMA(6)

Open-angle Glaucoma

Open-angle glaucoma, the most common form, initially has no symptoms. The pressure in the eye builds up gradually. At some point, the optic nerve is damaged and side vision (peripheral vision) is lost. Without treatment, total blindness will occur. Similarly, people with normal-tension glaucoma will not experience any symptoms until they begin to lose peripheral vision.

Acute Angle-closure Glaucoma

Acute angle-closure glaucoma is the result of a sudden blockage in the normal flow of eye fluid (aqueous humor) between the iris and the lens.

Symptoms of acute angle-closure glaucoma may include

- severe pain,
- nausea,
- vomiting,
- blurred vision
- seeing a rainbow halo around lights.

Acute angle-closure glaucoma is a medical emergency and must be treated immediately or blindness could result in one or two days. Chronic angle-closure glaucoma progresses more slowly and can damage the optic nerve without symptoms as in open-angle glaucoma.

1.3 TYPES OF GLAUCOMA(5)

There are many different types of glaucoma. Most, however, can be classified as either open-angle glaucomas, which are usually conditions of long duration (chronic), or angle-closure (closed angle) glaucomas, which include conditions occurring both suddenly (acute) and over a long period of time (chronic). The glaucomas usually affect both eyes, but the disease can progress more rapidly in one eye than in the other. Involvement of just one eye occurs only

when the glaucoma is brought on by factors such as a prior injury, inflammation, or the use of steroids only in that eye.

1.6 Ophthalmic dosage forms:

A multitude of ocular dosage forms are available for delivery of drugs to the eye. These can be classified on the basis of their physical forms as follows:

1) Liquids:	Solutions, Suspensions, Sol to gel systems, Sprays
2) Solids:	Ocular inserts, Contact lenses, corneal shield, Artificial tear inserts, Filter paper strips
3) Semi-solids:	Ointments, Gels
4) Miscellaneous:	Ocular iontophoresis, Vesicular systems, Mucoadhesive dosage forms, Particulates, Ocular penetration.

1.6.1 Liquids

Liquids are the most popular and desirable state of dosage forms for the eye. This is because the drug absorption is fastest from this state. The slow release of the drug from the suspended solids provides a sustained effect for a short duration of time.

1.6.1.1 Solutions and Suspensions

Solutions are the pharmaceutical forms most widely used to administer drugs that must be active on the eye surface or in the eye after passage through the cornea or the conjunctiva. The drug in the solution is in the solved state and may be immediately active. This form also has disadvantages; the very short time for which solution stays at the eye surface, its poor bioavailability (a major portion i.e. 75% is lost via nasolacrimal drainage), the instability of the dissolved drug and the necessity of using preservatives. The rapid elimination is due to solution state of the preparation and may be influenced by the composition of the solution. The retention of a solution in the eye is influenced by viscosity, hydrogen ion concentration, the osmolality and the instilled volume. Extensive work has been done to prolong ocular retention of drugs in the solution state by enhancing the viscosity or altering the pH of the solution [11].

1.6.1.2 Sol to gel Systems:

The new concept of producing a gel in situ (eg. in the cul-de-sac of the eye) was suggested for the first time in the early 1980s. It is widely accepted that increasing the viscosity of a drug formulation in the precorneal region will leads to an increased bioavailability, due to slower drainage from the cornea. There are four broadly defined mechanisms used for triggering the in situ gel formation of biomaterials: Physiological stimuli (e.g., temperature and pH), physical changes in biomaterials (e.g., solvent exchange and swelling), chemical reactions (e.g., enzymatic, chemical and photo-initiated polymerization).

1.6.1.2.1 In situ formation based on physiological stimuli:

Thermally trigged systems are the most commonly studied class of environment-sensitive systems**5**. The ideal critical temperature range for such system is ambient and physiologic temperature, such that clinical manipulation is faccilated and no external source of heat other than that of body is required for trigger gelation.

pH triggered systems cause formation of gel which is induced by pH changes**5**. All the pH-sensitive polymers contain pendant acidic or basic groups that either accept or release protons in response to changes in environmental pH . Swelling of hydrogel increases as the external pH increases in the case of weakly acidic (anionic) groups, but decreases if polymer contains weakly basic (cationic) groups**7**.

1.6.1.2.2. In situ formation based on physical mechanism:

Swelling of any material may occur when it absorbs water from surrounding environment and expands to desired space**14**. One such substance is myverol 18-99 (glycerol mono-oleate), which is polar lipid that swells in water to form lyotropic liquid crystalline phase

structures.It has some Bioadhesive properties and can be degraded invivo by enzymatic action**15**.

Diffusion type involves the diffusion of solvent from polymer solution into surrounding tissue and results in precipitation or solidification of polymer matrix. N-methyl pyrrolidone (NMP) has been shown to be useful solvent for such system**16**.

1.6.1.2.3. In situ formation based on chemical reactions:

Chemical reactions that results in *situ* gelation may involve precipitation of inorganic solids from supersaturated ionic solutions, enzymatic processes and photo-initiated processes.

Ionic crosslinking occurs in the presence of ions. Some of the polysaccharides fall into the class of ion-sensitive ones**17** .While k-carrageenan forms rigid, brittle gels in reply of small amount of K^+, i-carrageenan forms elastic gels mainly in the presence of Ca^{2+}. Gellan gum commercially available as Gelrite® is an anionic polysaccharide that undergoes *in situ* gelling in the presence of mono- and divalent cations, including Ca^{2+}, Mg^{2+}, K^+ and Na^+. Gelation of the low-methoxy pectins can be caused by divalent cations, especially Ca^{2+} . Likewise, alginic acid undergoes gelation in presence of divalent/polyvalent cations e.g. Ca^{2+} due to the interaction with guluronic acid block in alginate chains**18**.

Enzymatic cross-linking has not been investigated widely but seems to have some advantages over chemical and photochemical approaches. Intelligent stimuli-responsive delivery systems using hydrogels that can release insulin have been investigated. Cationic pH-sensitive polymers containing immobilized insulin and glucose oxidase can swell in response to blood glucose level releasing the entrapped insulin in a pulsatile fashion. Adjusting the amount of enzyme also provides a convenient mechanism for controlling the rate of gel formation, which allows the mixtures to be be injected before gel formation**19**.

Photo-polymerisation is commonly used for *in situ* formation of biomaterials. A solution of monomers or reactive macromer and initiator can be injected into a tissues site and the application of electromagnetic radiation used to form gel**5**.Acrylate or similar polymerizable functional groups are typically used as the polymerizable groups on the individual monomeers amd macromers because they rapidly undergo photo-polymerisation in the presence of suitable photoinitiator. Typically long wavelength ultraviolet and visible wavelengths are used.Short wavelength ultraviolet is not used often because it has limited penetration of tissue and biologically harmful.

1.6.1.3 Sprays

Although not commonly used, some practitioners use mydriatics or cycloplegics alone or in combination in the form of eye spray. These sprays are used in the eye for dilating the pupil or for cycloplegic examination.

1.6.2 Solids

The concept of using solids for the eye is based on providing sustained release characteristics.

1.6.2.1 Ocular inserts

Ocular inserts overcome the disadvantage reported with traditional ophthalmic systems like aqueous solutions, suspensions and ointments. The typical pulse entry type drug release behaviour observed with traditional systems is replaced by ocular inserts. A number of ocular inserts were prepared utilizing different techniques to make soluble, erodible, nonerodible, and hydrogel inserts[12]. Soluble ocular inserts, such as the poly(vinyl alcohol) insert (PVAI), the soluble ophthalmic drug insert (SODI) and polypeptide devises are matrix type polymeric devises used for drug delivery to eye. Poly(vinyl alcohol) inserts are characteristically thin, elastic and oval shaped plates and impregnated with antibiotics, sulfonamides, pilocarpine, atropine, or other drugs used in ophthalmology.

1.6.2.2 Contact lenses

These drug saturated contact lenses are placed in the eye for releasing the drug for long period of time. Hydrophilic soft contact lenses are made up of hydrogels that absorb certain

amounts of aqueous solutions, because of this property they have also been found useful for drug delivery to anterior segment of the eye. In humans, the Bionite lens which was made from hydrophilic polymer (2-hydroxy ethyl methacrylate) has been shown to produce a greater penetration of fluorescein.

1.6.2.3 Corneal shield

A non cross-linked homogenized, porcine scleral collagen slice is developed by a company (Biocor (Bausch and Lomb pharmaceuticals). Topically applied antibiotics have been used in conjunction with the shield to promote healing of corneal ulcers. Collagen shields are fabricated with foetal calf skin tissue and originally developed as a corneal bandage. These devices, once softened by the tear fluid, form a thin pliable film that confirms exactly to the corneal surface, and undergoes dissolution up to 10, 24 or 72 hours.

1.6.2.4 Artificial tear inserts

A rod shaped pellet of hydroxypropyl cellulose without preservative is commercially available (Lacrisert). This device is designed as a sustained release artificial tear for the treatment of dry eye disorders. It was developed by Merck, Sharp and Dohme in 1981[13].

1.6.2.5 Filter paper strips

Sodium fluorescein and rose Bengal dyes are commercially available as drug impregnated filter paper strips. These dyes are used diagnostically to disclose corneal injuries and infections such as herpes simplex, and dry eye disorders.

1.6.3 Semi-solids

A wide variety of semisolids vehicles are used for topical ocular delivery which falls into two general categories: simple and compound bases. Simple bases refer to a single continuous phase. These include white petrolatum, lanolin and viscous gels prepared from polymers such as PVA, carbopol etc. Compound bases are usually of a biphasic type forming either water in oil or oil in water emulsions. A drug in either a simple or compound base provide an increase in the duration of action due to reduction in dilution by tears, reduction in drainage by way of a sustained release effect, and prolonged corneal contact time. The most commonly used semisolid preparation is ointments consisting of dispersion of a solid drug in an appropriate vehicle base.

1.6.4 Miscellaneous:

1.6.4.1 Capsular-type Drug Delivery Systems:-

1.6.4.1.1 Ocusert® and Related Devices

Ocusert® is a drug delivery device for hydrophilic drugs. A truly continuous, controlled release and zero order kinetic fashion is achieved using ocusert e.g. Pilocarpine ocusert. This system consists of a pilocarpine-alginate core (drug) sandwiched between two transparent, rate-controlling ethylene-vinyl acetate copolymer membranes. When it is placed under the upper or lower eyelid, the pilocarpine molecules dissolved in the lacrimal fluid are released through the rate-controlling membranes at predefined rates.

Advantages of occuserts:

1. Increased ocular residence time hence a prolonged drug activity and a higher bioavailability with respect to standard vehicle.
2. Possibility of releasing drug at a slow constant rate.
3. Accurate dosing (contrary to eye drop that can be improperly instilled by the patient and are partially lost after administration each insert can be made to contain a precise dose which is fully retained at the administration site).
4. Reduction of systemic absorption (which occurs freely with eye drops via the nasolacrimal duct and nasal mucosa).
5. Better patient compliance, resulting from a reduced frequency of administration and a lower incidence of visual and systemic side effects.

6. Possibility of targeting internal ocular tissues through non corneal(conjunctival sectional) routes.
7. Increased shelf life with respect to aqueous solutions.
8. Exclusion of preservative, thus reducing the risk of sensitivity reactions.
9. Possibility of incorporating various level novel chemical/technological approaches, such as prodrugs, mucoadhesives, permeation enhancers, microparticulates, salts acting as buffers etc.

Disadvantages of occusert:

1. Initial discomfort, their movement around the eye.
2. Occasional inadvertent loss during sleeps or while rubbing the eye.
3. Interference with vision and a difficult placement.

1.6.4.1.2 Implantable Silicone Rubber Devices

Refojo et al., 1978 developed a constant release rate implantable silicone rubber device for hydrophobic drug, BCNU (1, 3-bis (2-chloroethyl)-1-niurosourea an intraocular malignancy agent. This device consists of two sheets of silicone rubber (Silastic® 500-1, 0.13 mm thick) glued together only at the edges with silicone adhesive. A tube of the same material extends from device. The device released BCNU at a nearly constant rate (about 200-400 µg/h) for a time determined by amount of drug in the device.

1.6.4.2. Implantable Drug Delivery Pumps

1.6.4.2.1. Osmotic Mini pump and Implantable Infusion System

Implantable devices include an osmotic minipump, a drug pellet coated with polyvinyl alcohol and ethylene vinyl acetate, and polysulfone capillary fiber. The generic osmotic minipump (ALZET®) is a useful implantable drug delivery system with a constant drug delivery rate with a pumping duration of up to 2 weeks. Another drug delivery pumping system is the Infusaid®, which is an implantable infusion system (Infusaid Corporation, USA). The pumping force is generated by an expanding fluid (a flurocarbon in liquid-gas equibillirium) at body temperature.

CONCLUSION:

The main efforts in ocular drug delivery during the past two decades has been on the design of systems to prolong the residence time of topically applied drugs in conjunctival sac. The advantages offered by ophthalmic inserts are numerous but only few of them have gained commercial acceptance. This is because of its comparatively high cost and reluctance of the patients to use unfamiliar types of ophthalmic medication. There is need to apprise the patients with the benefits of such systems and also to make them familiar with the methods of using such devices. Ophthalmic inserts design, construction and technology is witnessing a rapid improvement, so in near future its use is expected to increase tremendously in ophthalmic therapy.

REFERENCES

1. Quigley HA, Broman AT. The number of people with glaucoma worldwide in 2010 and 2020. Br J Ophthalmol. 2006; 90(3):262-267.
2. Chiang Chiao-Hsi. Ocular drug delivery systems of antiglaucoma agents. J Med Sci. 1991; 12(3):157-170.
3. Gupta SK, Galpalli ND, Agarwal SS, Srivastava S, Saxena R. Recent advances in pharmacotherapy of glaucoma. Indian Journal of Pharmacology. 2008; 40(5):197-208.
4. AXS Biomedical Animation Studio. Inc. Glaucoma: Challenge (Internet). Available from: http://www.cnib.ca/en/your-eyes/eye-conditions/Glaucoma/Glaucoma Challenge/glaucoma_summary_EN.pdf
5. Dahl AA, Glaucoma (Internet). Available from: http://www.medicinenet.com/glaucoma/article.htm
6. Bright focus foundation. About glaucoma (Internet). Available from: http://www.brightfocus.org/glaucoma/about/understanding/facts.html
7. Patel HA, Patel JK, Patel KN, Patel RR. Ophthalmic drug delivery system-a review. Scholars research library. Der Pharmacia Lettre, 2010; 2(4):100-115.

8. Nirmal HB, Bakliwal SR, Pawar SP. In-situ gel: new trends in controlled and sustained drug delivery system. International Journal of Pharm Tech Research. 2010 April; 2(2): 1398-1408.
9. Marsha Ritter Jones, MS, Philip B. Massersmith, In-situ gel forming biomaterials. Oral Maxillofacial Surg Clin N Am 14 (2002):29-38.
10. Qiu Y, Park K. Environment-sensitive hydrogels for drug delivery. Adv Drug Deliv Rev 2001; 53:321-39.
11. Esposito E, Carratto V et al. Comparative analysis of tetracycline containing dental gels; poloxomers and mono-glycerides based formulation. Int. J. Pharm. 1996; 142:9-23.
12. Geraghaty P, Attwood D et al. An investigation of parameters influencing the Bioadhesive properties of Myverol 18-99/ water gels. Biomaterials 1997, 18:63-7.
13. Motto F, Gailloud P, et al., In-vitro assessment of new embolic liquids prepared from preformed polymers and water miscible solvents aneurysm treatment. Biomaterials 2000; 21:803-11.
14. Bhardwaj TR, Kanwar M, Lal R, Gupta A. Natural gums and modified natural gums as sustained release carriers. Drug Devel Ind Pharm 2000; 26:1025-1038.
15. Guo J-H, Skinner GW, Harcum WW, Barnum PE. Pharmaceutical applications of naturally occurring water-soluble polymers. Pharm Sci & Technol Today 1998; 1:254-261.
16. Podual K, Doyle III FJ, Peppas NA. Dynamic behavior of glucose oxidase-containing microparticles of poly(ethylene)- grafted cationic hydrogels in an environment of changing pH. Biomaterials 2000; 21:1439-50.
17. Sawhney AS, Pathak CP, Hubbell JA, Hill JL, Desai NP. Photopolymerizable biodegradable hydrogels as tissue contacting materials and controlled release carriers.US Patent 5410016. 1995.
18. Ritch R. Potential role for Ginkgo biloba extract in the treatment of glaucoma. Med Hypotheses. 2000; 54:221-35.
19. Robert NW. Pathophysiology, classification and treatment options of Glaucoma. Can J Opthalmol. 2007; 42:396-8

CARBON NANOTUBES AND ITS APPLICATION IN PHARMACY: A REVIEW

Bijender Sagar[1*]

Assistant Professor,[1]Shri Baba Mastnath Institute of Pharmaceutical Sciences and Research, Asthal Bohar, Rohtak-124001 (Haryana) E-mail: bijendersagar@rediffmail.com

1. INTRODUCTION

CNTs are allotropes of carbon. They are tubular in shape, made of graphite. CNTs possess various novel properties that make them useful in the field of nanotechnology and pharmaceuticals. They are nanometers in diameter and several millimeters in length and have a very broad range of electronic, thermal, and structural properties. These properties vary with kind of nanotubes defined by its diameter, length, chirality or twist and wall nature. Their unique surface area, stiffness, strength and resilience have led to much excitement in the field of Pharmacy.

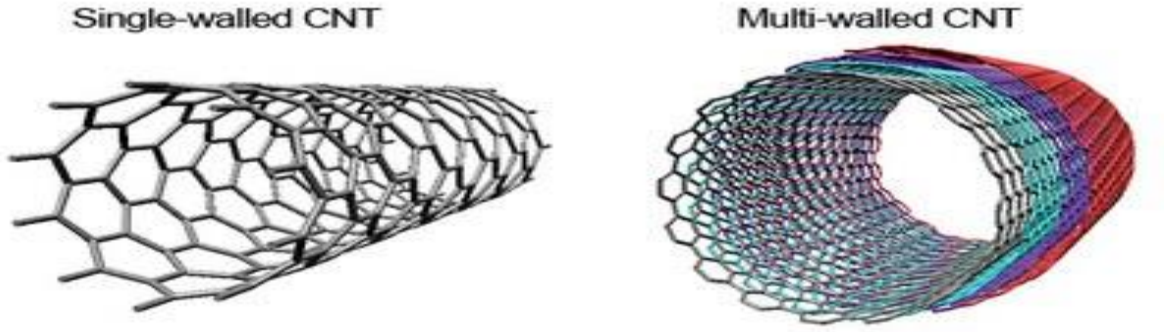


Fig: Fullerene molecules

2. CNT Morphology and Structure

CNTs belong to the fullerene family of carbon allotropes. They are cylindrical molecules consisting of a hexagonal arrangement of sp^2-hybridized carbon atoms (C-C distance of 1.4 Å). They are described as hollow cylinders formed by rolling single or multiple layers of grapheme sheets into seamless cylinders. These cylindrical structures have two forms (SWNTs) and multi walled carbon nanotubes (MWNTs). MWNTs comprise several to tens of concentric cylinders of graphitic shells, each one forming a SWNT. MWNTs generally have a larger outer diameter (2.5–100 nm) than SWNTs (0.6–2.4 nm) and consist of a varying number of concentric SWNT layers, with an interlayer separation of about 0.3nm.(R. Chavan *et al.*,2012)

3. Synthesis of Carbon Nanotubes

Carbon nanotubes are generally produced by three main techniques, arc discharge, laser ablation and chemical vapour deposition.

- In arc discharge, a vapour is created by an arc discharge between two carbon electrodes with or without catalyst. Nanotubes self-assemble from the resulting carbon vapour.
- In the laser ablation technique, a high-power laser beam impinges on a volume of carbon containing feed stock gas (methane or carbon monoxide). At the moment, laser ablation produces a small amount of clean nanotubes, whereas arc discharge methods generally produce large quantities of impure material.
- In general, chemical vapour deposition (CVD) results in MWNTs or poor quality SWNTs.(K. Anazawa *et al.*,2002)

4. Characterisation and Properties of of CNTs

- TEM allowing for the assessment of detailed structures.
- SEM providing overviews of sample structures while less sensitive to sample preparation and homogeneity than TEM.
- RAMAN Spectroscopy suitable for the quick and reliable screening of the presence of SWCNT.
- Thermogravimetric analysis giving information about relative abundance of catalyst particles, nanotubes and other carbonaceous structures.

  CNTs have very interesting physicochemical properties such as ordered structure with high aspect ratio, ultralight weight, high mechanical strength, high electrical conductivity, high thermal conductivity, metallic or semi-metallic behaviour and high surface area.(w. Yang *et al.*,2007)

5. Application of CNTs (R. Hirlekar *et al.*,2009)

Various applications of CNTs are as follows:

5.1 Carrier for Drug delivery:

- Carbon nanohorns (CNHs) are the spherical aggregates of CNTs with irregular horn like shape. Research studies have proved CNTs and CNHs as a potential carrier for drug delivery system.
- Functionalised carbon nanotubes are reported for targeting of Amphotericin B to Cells.
- Cisplatin incorporated oxidized SWNHs have showed slow release of Cisplatin in aqueous environment. The released Cisplatin had been effective in terminating the growth of human lung cancer cells, while the SWNHs alone did not show anticancer activity
- Anticancer drug (Polyphosphazene platinum) given with nanotubes had enhanced permeability, distribution and retention in the brain due to controlled lipophilicity of nanotubes
- CNT-based carrier system can offer a successful oral alternative administration of Erythropoietin (EPO), which has not been possible so far because of the denaturation of EPO by the gastric environment conditions and enzymes.
- They can be used as lubricants or glidants in tablet manufacturing due to nanosize and sliding nature of graphite layers bound with van der waals forces.

5.2 Genetic Engineering

In genetic engineering, CNTs and CNHs are used to manipulate genomes and atoms in the development of bioimaging genomes, proteomics and tissue engineering.

5.3 Preserative

Carbon nanotubes and nanohorns are antioxidant in nature. Hence, they are used to preserve drugs formulations prone to oxidation.

5.4 Diagnostic Tool

Protein-encapsulated or protein/enzyme filled nanotubes, due to their fluorescence ability in presence of specific biomolecules have been tried as implantable biosensors.

6. Limitation

- Difficulty in maintaining high quality and minimal impurities.
- Lack of solubility in most solvents compatible with the biological milieu (aqueous based).

7. CONCLUSION

This review on carbon nanotubes reveals the overview on structure, morphology, synthesis of carbon nanotubes along with their properties benefits and applications. The properties and characteristics of CNTs are still being researched heavily and scientists have barely begun to

tap the potential of these structures. Single and multiplewalled carbon nanotubes have already proven to serve as safer and more effective alternatives to previous drug delivery methods. They can pass through membranes, carrying therapeutic drugs, vaccines, and nucleic acids deep into the cell to targets previously unreachable. Overall, recent studies regarding CNTs have shown a very promising glimpse of what lies ahead in the future of medicine.

8.REFERENCES

1. Anazawa K., Shimotani K., Manabe C., Watanabe H., Shimizu M. High purity carbon nanotube synthesis method. *Applied Physics Letters*.2002;8(1):739-741.
2. Yang W., Thordarson P., Gooding JJ., Ringer SP., Braet F. Carbon nanotubes for biological and biomedical applications. Nanotechnology2007;18: 412001.
3. Chavan R., Desai U., Mhatre P., Chinchole R. Carbon Nanotubes: a review. *Int. J of Pharmaceutical Sciences Review and Res.*2012;13(1):125-134
4. Hirlekar R., Yamagar M., Garse H., Vij M.,Kadam V. Carbon Nanotubes and its application: a review. *Asian J of Pharmaceutical and Clinical Res.*2009;2(4):17-27

CLASSIFICATION AND METHOD OF PREPARATION OF LIPOSOMES

Anusha[1], Minakshi Gupta[1]

Assistant Professor, Shri Baba Mastnath University of Pharmaceutical Sciences And Research, Asthal Bohar, Rohtak (124001)

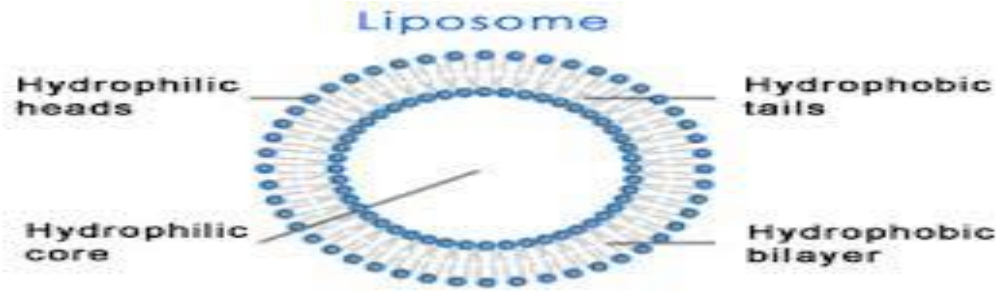
INTODUCTION

Liposomes, defined as microscopic spherical-shaped vesicles, consist of an internal aqueous compartment entrapped by one or multiple concentric lipidic bilayers. Liposomes membrane is composed of natural and/or synthetic lipids which are relatively biocompatible, biodegradable and non-immunogenic material. Because of their unique bilayer-structure properties, liposomes are used as carriers for both lipophilic and water-soluble molecules. Hydrophilic substances are encapsulated in the interior aqueous compartments. Lipophilic drugs are mainly entrapped within lipid bilayers. Liposomes are a form of vesicles that consist either of many, few or just one phospholipid bilayers. The polar character of the liposomal core enables polar drug molecules to be encapsulated. Amphiphilic and lipophilic molecules are solubilized within the phospholipid bilayer according to their affinity towards the phospholipids. Participation of nonionic surfactants instead of phospholipids in the bilayer formation results in niosomes. Channel proteins can be incorporated without loss of their activity within the hydrophobic domain of vesicle membranes, acting as a size-selective filter, only allowing passive diffusion of small solutes such as ions, nutrients and antibiotics. Thus, drugs that are encapsulated in a nanocagefunctionalized with channel proteins are effectively protected from premature degradation by proteolytic enzymes. The drug molecule, however, is able to diffuse through the channel, driven by the concentration difference between the interior and the exterior of the nanocage. Liposomes have the distinct advantages of being both nontoxic and biodegradable because they are composed of naturally occurring substances. Biologically active materials encapsulated within liposomes are protected to varying extent from immediate dilution or degradation, suggesting drug carrier systems for the transport of drugs and other bioactive capsules to disease-affected organs. The unique ability of liposomes to entrap drugs both in an aqueous and a lipid phase make such delivery systems attractive for hydrophilic and hydrophobic drugs. Because of advancements in the methods of preparing and formulating liposomes, high-entrapment efficiencies are possible for incorporating drugs into liposomes, creating a tremendous pharmaceutical impact. Furthermore, such encapsulation has been shown to reduce drug toxicity while retaining or improving the therapeutic efficacy. Several laboratories have reported the use of liposomes as drug carriers in the treatment of cancer, leishmaniasis, metabolic disorders, and fungal diseases. Innovative research in liposomal drugs has led to commercialization of several anticancer therapeutics such as Doxil, Myocet, two liposome-based anticancer drugs; doxorubicin; and an antifungal drug formulation, AmBisome, which is a liposomal formulation of amphotericin B used for systemic therapy. Liposomes may have a use in gene delivery to correct gene-associated disorders or for vaccine therapy. A quantitative entrapment of DNA can be achieved using the preparation of empty liposomes with cationic

lipids followed by mixing with DNA or a plasmid of interest. Because of its convenience and efficacy, cationic lipid mediated gene delivery technology is a promising system for in vivo gene therapy. Clinical trials of large-size lipid–DNA complexes have mostly shown a lack of adverse effects and moderate expression in a relatively low fraction of cells, but no decisive clinical disadvantages.

CLASSIFICATION OF LIPOSOMES

Liposomes can be classified either on the basis of their structural properties or on the basis of the preparation method used. These two classification system are in principle, independent of each other. Dependent on the selection of lipids, the preparation technique, and preparation conditions, liposomes can vary widely in size, number, position of lamellae. These parameters influence the behaviour of liposomes both in vivo and in vitro.



Liposome classification based on structural features

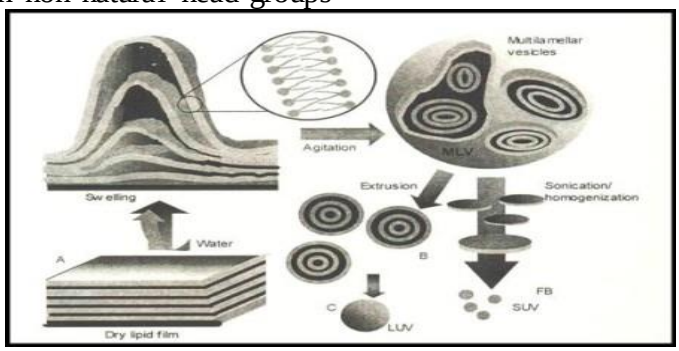
- (MLV) Multilamellar large vesicles
- (OLV) Oligolamellar vesicles
- (UV) Unilamellar vesicles
- (SUV) Small unilamellar vesicles
- (MUV) Medium sized unilamellar vesicles
- (LUV) Large unilamellar vesicles
- (GUV)Giant unilamellar vesicles
- (MVV)Multivesicular vesicles

Liposome classification based on method of liposome preparation

- (REV) Single or oligolamellar vescile made by reverse phase evaporation method.
- (MLV / REV) Multilamellar vesicles made by reverse phase evaporation method.
- (SPLV) Stable plurilamellar vesicles.
- (FATMLV) Frozen and thawed MLV
- (VET) Vesicles prepared by extrusion method.
- (FUV) Vesicles prepared by fusion
- (FPV) Vesicles prepared by french press
- (DRV) Dehydration- rehydration vesicles

Five groups of phospholipids that can be used for the liposomal preparation can be discerned:

1. Phospholipids from natural sources
2. Modified natural phospholipids
3. Semi synthetic phospholipids
4. Fully synthetic phospholipids
5. Phospholipids with non-natural head groups



Mechanism of preparation of liposomes

ADVANTAGES OF LIPOSOME

- Liposomes increased efficacy and therapeutic index of drug (actinomycin-D)

- Liposome increased stability via encapsulation
- Liposomes are non-toxic, flexible, biocompatible, completely biodegradable, and nonimmunogenic
- for systemic and non-systemic administrations
- Liposomes reduce the toxicity of the encapsulated agent (amphotericin B, Taxol)
- Liposomes help reduce the exposure of sensitive tissues to toxic drugs
- Site avoidance effect
- Flexibility to couple with site-specific ligands to achieve active targeting

DISADVANTAGES OF LIPOSOME

- Low solubility
- Short half-life
- Sometimes phospholipid undergoes oxidation and hydrolysis-like reaction
- Leakage and fusion of encapsulated drug/molecules
- Production cost is high
- Fewer stables

METHODS OF LIPOSOME PREPARATION

General methods of preparation

All the methods of preparing the liposomes involve four basic stages:

1. Drying down lipids from organic solvent.
2. Dispersing the lipid in aqueous media.
3. Purifying the resultant liposome.
4. Analyzing the final product.

Method of liposome preparation and drug loading

The following methods are used for the preparation of liposome:

1. Passive loading techniques
2. Active loading technique.

Passive loading techniques include three different methods

1. Mechanical dispersion method.
2. Solvent dispersion method.
3. Detergent removal method (removal of nonencapsulated material).

MECHANICAL DISPERSION METHOD

The following are types of mechanical dispersion methods:

1. Sonication.
2. French pressure cell: extrusion.
3. Freeze-thawed liposomes.
4. Lipid film hydration by hand shaking, non-hand. shaking or freeze drying.
5. Micro-emulsification.
6. Membrane extrusion.
7. Dried reconstituted vesicles.

CONCLUSIONS

Liposomes have been used in a broad range of pharmaceutical applications. Liposomes are showing particular promise as intracellular delivery systems for anti-sense molecules, ribosomes, proteins/peptides, and DNA. Liposomes with enhanced drug delivery to disease locations, by ability of long circulation residence times, are now achieving clinical acceptance. Also, liposomes promote targeting of particular diseased cells within the disease site. Finally, liposomal drugs exhibit reduced toxicities and retain enhanced efficacy compared with free complements. Only time will tell which of the above applications and speculations will prove to be successful. However, based on the pharmaceutical applications and available products, we can say that liposomes have definitely established their position in modern delivery systems.

REFERENCE:

1. Ahmad et al., Antibody-Targeted Delivery of Doxorubicin Entrapped in Sterically Stabilized Liposomes Can Eradicate Lung Cancer in Mice, Cancer Res., 1993, 53, 1484–1488.
2. Mayhew F. J., Goldrosen R. and Vaage J., Effects of Liposome- Entrapped Doxorubicin on Liver Metastases of Mouse Colon Carcinomas 26 and 38, J. Natl. Cancer Inst., 1987, 78, 707–713. 3. Alving C. R. et al., Therapy of Leishmaniasis: Superior Efficacies of Liposome-Encapsulated Drugs, Proc. Natl. Acad. Sci. (USA), 1978, 75, 2959–2963.
3. Lopez-Berestein G. et al., Liposomal Amphotericin B for the treatment of Systemic Fungal Infections in Patients with Cancer: A Preliminary Study, J. Infect. Dis., 1985, 151, 704–710.
4. Gregoriadis G. Genetic Vaccines: Strategies for Optimization, Pharm. Res., 1998, 15, 661–670.
5. Spanjer H. H., Scherphof G., Targetting of lactosylceramidecontaining liposomes to hepatocytes in vivo. Biochim Biophys. Acta, 1983, 174, 40-47.
6. Nicolau C., Legrand A., Grosse E., Liposomes as carriers for in vivo gene transfer and expression, Methods in enzymology, 1987, 149, 157-176.
7. Barenholz Y., Crommelin D. J. A., Liposomes as pharmaceutical dosage forms. Encyclopedia of pharmaceutical Technology, 1994, 1-39.
8. Lichtenberg D., Bareholz Y., "Liposomes: Preparation, Characterisation and preservation," Methods of Biological analysis 33, New york. John wiley, 1988, 337-461.
9. Eibl H., Woolley P., Synthesis of enatiomerically pure glyceryl esters and ethers, I, Methods employing the precursor 1, 2- isopropylidene-sn-glycerol., Chem. Phys. Lipids 1986, 41, 5363.
10. Allen T. M., Hansen C., Martin F., Redemann C., Yau- young A., Liposomes containing synthetic lipid derivatives of poly(ethylene glycol) show prolonged circulation half lives in vivo", Biochem. Biophys. Acta, 1990, 1066, 29-36.

DEPARTMENT OF ENGINEERING

SYSTEM APPROACH TO TECHNICAL EDUCATION

V.K AHUJA

Head of Deptt. Civil Engg, Sbmnec,Asthal Bohar Rohtak-124001, Haryana

Main Features

1. Understand the system concept
2. Evolve Technical Education as a System
3. Identify the Various Subsystems and their Function
4. Necessity of Evaluation on the Technical Education System

1. Introduction

In present scenario the system concept is widely being used in Engineering & Technology. This helps in getting a clear picture of behavior and cause effect relationship in processes. System approach in problem solving and decision making is an overall integrated approach. It helps in designing an optimal system with a view to maximize the benefit for a given cost or to minimize the cost for a specified output. It helps in analyzing the system by identifying the actions interrelations dependencies and feedback between the subsystems of facilitates to evaluate the system to know weather they are meeting or some error is there so that to evolve an alternative. It provides logical, systematic, comprehensive, critical, judgmental and rational view to look at the system and its operations. Functionally, system approach may be taken as a technique for understanding, predicting and controlling the interrelations and interdependence of the major parts of the system in a given situation to achieve the stated objectives

2 .System concept & Terminology

(A) System
(B) Subsystem
(B) System Boundaries
(D) System Feedback
(E) System Functioning

SYSTEM: A system is a unity which is conceptual. It consists of a set of parts of elements which are interrelated interacting and interdependent. They work in particular way to achieve a set of predetermined objectives. They respond and obey some from of control.

They are not interdependent of the environment in which work. They are influenced by the changes occurring in the environment.

SUBSYSTEM:- These are components of a system. These can also be considered as systems of lower order.

SYSTEM BOUNDARIES ;- As system operates in an environments, an

Open system has interrelationship with the environment. A closed system works within the rigid boundaries its origination and the environment. It has no effect on the changes in environment due to techno, socio economic and political changes. A closed system sometime starts becoming open due to the effects of penetration of the effects of environinent which may penetrate boundaries. Unbounded open system may create some problems. So it must be restricted to a manageable level. This process helps in boundary building of the system.

- SYSTEM FEEDBACK:- Feedbck of any system is very vital to control a system to achieve and maintain its objectives without which the system may not be purposive system. Feedback may be obtained through the control function of a system which a my be closed loop or open lop. Closed loop control is a self controlling integrated part of the system. It does not only control the system output but regulates the system at every stage

to achieve the set goals. Open loop control is a externally built into the system and provides feedback intermittently. Hence it is incomplete and involves time loop. Feedback may be positive or negative. Negative feedback has a corrective function while Positive feedback has reinforcing function. Delayed, Disturbed & Distorted feedback results in creating instability in the system.

- **SYSTEM FUNCTIONING:-**The system as a whole & its subsystem & its component have their inputs which are the elements of the system. The operating processes of the system converts theses inputs into system inputs. The output represents the achievement of the system. The control of the system gives the output & feedback about the system against the set standards. The subsystem helps the system towards its goal state. Environmental pressure priorities, trends, problems and influences set the boundary of system operation.

3. Technical Education - It satisfies all the requirements of an open system. It may be taken as system for all purposes. It possess the following characteristics:
   - **I.** Supra system education of all levels.
   - **II.** Consists of number of interrelation, Interdependent and interacting system.
   - **III.** Variety of inputs in the form of students, staff courses, curricular griddling, equipments, and grounds funds as operating cost.
   - **IV.** Having societal needs, priorities, problems, future trends, linkage with other system influencing the system goals and objectives.
   - **V.** Number of operating processes converting inputs into outputs
   - **VI.** Outputs in the form of qualified technicians
   - **VII.** processes formative & summative evaluation approaches
   - **VIII.** Has a outward perspective boundary interacting with the techno, Socio, economic and political forces in the environment & adopting itself to the input characteristics & output requirements.

Subsystem of Technical Education System

(A) Classification according to Structure
   a) Teacher controlled classrooms & labs
   b) Section head controlled discipline wise departments
   c) Principal controlled institutions
   d) Directorate controlled
   e) State & central controlled

(B) Classification According to Functions.
   a) Management & Administration Subsystems.
   b) Goal setting Subsystems.
   c) Course structuring subsystems.
   d) Curricular designing.
   e) Institutional building subsystems.
   f) Teaching, learning & testing evaluation subsystem.
   g) Product evaluating subsystem.
   h) System evaluating subsystem.

4. **Grouping of Subsystem :- It helps in decision taking & evaluation. It includes the following:**
   (A) Planning Subsystem
   (B) Programming Subsystem
   © Processing Subsystem
   (D) Controlling Subsystem

5. Need for total evaluation of the technical education system

The basic objective of evaluation of a system is to assess the worth quality, quantity,

relevance, effectiveness future capabilities, cost benefit analysis. It provides guideline regarding continuing, modifying, terminating, replacing, rating, correcting and changing its outputs, outcomes and accomplishments. Any proposal for evaluation must suggest total evaluation of the system. Evaluation of individual subsystem will provide useful information about their individual functional effectiveness only.

LINEAR PREDICTIVE CODING ALGORITHM WITH ITS APPLICATION TO SOUND SIGNAL COMPRESSION

Anil Dudy

Assistant Professor, Dept. Of ECE, Shri Baba Mastnath Engineering College, Rohtak, Haryana

ABSTRACT

Recently, with the advances in digital signal processing, compression of audio signals has received great attention for speech processing applications. The perceived loudness of a sound signal by human is spectrally masked by background noises. This effect causes a shift of audible sound pressure level. This paper investigates the application of linear predictive coding (LPC) algorithm as compression of recorded sound signal. LP provides parametric modelling techniques which are used to model the spectrum as an autoregressive process for sound signal compression. Linear prediction and its mathematical derivation will be described. Finally, simulation results show the original recorder signal and the compressed signal sinusoidal with different audio signal. Results indicate that the linear predictive coding compresses the signal to a great extent with small distortion of signal.

Index Terms: Linear predictive coding (LPC), Autoregressive process.

I. INTRODUCTION

Uncompressed audio require substantial storage capacity. Data transfer of uncompressed audio data over digital networks requires that very high bandwidth be provided for a single point-to-point communication. To be cost-effective and feasible, multimedia systems must use compressed audio streams. The importance of data compression is not likely to diminish, as a key technology to allow efficient storage and transmission [1]. With recent advances in digital signal processing, compression of sound signals has received great attention in speech signal processing applications [2]–[3]. Different compression methods used for speech signals [4]– [9] were investigated. Here,, either wavelet transforms or predictive coding was used. Other similar application are, electrocardiogram (ECG) [10], and electromyogram (EMG) [11] signals compression. Two families of algorithms that used for the compression of the sound signal namely lossy compression and Lossless compression. Lossy compres-sion [14] means that some data is lost when it is decom-pressed. Lossy compression bases on the assumption that the current data files save more information than human beings can "perceive". Thus the irrelevant data can be re-moved. Lossless compression means that when the data is decompressed, the result is a bit-for-bit perfect match with the original one. The name lossless means "no data is lost", the data is only saved more efficiently in its compressed state, but nothing of it is removed. Here, we are using sub band coding compression technique is used for the compression of the speech signal. Subband coding [12][13] considers the signal only in predefined regions of the spectrum, such as frequency bands. Audio techniques apply Differential Pulse Code Modulation (DPCM) to a sequence of PCM-coded samples. This technique requires a linear characteristic curve for quantization. This paper investigates the application of linear predictive coding (LPC) algorithm as compression of recorded sound signal. LP provides parametric modelling techniques which are used to model the spectrum as an autoregressive process for sound signal compression. Linear prediction and its mathematical derivation will be described. The outline of this paper is as follows. In Section II, we provide the basic of the linear prediction and block diagram is presented for linear predictor. Linear prediction error

function is described in Section III. In Section IV levinson-durbin algorithm is explained with mathematical analysis.. Section V present the audio compression techniques based on linear prediction. Simulation results are presented in section VI. Finally, Section VII presents our conclusions and final comments.

II. LINEAR PREDICTION

A linear prediction (LP) model [4] predicts/forecasts the future values of a signal from a linear combination of its past values. A linear predictor model is an all-pole filter that models the resonance (poles) of the spectral envelope of a signal or a system. LP models are used in diverse areas of applications, such as data forecasting, speech coding, video coding, speech recognition, model-based spectral analysis, model-based signal interpolation, signal restoration, noise reduction, impulse detection, and change detection. In the statistical literature, linear prediction models are often referred to as autoregressive (AR) processes. The all-pole LP model shapes the spectrum of the input signal by transforming an uncorrelated excitation signal to correlated output signal whereas the inverse LP predictor transforms a correlated signal back to an uncorrelated flat-spectrum signal. Inverse LP filter is an all-zero filter, with the zero situated at the same position in pole-zero plot as the poles of the all-pole filter and is also known as a spectral whitening, or de-correlation filter . Poles are at denominator of the polynomial and zeros are at numerator of the polynomial.

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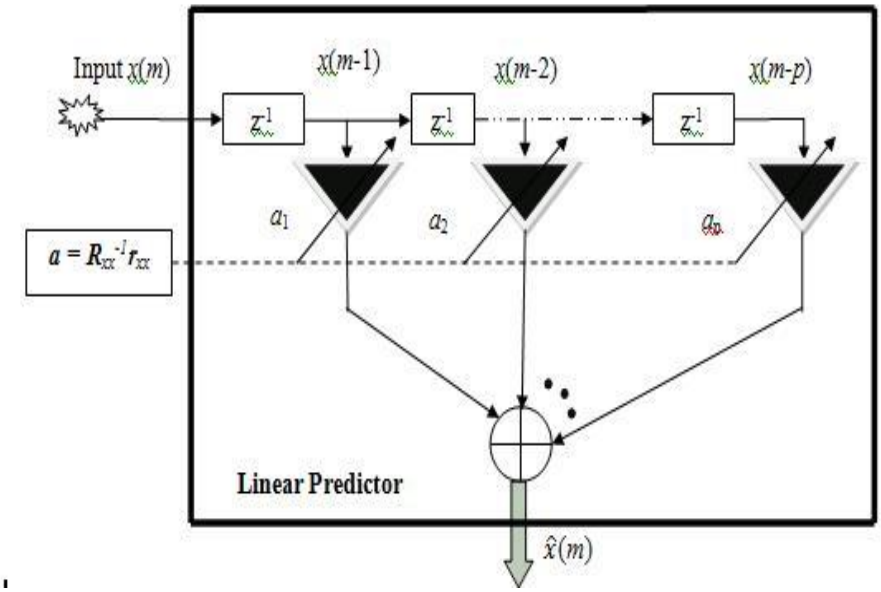


Fig. 1 Diagram of the Linear Predictor

III. LINEAR PREDICTION ERROR

The mean square prediction error becomes zero only if the following three

conditions are satisfied: (a) the signal is deterministic, (b) the signal is correctly modelled by a predictor of order P, and (c) the signal is noise-free. For example, a mixture of $P/2$ sine waves can be modelled by a predictor of order P, with zero prediction error. However, in practice, the prediction error is non-zero because information-bearing signals are random, often only approximately modelled by a linear system, and usually observed in noise. The least mean square prediction error, btained is

$$E^{(P)} = E[e^2(m)] = r_{xx}(0) - \sum_{k=1}^{p} a_k\, r_{xx}(k) \quad (1)$$

Where $E^{(P)}$ denotes the prediction error for a predictor of order P. The prediction error decreases, initially rapidly and then slowly, with the increasing predictor order up to the correct model order. For the correct model order, the signal $e(m)$ is an uncorrelated zero-mean random process with an autocorrelation function defined as

$$E[e(m)\, e(m-k)] = \begin{cases} \sigma_e^2 = G^2 & \text{if } m = k \\ 0 & \text{if } m \neq k \end{cases} \quad (2)$$

Where σ_e^2 is the variance of $e(m)$.

IV. LEVINSON-DURBIN ALGORITHM

The Durbin algorithm starts with a predictor of order zero for which $E^{(0)}=r_{xx}(0)$. The algorithm then computes the coefficients of a predictor of order i, using the coefficients of a predictor of order i-1. In the process of solving for the coefficients of a predictor of order P, the solutions for the predictor coefficients of all orders less than P are also obtained:

$$E^{(0)}=r_{xx}(0) \quad (3)$$

For $i = 1,\ldots, P$

$$\Delta^{(i-1)}=r_{xx}(i)-\sum_{k=1}^{i-1} a_k^{(i-1)} r_{xx}(i-k) \quad (4)$$

$$k_i= -\frac{\Delta^{(i-1)}}{E^{(i-1)}} \quad (5)$$

$$a_i^{(i)} = -\,k_i \quad (6)$$

$$a_j^{(i)} = a_j^{(i-1)} + k_i\, a_{i-j}^{(i-1)} \qquad \text{where } 1 \le j \le i-1 \quad (7)$$

$$E^{(i)}=(1-k_i^2)E^{(i-1)} \quad (8)$$

Audio Compression Based on Linear Prediction

Audio waveforms exhibit a high degree of continuity or sample-to-sample correlation, i.e., the tendency of samples to be similar to their neighbours [3]. The reason is that audio samples are digitized from continuous waveforms, and the sampling rate is usually higher than the rate needed at any particular time. To take advantage of this correlation, prior to the encoding process most audio compressors apply a pre-processing component called a predictor [4]. The predictor can reduce the sample magnitude by making a prediction of the current sample based on the knowledge of some given preceding samples, and then

subtracting the prediction from the current sample value. As a result, the predictor eliminates the correlation inherent in samples before encoding. An audio file in compacted form is comprised of a header and a sequence of frames. The file header contains properties of the audio signal stream. Each frame consists of its own frame overhead and a sequence of residual codes. The frame overhead provides enough information so that the decoding process can start working without the knowledge of other frames. The frame overhead contains the corresponding block size, the prediction model, the residual coding algorithm, and all relevant parameters. Decoding, the reverse process, retrieves information from the compacted form and reconstructs audio samples.

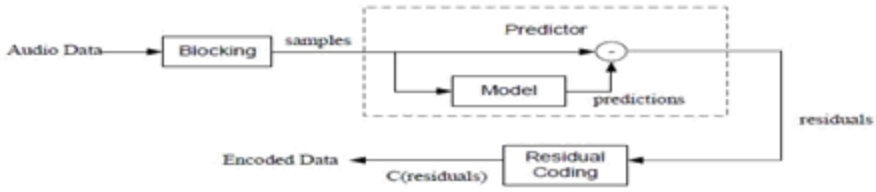


Fig. 2 Encoding Process

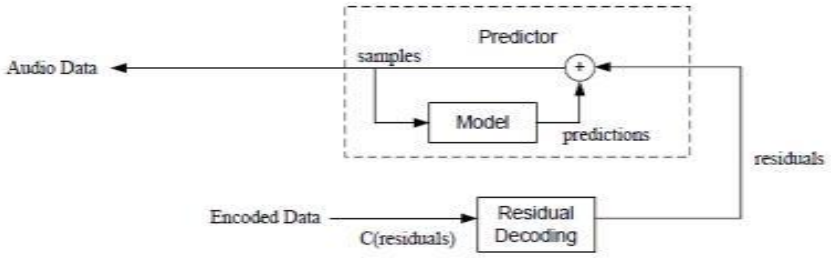


Fig. 3 Decoding Process

V. SIMULATION RESULTS

This section present the audio signal compression is performed using linear predictive coding. Simulation is presented using Matlab.

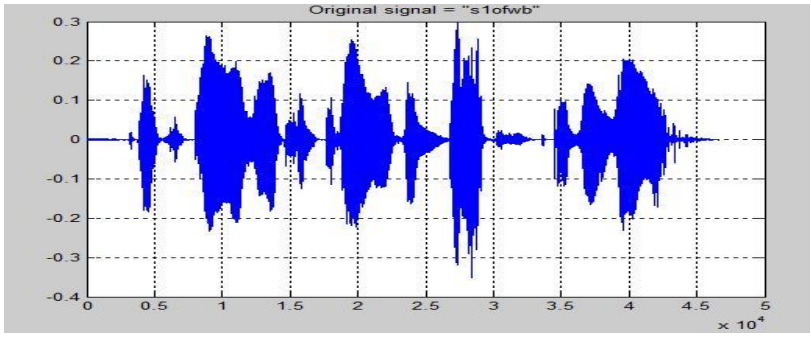


Fig. 4 Original Recorded Speech Signal

We are performing the experiments, with five different recorded voice signals to analysis the level of compression as well as performance variation in sound signal due to compression. The problem of signal compression or source coding is to achieve a low bit rate in the digital representation of an input signal with minimum perceived loss of signal quality. Fig. 4 shows the Original Recorded Speech Signal in different environment condition. Fig. 5 depicts the LPC compressed Speech Signal. Fig. 6 shows the Original Recorded Speech Signal in different environment condition. Fig. 7 depicts the LPC compressed Speech Signal.

Fig. 8 shows the Original Recorded Speech Signal in different environment condition. Fig. 9 depicts the LPC compressed Speech Signal.

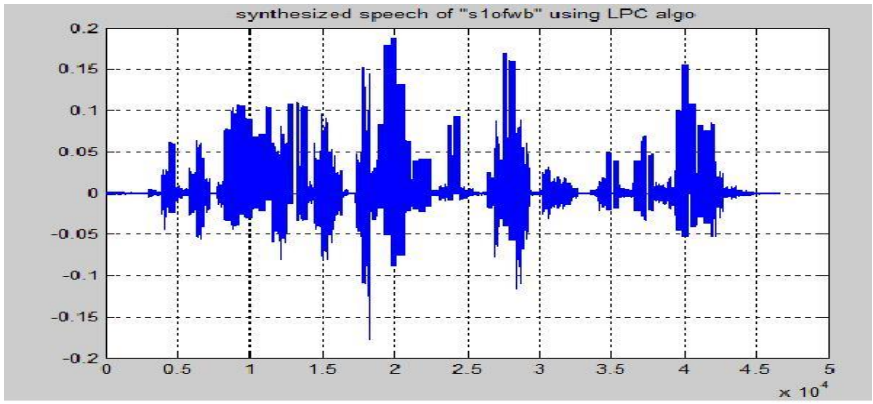


Fig. 5 LPC compressed Speech Signal

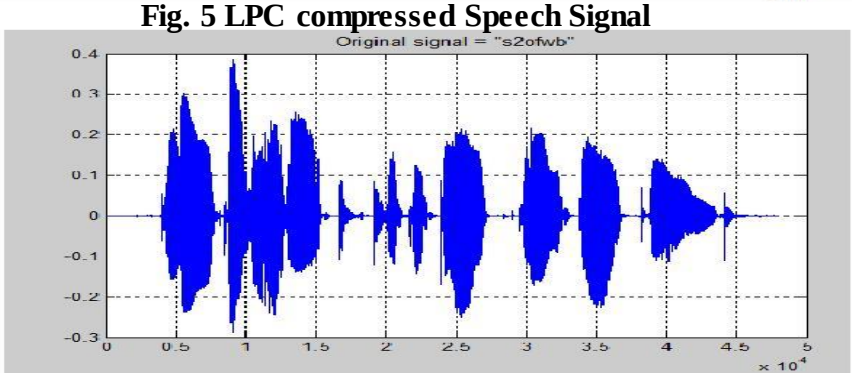


Fig. 6 Original Recorded Speech Signal

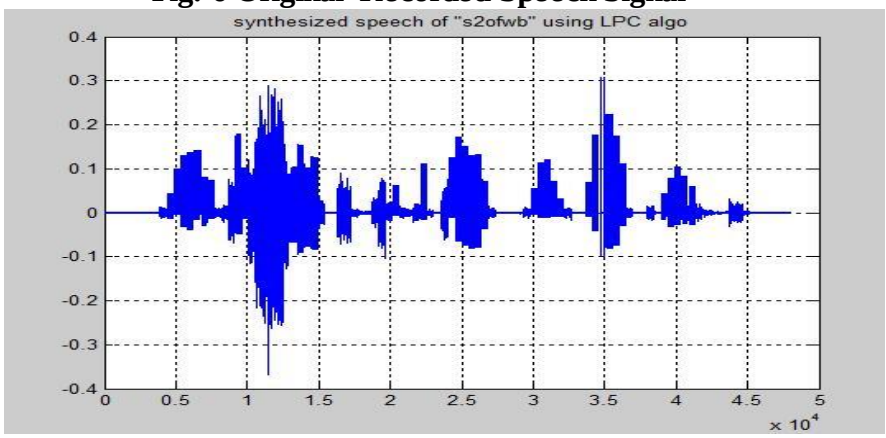


Fig. 7 LPC compressed Speech Signal

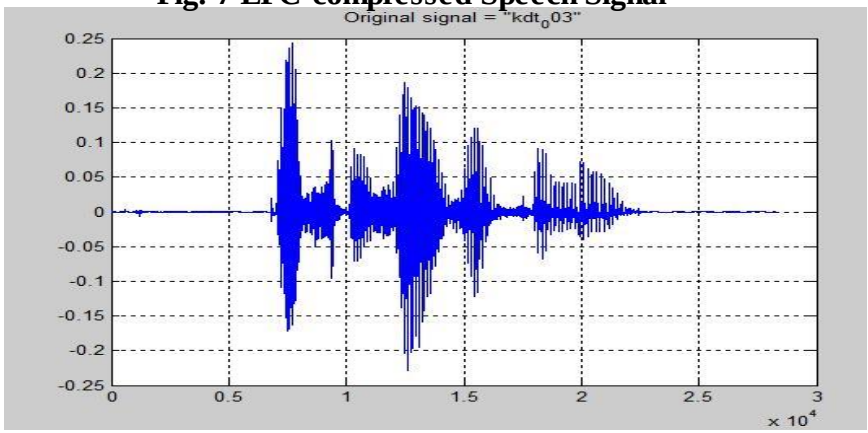


Fig. 8 Original Recorded Speech Signal

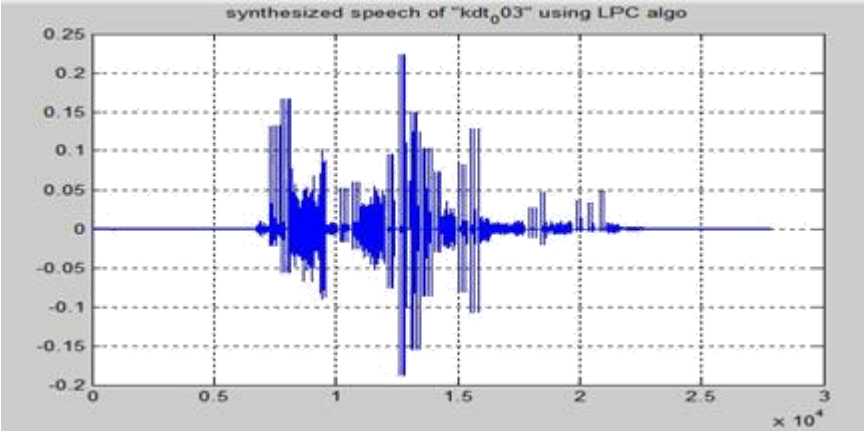


Fig. 9 LPC compressed Speech Signal

VI. CONCLUSIONS

In this paper, a method for compression of sounds signals was investigated. The method is based on linear predictive coding algorithm. The listening of the results of the sound signal and compression signal indicate that the signal is compressed to great extent with less distortion. Results indicate that the linear predictive coding compresses the signal to a great extent with small distortion of signal.

REFERENCES

[1] A. V. Oppenheim R. W. Schafer. Digital Signal Processing. Prentice Hall, Englewood Cliffs, New Jersey, 1975.
[2] T. W. Parsons. Voice and speech processing. McGraw-Hill New York, 1987.
[3] L. R. Rabiner and R. W. Schafer. Digital Processing of Speech Signals. Prentice-Hall Signal Processing Series. Prentice-Hall Inc., Engelwood Cliffs, New Jersey, 1978.
[4] R. Zelinski and P. Noll, "Adaptive transform coding of speech signals," IEEE Trans. Acoust. Speech Signal Process. Vol. ASSP-25, no. 4, pp. 299–309, Aug. 1977.
[5] J. Tribolet and R. Crochiere, "A vocoder-driven adaptation strategy for low bit-rate adaptive transform coding of speech signals," presented at the Int. Conf. Digit. Signal Process. Florence, Italy, 1978.
[6] R. Zelinski and P. Noll, "Approaches to adaptive transform speech coding at low bit rates," IEEE Trans. Acoust. Speech Signal Process. Vol. ASSP-27, no. 1, pp. 89–95, Feb. 1979.
[7] J. Tribolet and R. Crochiere, "Frequency domain coding of speech," IEEE Trans. Acoust. Speech Signal Process. Vol. ASSP-27, no. 5, pp. 512–530, Oct. 1979.
[8] R. Cox and R. Crochiere, "Real-time simulation of adaptive transform coding," IEEE Trans. Acoust. Speech Signal Process. Vol. ASSP-29, no. 2, pp. 147–154, Apr. 1981.
[9] A. Spanias, "Speech coding: A tutorial review," Proc. IEEE, vol. 82, no. 10, pp. 1539–1582, Oct. 1994.
[10] A. Cohen and Y. Zigel, "Compression of multichannel ECG through multichannel long-term prediction," IEEE Eng. Med. Biol. Mag., vol. 17, no. 1, pp. 109–115, Jan.– Feb. 1998
[11] E. Carotti, J. D. Martin, D. Farina, and R. Merletti, "Linear predictive coding of myoelectric signals," in Proc. Int. Conf. Acoust. Speech Signal Process. Philadelphia, PA, 2005, vol. 5, pp. 629–632.
[12] Luneau, J.-M. ; Lebrun, J. ; Jensen, S.H., "Complex Wavelet Modulation Sub bands for Speech Compression", Data Compression Conference, 2009. DCC '09, Page(s): 457, 2009.

LIME-SOIL-FLY ASH BRICKS

Ajay Sharma

*Civil Engineering Department, Shri Baba Mastnath Engineering College, Asthal Bohar, Rohtak, Haryana, India aajay.055@ymail.com

Fly ash is generated in large quantities especially by thermal power plants. A lot of research has been carried out for effective utilization of fly ash in building industry. Use of fly ash in manufacturing brick is one such subject which is being studied by researchers. The aim of the present study is to investigate the strength and water absorption characteristic of fly ash bricks made of lime (L), local soil (S) and fly ash (FA). The experiments were conducted both on Hand moulded and Pressure moulded fly ash bricks. It was observed that none of the L-S-FA bricks satisfy all the requirements of standard codes. While some of the bricks satisfy the provisions in respect of strength only the L-FA (40: 60) bricks satisfy the requirement of Indian Standard Code in respect of strength as well as water absorption characteristics.

Fly ash is finely divided residue resulting from the combustion of powdered coal, transported by the flue gases and collected by electrostatic precipitators. Its proper disposal has been a cause of concern since long, which otherwise leads to pollution of air, soil and water. The disposal and utilization of this fly ash is a matter to ponder. The World Bank has cautioned India that by 2015, land disposal of coal ash would require about 1000 Km2of land. To overcome this problem and to encourage the utilization of fly ash, Government of India in 2003 made it mandatory to use at least 25% fly ash with soils on weight to weight basis for manufacture of bricks within a radius of 100 Km from coal or lignite based thermal power plants. Several researchers and organizations have put forward the methods for use of fly ash in brick making. For last several decades attempts are being made to find a suitable method for the disposal and proper utilization of fly ash.

MATERIALS

LIME (L)

The lime was tested as per the provisions of IS: 6932 -1973. The impurities present in lime were less than 5%. The OMC and MDD were found to be 42.5% and 1080 kg/m3 respectively.

SOIL (S)

The soil available in MNNIT Allahabad Campus was taken and tested as per the provisions contained in IS: 2720 -1983. The specific gravity of the soil was 2.65. In all the samples, the fraction finer than 2 was maintained as 11%. Its liquid limit (LL), plastic limit (PL), and plasticity index (PI) were 30%, 21%, and 9% respectively. The OMC and MDD were 14.5% and 1780 kg/m³. Its unconfined compressive strength was 0.144 N/mm². Fly ash (FA) The fly ash for the present investigation was procured from IFFCO, Phulpur, Allahabad. The specific gravity of fly ash was 2.08. In all the samples, fraction finer than 2was maintained as 7.7%. Its LL, PL, and PI were 15%, 15%, and 0% respectively. The OMC and MDD were 45% and 800 kg/m3 respectively. The chemical composition of fly ash is presented in Table 1.

Chemical Composition of Fly Ash

CHEMICAL COMPOSITION	% BY WEIGHT
UNBURNT CARBON	12.00

SiO2	**57.77**
Al2O3	**23.92**
Fe2O3	**9.56**
TiO2	**1.63**
CaO	**2.24**
K2O	**0.60**
MgO	**1.28**
Mo2O	**0.13**

in lab using the lime, soil and FA in ratios of: 15: 5: 80 and 10: 10: 80; 25: 5: 70 and 20: 10: 70; 35: 5: 60 and 30: 10: 60 respectively. Similarly modular bricks made of L and FA in the ratio of 20: 80, 30: 70 and 40: 60 respectively were cast. The sample was mixed with sufficient quantity of water to obtain working consistency for moulding. The clean mould was filled with the lime fly ash and soil mixture without allowing any air bubble. The surplus mix was removed and top surface was leveled. For the hand moulded bricks no pressure was applied on the mould. The pressure moulded bricks were prepared by applying load of 10, 30 and 50 kN, respectively. The moulded brick were allowed to dry for two days, protecting from direct sun light. The specimens were immersed in water at room temperature for 24 hours and there after, the specimens were taken out of water. These samples were cured by moist jute bags for 7 and 28 days. The samples were tested after 7 and 28 days respectively for compressive strength as per the provisions of IS: 3495 (Part 1)-1992. The water absorption of the bricks was tested as per the provisions contained in IS: 3495 (Part 2)-1992. Before testing, the frogs and voids of the specimen were filled up with cement sand mortar (1: 1).

EXPERIMENTAL RESULTS

The results of the present investigation are presented in Figures 1 to 3 and compared with the Indian Standards of clay and fly ash bricks (IS: 1077-19992 and IS: 12894-2002).

CONCLUSION

1. The compressive strength of bricks increases with lime proportion.
2. The bricks made under pressure has increased compressive strength according as the pressure was applied.
3. As long as the percentage between lime and fly ash is unchanged, the change in soil percentages does not affect compressive strength significantly.
4. Most of the L-S-FA bricks belong to class 3.5 and 5 in respect of strength only. In respect of water absorption all L-S-FA bricks fail.

Only L-FA (40: 60) brick satisfies the criterion of class 3.5 in respect of both strength and water absorption.

BIGGER AND SMARTER LED TVS: THE NEXT LEVEL OF HOME ENTERTAINMENT

Ajay Madan

Assistant Professor S.B.M.N Engineering College, Rohtak

Gone are the days when bulky and affordable televisions ruled typical Indian households. Now sleek and less-space-hogging light-emitting diode (LED) and liquid-crystal display (LCD) TVs are fast replacing box-like bulky television sets. The demand for LED TVs started gaining traction with the prices of LED technology and products coming down. At present, you can buy an LED TV at around Rs 25,000, which is fairly reasonable for a 'good' television. It doesn't end with a good television coming home. Indian consumers now want entertainment in bigger, better and larger than life avatar. One can see an apparent rise in the demand for big-size televisions. Samsung, which enjoys the credit of launching big-screen smart phones, had recently introduced its 190cm (75-inch) LED TV. Obviously, the bigger the size, the higher the price. In fact, LED TVs as big as 254cm (100-inch) are also available in the market. Yes, one cannot deny that such huge televisions are recognized as merely statement products, but there are many takers of such huge TVs as well.

It goes without saying that LED TVs have taken the Indian television market to the next level. After using bulky colour TV for years, people are switching to the sleeker but bigger-screen LED TVs. The trend started with the use of plasma TVs, which later moved onto LCD TVs. In the last couple of years, LED TVs have gained acceptance among the consumers even as many are still having tough time selecting the right one at the right price. Indian consumers are price-conscious, but, of late, the pattern has changed its course with people going for 102cm (40-inch) or larger LED TVs, which cost anywhere between ` 40,000 and ` 350,000, depending on the size and technology loaded. Consumers of these LED TVs are in the age group of 30-45 years, and have an eye for design, aesthetics and nuances of high picture quality.

OLED TV's (Organic light emitting diode)

With LED revolution already in place, the world is getting ready to welcome the future of television with OLED (organic light emitting diode) technology, which was showcased at the Consumer Electronics Show in Las Vegas this year.

"In the future, new technologies such as OLED and bigger screen sizes will come to the market. The market at the moment prefers the 102cm (40-inch) display variant, but this trend could move further up towards the 140cm (55-inch) mark," H.S. Kim, executive vice-president and head of visual and display, Global Business, Samsung, was quoted saying in a report recently.

OLED TVs are still four to five years away, so for the time being you could buy an LED TV. However, there are some basics that you should know before shelling out a big amount on these TVs.

Selection Criterion:

Type of LED TV LED TVs come with LED backlighting. There are two types: One comes with edge backlighting and is super-thin. The other is the full-array or full-LED TV. It is a lot thicker than edge-backlit type, which results in better contrast and image quality. Full-LED TV offers better picture quality than slim version because LEDs are positioned behind the screen in a couple of rows, which results in equal light distribution across the display panel. The super-thin version, on the other hand, has LEDs placed on the outside edge of the panel, which makes a lot of difference to the final picture.

HD or Full HD display? Currently, the market is being spearheaded by High Definition (HD) or 720p resolution. It offers the best quality for normal content like serials, sitcoms, sports action and even movies played from a DVD player. But if you are a gamer who likes to play around with your PlayStation or Xbox, you should definitely go for Full HD or 1080p resolution, which is already present on a few smart phones available across the globe. It will also do justice to any Full HD or Blu Ray movie that you download.

Contrast ratio Contrast ratio is primarily defined as the difference in contrast between the black and white portions of the TV screen. Higher contrast ratio will offer higher image quality. You must try to identify the actual contrast ratio of the TV that you intend to buy.

Design element: While screen thickness does not affect the overall picture quality, LED TVs add some aesthetic element to your living room. Sleek TVs look much better and are easier to mount on a wall.

Smart features Smart TVs were initially known for their integration with the Internet, which was possible via the local Wi-Fi network. But gradually brands like Samsung, LG and Sony have realised that relying solely on access to Facebook, Twitter and YouTube will not offer much resilience in the current scheme of things, which is why they have also introduced new innovative features that vindicate the term 'Smart.' Smart TVs are not just about Internet connectivity anymore, features like motion control and voice recognition are an integral part of LED TVs available in the country.

Less power consumption: In a country like India, where power is a big problem, it is imperative that the television supports the core objective of power saving. Previous technologies like plasma and LCD failed to solve the problem but with LED TVs, the market has found its perfect match. LED TVs consume lesser power than LCD and plasma TVs.

FRICTION WELDING: A WELDING PROCESS HELPS IN AUTOMOTIVE INDUSTRY

Aditya Kaushik[1], Sumit[2]

[1,2]Asst. Prof., S.B.M.N.E.C., Asthal Bohar, Rohtak Email: Adityakaushik15@gmail.com, Engr.sumit1990@gmail.com

INTRODUCTION

Friction welding (FRW) is a solid-state welding process that generates heat through mechanical friction between a moving workpiece and a stationary component, with the addition of a lateral force called "upset" to plastically displace and fuse the materials. Technically, because no melt occurs, friction welding is not actually a welding process in the traditional sense, but a forging technique. However, due to the similarities between these techniques and traditional welding, the term has become common. Friction welding is used with metals and thermoplastics in a wide variety of aviation and automotive applications. (In friction welding, the heat is obtained by the friction between two parts which are to be joined. These parts are held under pressure where one part is stationary and other part is made to rotate at high speed. The welded joints is obtained when a force on the stationary part is applied after stopping the rotation of the part to get the welded joint)

The combination of fast joining times (on the order of a few seconds), and direct heat input at the weld interface, yields relatively small heat-affected zones. Friction welding techniques are generally melt-free, which avoids grain growth in engineered materials, such as high-strength heat-treated steels. Another advantage is that the motion tends to "clean" the surface between the materials being welded, which means they can be joined with less preparation. During the welding process, depending on the method being used, small pieces of the plastic or metal will be forced out of the working mass (flash). It is believed that the flash carries away debris and dirt

Another advantage of friction welding is that it allows dissimilar materials to be joined. This is particularly useful in aerospace, where it is used to join lightweight aluminum stock to high-strength steels. Normally the wide difference in melting points of the two materials would make it impossible to weld using traditional techniques, and would require some sort of mechanical connection. Friction welding provides a "full strength" bond with no additional weight. Other common uses for these sorts of bi-metal joins is in the nuclear industry, where copper-steel joints are common in the reactor cooling systems; and in the transport of cryogenic fluids, where friction welding has been used to join aluminum alloys to stainless steels and high-nickel-alloy materials for cryogenic-fluid piping and containment vessels.

Friction welding is also used with thermoplastics, which act in a fashion analogous to metals under heat and pressure. The heat and pressure used on these materials is much lower than metals, but the technique can be used to join metals to plastics with the metal interface being machined. For instance, the technique can be used to join eyeglass frames to the pins in their hinges. The lower energies and pressures used allows for a wider variety of techniques to be used.

METAL TECHNIQUES

Spin welding

Spin welding systems consist of two chucks for holding the materials to be welded, one of which is fixed and the other rotating. Before welding one of the work pieces is attached to the rotating chuck along with a flywheel of a given weight. The piece is then spun up to a high rate of rotation to store the required energy in the flywheel. Once spinning at the proper speed, the motor is removed and the pieces forced together under pressure. The force is kept on the pieces after the spinning stops to allow the weld to "set". This technique is also known as inertia welding, rotational welding or inertial friction welding.

Linear friction welding

Linear friction welding (LFW) is similar to spin welding except that the moving chuck oscillates laterally instead of spinning. The speeds are much lower in general, which requires the pieces to be kept under pressure at all times. This also requires the parts to have a high shear strength. Linear friction welding requires more complex machinery than spin welding, but has the advantage that parts of any shape can be joined, as opposed to parts with a circular meeting point. Another advantage is that in most instances quality of joint is better than that obtained using rotating technique.

Friction surfacing

Friction surfacing is a process derived from friction welding where a coating material is applied to a substrate. A rod composed of the coating material (called a mechtrode) is rotated under pressure, generating a plasticised layer in the rod at the interface with the substrate. By moving a substrate across the face of the rotating rod a plasticised layer is deposited between 0.2–2.5 millimetres (0.0079–0.0984 in) thick depending on mechtrode diameter and coating material.

QUALITY FACTOR EVALUATION OF STIMULATED RAMAN SCATTERING (SRS) AND FOUR-WAVE MIXING (FWM) IN PASSIVE OPTICAL NETWORKS

Jitender Khurana

Professor, Dept. Of ECE, Shri Baba Mastnath Engineer-ing College, Rohtak, Haryana

ABSTRACT

In this paper, we presents the quality factor eval-uation of stimulated Raman scattering (SRS) and fourwave mixing (FWM) in optical networks. The nonlinearity ef-fect of stimulated Raman scattering (SRS) and four wave mixing (FWM) with 16-channel optical system. Two paramet-ers are used to evaluate the performance of the optical passive networks. These parameters include transmitting power and quality factors. Simulation result shows are performed based on Q-factor of the optical system and based on transmitting power of the optical system. Simulation results indicate that the quality factor of the optical system goes on decreasing with the addition of nonlinearities of optical system. Results also reveals that the quality factor of the optical system increase for a limited value of transmitting power but after a particular level this goes on decreasing.

Keywords: Stimulated Raman scattering (SRS), four wave mixing (FWM), Q-factor, passive optical Networks.

I. INTRODUCTION

Optical communication based on Wavelength-Division multiplexing (WDM) [1]-[9] has become the key technology to enable the very high capacity networks required by our communication thirsty society. WDM systems dominate long-haul and ultra-long-haul networks due to performance and cost advantages. To quench the rapidly increasing capacity requirement for further progress of information technology, WDM networks with narrower channel spacing are being used [4]. As a result, the dominant nonlinearities become more and more pronounced which puts a challenge to system design engineers. Also, the desired increase in launched power in order to expand the WDM network is limited by these nonlinearities. Among the nonlinearities known to limit the throughput of WDM system, four-wave mixing (FWM), cross-phase modulation (XPM) and stimulated Raman scattering (SRS) are the dominant effects [1] [5].

SRS is significant when there are a number of signals on different wavelengths and it induces power transfer from the shorter wavelength channels to longer wavelength channels leading to power penalty in the shorter wavelength channels [1][5]. FWM acts as crosstalk between channels as it results in the mixing of two signals at different frequencies, which leads to the generation of "sum and difference" frequencies [10]-[15]. XPM leads to phase change of one channel according to power of the other channels and the presence of group velocity dispersion (GVD) transforms this phase-modulation (PM) into intensity-modulation (IM) [1][5]. This PM-IM conversion results in XPM acting like crosstalk between channels leading to deterioration of the signal quality. In long haul WDM systems, Erbium-Doped fiber amplifiers (EDFAs) are used to compensate for signal attenuation, thus allowing high data rate transmission over a long distance. The high optical power level available from EDFAs though, leaves the system performance more vulnerable to various nonlinear effects [3]. Also, ASE noise of all the amplifiers accumulates at the receiver and results in degradation of system performance. In this paper, we presents the quality factor evaluation of stimulated Raman scattering (SRS) and four wave mixing (FWM) in optical networks. The nonlinearity effect of stimulated Raman scattering (SRS) and four wave mixing (FWM) with 16-channel optical system. Two parameters are used to evaluate the performance of the optical passive networks. These parameters include transmitting power and quality factors.

Simulation result shows are performed based on Q-factor of the optical system and based on transmitting power of the optical system. The rest of the paper is organized as follows: In section II, explain the basic four wave mixing (FWM) in optical networks.. In Section III, detail of the stimulated Raman scattering (SRS) is given and how this affects the performance of optical system. In Section V, shows the simulation results of optical system in the presence of the nonlinearities such as four waves mixing (FWM) and stimulated Raman scattering (SRS). Finally, a conclusion is made.

II. FOUR-WAVE MIXING (FWM)

Four-wave mixing [2] is an intermodulation phenomenon in non-linear optics, whereby interactions between two wavelengths produce two extra wavelengths in the signal. It is similar to the third-order intercept point in electrical systems. Four-wave mixing is a nonlinear effect arising from a third-order optical nonlinearity, as is described with a $X(3)$ coefficient. It can occur if at least two different frequency components propagate together in a nonlinear medium such as an optical fiber. Assuming just two input frequency components $f1$ and $f2$ (with $f2 > f1$), a refractive index modulation at the difference frequency occurs, which creates two additional frequency components. In effect, two new frequency components are generated: $f3 = f1 - (f2 - f1) = 2f1 – f2$ and $f4 = f2 + (f2 - f1) = 2f2 - f1$. Furthermore, a pre-existing wave the frequency $f3$ or $f4$ can be amplified.

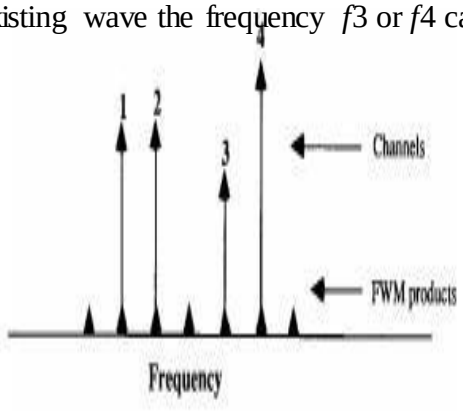


Fig. 1 Effect of four wave mixing

The simplest picture of the four-wave mixing process in fibers can be illustrated by the transmission and cross-phase modulation of four equally spaced channels shown in Fig. 1 Channels 1 and 2 interfere, producing an index of refraction which oscillates at the difference frequency. This modulation in refractive index modulates channel 4, producing sidebands at channels 3 and 5. This is only the simplest combination of frequencies. Four-wave mixing allows any combination of three frequencies beating together to produce a fourth. If the fourth frequency lies within a communication band, that channel can be rendered unusable. This channel interference can affect either closely spaced channels, as one encounters with coherent communications, or the rather widely separated channels of a WDM system. Efficient four-wave mixing requires phase matching of the interacting waves throughout the interaction length widely separated channels will therefore be phase matched only in a region of low-fiber dispersion.

III. STIMULATED RAMAN SCATTERING (SRS)

SRS is due to interaction of incident light wave with vibrational modes of silica molecule i.e., if two or more optical signals at different wavelengths are injected into a fiber, SRS causes energy from lower wavelength channels to be transferred to the higher

wavelength channels [1][2]. This in turn reduces the signal-to-noise ratio of the lower wavelength channels and introduces crosstalk on higher wavelength channels which can lower the information carrying capacity of the system. The threshold power in case of SRS can be estimated as

$$P_{th} \approx 16 A_{eff} / g_R L_{eff} \quad (1)$$

where g_R is the Raman gain and L_{eff} the effective length of the fiber. If the fiber is sufficiently long, then $L_{eff} \cong 1/\alpha$. In that case

$$P_{th} \approx 16 A_{eff} \alpha / g_R \quad (2)$$

The value of g_R is 1 x 10^{-13} m/W for silica at λ =1550 nm.

The value of α as 0.2 dB/km and A_{eff} as 55 2 μm, results in P_{th} equal to 570 mW. Hence, the effect of SRS is insignificant in case of single channel. In WDM systems, where there are number of signals on different wavelengths, the effect of SRS is significant and it results in transfer of energy from signal at shorter wavelength to signal at longer wavelength. In case of N equally spaced channels with frequency separation between adjacent channels f Hz, assuming scrambled polarization and Raman gain g_R to be linear, the power loss due to SRS by the shortest wavelength (i.e., first) channel is given by

where m_i is the modulation of i^{th} channel, P_i the power injected in i^{th} channel in watts and γ_i the Raman gain coefficient coupling the i^{th} channel and first channel.

Assuming the Raman gain profile to be triangular, γ_i is given by

Where γ_p is the peak Raman gain coefficient. Since m_i are independent random variables, power depletion D converges to a Gaussian random variable. Equally spaced channels with intensity-modulated/direct-detection (IM/DD) system have been assumed for analysis. It is observed that the presence of SRS along with FWM further degrades the network performance. This is due to the fact that the power received at the receiver due to SRS is lower than the power received at the receiver when SRS is not present.

IV. SIMULATION RESULTS

In this section, we present the Q-factor and Transmitted Power Analysis with 16 channel optical communication System. Simulations are performed using MATLAB 2012a. Fig.2 shows the Q-factor analysis with time in dB for FWM-XPM nonlinear effect for 16-channel optical communication System. Fig.3 shows the Q-factor analysis with time in dB for SRS-FWM nonlinear effect for 16-channel optical communication System. Fig.4 shows the Q-factor analysis with time in dB for SRS-XPM nonlinear effect for 16-channel optical communication System. Fig.5 shows the Q-factor analysis with time in dB for SRS-XPM-FWM-ASE nonlinear effect for 16-channel optical communication System. Fig.6 shows the Q-factor analysis with time in dB for different combination of nonlinear effect for 16-channel optical communication System.

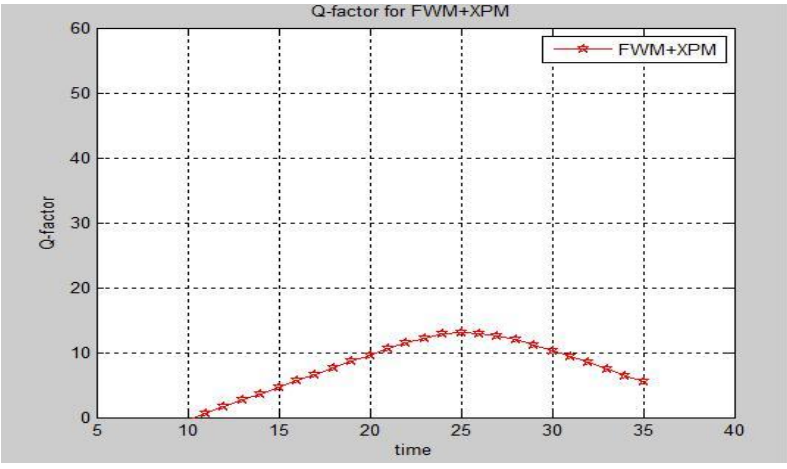


Fig.11 shows the Q-factor analysis with changing transmitted power for different combination of nonlinear effect for 16-channel optical communication System.

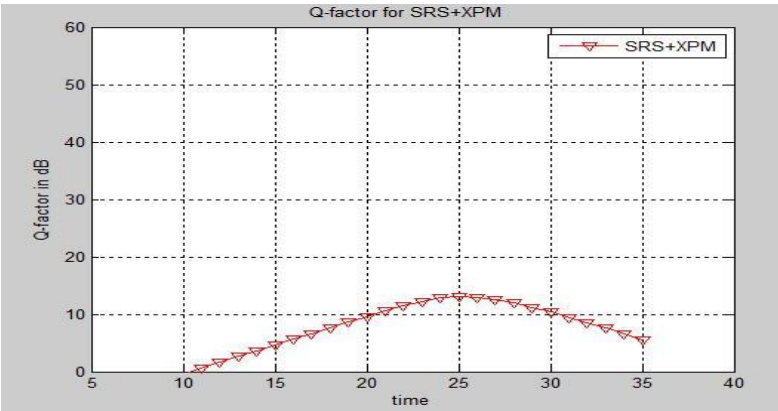


Fig. 2 Q-factor analysis FWM-XPM nonlinear effect for 16-channel optical communication System

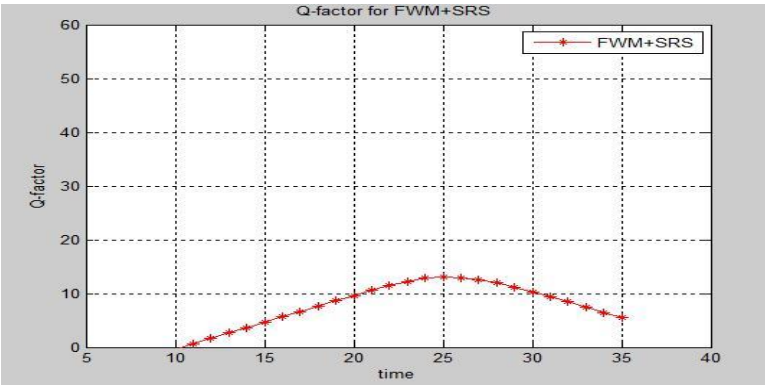


Fig. 3 Q-factor analysis FWM-SRS nonlinear effect for 16-channel optical communication System

Fig.7 shows the Q-factor analysis with changing transmitted power for FWM-XPM nonlinear effect for 16-channel optical communication System. Fig.8 shows the Q-factor analysis with changing transmitted power for SRS-FWM nonlinear effect for 16-channel optical communication System. Fig.9 shows the Q-factor analysis with changing transmitted power for SRS-XPM nonlinear effect for 16-channel optical communication System. Fig.10 shows the Q-factor analysis with changing transmitted power for SRS-XPM-FWM-ASE nonlinear effect for 16-channel optical communication System.

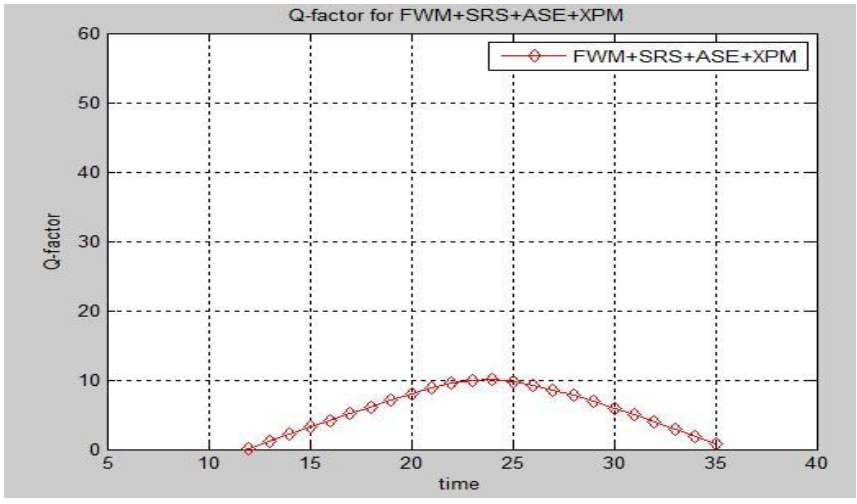


Fig. 4 Q-factor analysis SRS-XPM nonlinear effect for 16-channel optical communication System

Fig. 5 Q-factor analysis FWM-SRS-ASE-XPM nonlinear effect for 16-channel optical communication System

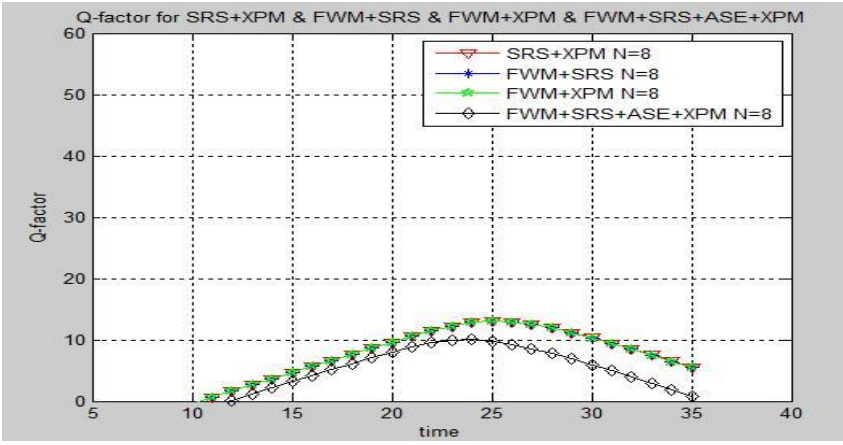


Fig. 6 Combined Q-factor analysis for different nonlinear and linear effect for 16-channel optical communication System

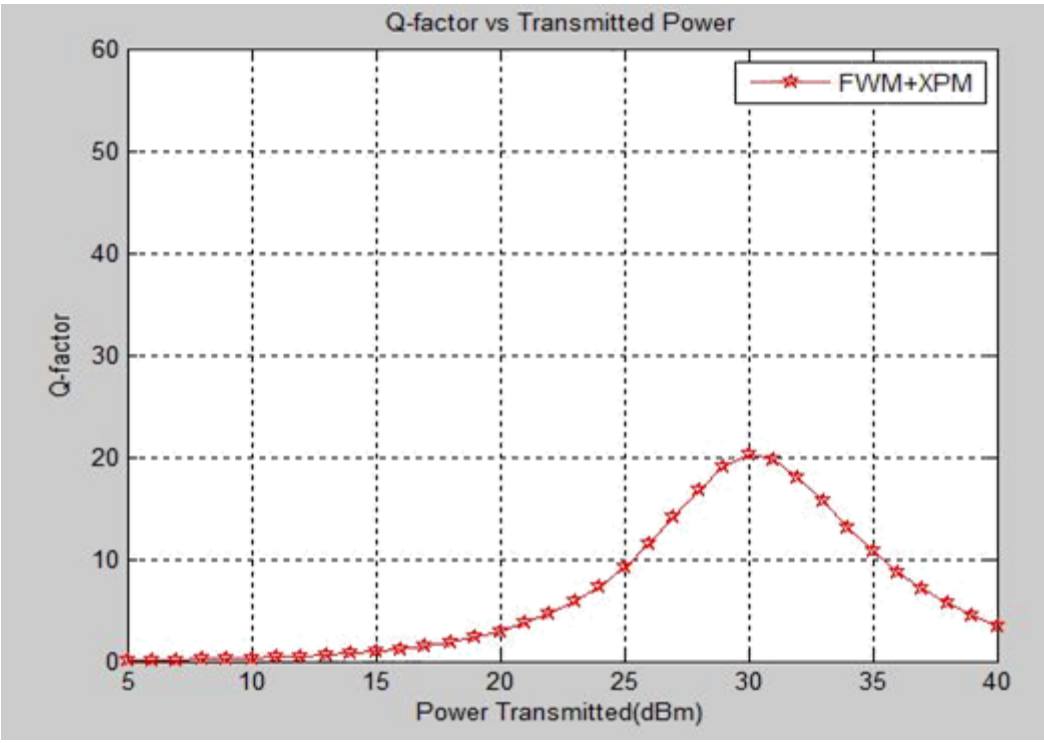


Fig. 7 Transmitted Power Analysis FWM-XPM nonlinear effect for 16-channel optical communication System

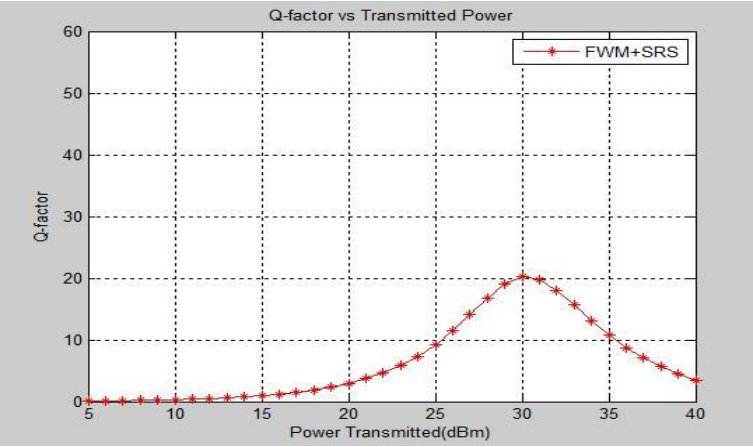


Fig. 8 Transmitted Power Analysis FWM-SRS nonlinear effect for 16-channel optical communication System

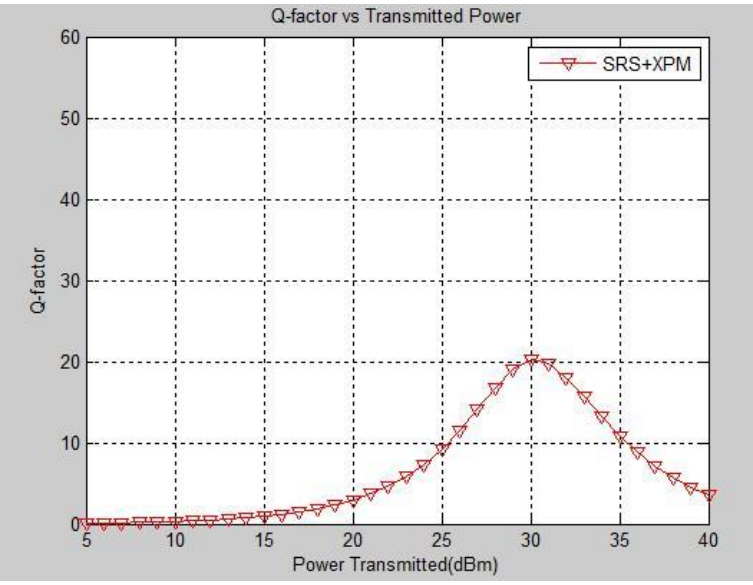


Fig. 9 Transmitted Power Analysis SRS-XPM nonlinear effect for 16-channel optical communication System

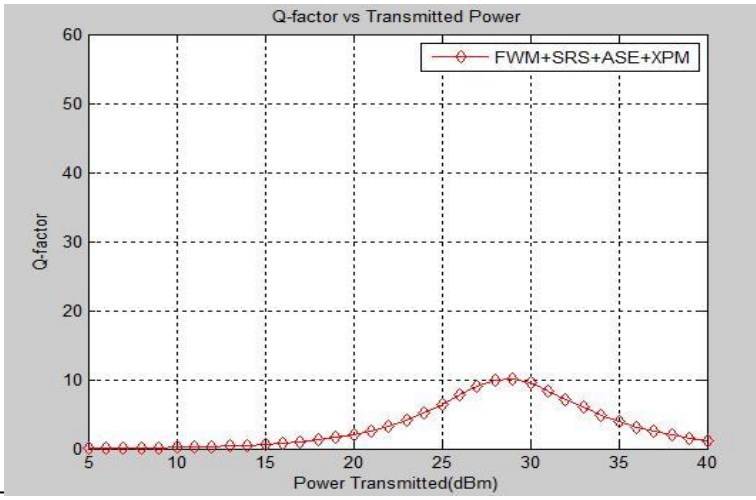


Fig. 10 Transmitted Power Analysis FWM-SRS-ASE-XPM nonlinear and linear effect for 16-channel optical communication System

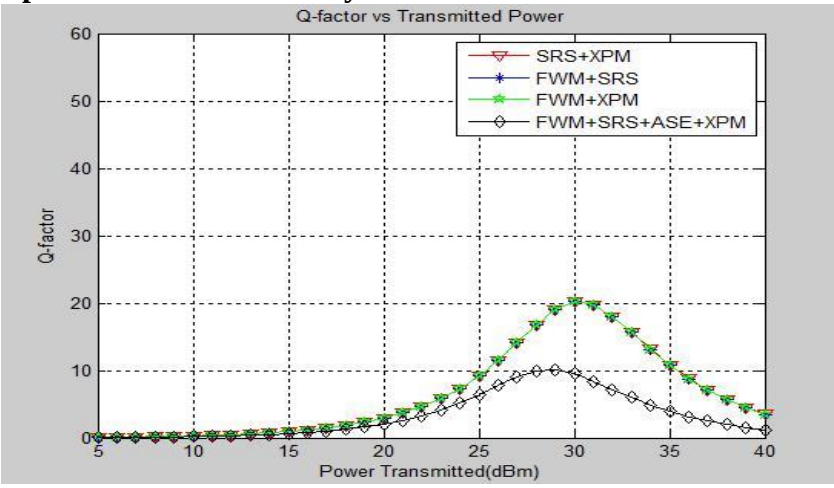


Fig. 11 Combined Transmitted Power Analysis for different nonlinear and linear effect for 16-channel optical communication System

V. CONCLUSIONS

In this paper, we presents the quality factor evaluation of stimulated Raman scattering (SRS) and four wave mixing (FWM) in optical networks. The nonlinearity effect of stimulated Raman scattering (SRS) and four wave mixing (FWM) with 16-channel optical system. Two parameters are used to evaluate the performance of the optical passive networks. These parameters include transmitting power and quality factors. Simulation result shows are performed based on Q-factor of the optical system and based on transmitting power of the optical system. Results also reveals that the quality factor of the optical system increase for a limited value of transmitting power but after a particular level this goes on decreasing.

REFERENCES

[10] G. P. Agrawal, Nonlinear Fiber Optics. 4th edition. Academic Press, San Diego, USA, 2001.

[11] Gerd Kaiser, Optical Fiber Communications, 4th edition. Tata McGraw-Hill, 2008.

[13] G. P. Agarwal, Fiber optics communication systems, 3rd edition, John Wiley & Sons, 2002.

[14] F. Forghieri, R. W. Tkach and A. R. Chraplyvy, "Fiber nonlinearities and their impact on transmission systems", Optical Fiber Telecommunications, I. P. Kaminov and T. L. Koch, Eds. San Diego, CA: Academic, 1997, vol. IIIA, ch. 8.pp. 196-264, 1997.

[15] Mukherjee B. 1997. Optical Communication Networks. New York: McGraw Hill. Pp1-10.

[16] Mukherjee B. 1992. WDM-based local light wave networks part-1: Single-hop systems. IEEE Network. 7(5) :12.

[17] Chraplyvy A. and J. Stone. 1984. Measurement of crossphase modulation in coherent wavelength division multiplexing using injection laser. Electronics Letters. 20(11): 996

[18] Mingchia Wu and Winston I. Way, "Fiber Nonlinearity Limitations in Ultra-Dense WDM Systems", Journal of Light wave Technology, vol. 22, no. 6, Jun 2004.

[19] Bo Xu and Maïté Brandt-Pearce, "Comparison of FWM-and XPM-Induced Crosstalk Using the Volterra Series Transfer Function Method", Journal of Light wave Technology, vol. 21, no. 1, pp. 40-53, Jan 2003.
Journal of Light wave Technology, Volume: 24 , Issue: 5 , Page(s): 1955 – 1967, 2006.

[10] Sankawa, Izumi ; Yamamoto, F. ; Okumura, Y. ; Ogura, Y., "Cost and quantity analysis of passive double-staroptical-access-network facilities for broadband service multiplexing", Journal of Light wave Technology, Vol: 24 , Issue: 10 , Page: 3625 – 3634, 2006.

AN APPROACH TO ENHANCE THE RELIABILITY OF THE INDUSTRIAL SYSTEM

Pooja Budhiraja and Vinod Kumar

Department of Mathematics, B. M. University, Rohtak

Science and technology have been progressing at a dumfounding rapid rhythm. The dynamic role of science and technology in modern life is not overstated in view of today's world as they have profoundly influenced our living style. Our life will be paralyzed or come to a halt without physical and abstract conceptual systems that have been developed over the years such as transportation system, communication system, power generation and control system, oil or water purification system etc. In today's scientific world, people depend on the advancement whether in the house or as part of a large industry. To achieve the maximum benefit of advancement in technology improved, trustworthy and reliable systems are in great demand. But improvement in the systems should not increase the cost as it affects and restricts the number of users. In addition, analysis of systems in terms of the reliability and cost effectiveness helps to formulate proper policies to produce and provide the cost effective, efficient, user friendly systems. Therefore the study of reliability and cost effectiveness of the system has become not only inevitable but also of primal significance to the stakeholders.

The probabilistic theory of reliability has grown out of the demand of modern technology and particularly out of the experience of the World War-II with complex military system. Due to the complexity and automation of equipments used during the war resulted in several problems of maintenance and repair. It is in the early 1950's that certain areas of reliability, especially life testing and electronics and missile reliability problems started to receive a great deal of attention from mathematicians and engineers. A commercial organization, Aeronautical Radio, Inc. (ARINC), was set up the airlines of Ohio in 1950 which among several functions, collected and analysed defective tubes and resulted in a remarkable success in improving the reliability of different types of tubes. In 1952, the U.S. department of defence established the Advisory Group on reliability of Electronic Equipment (AGREE) and got its first report published in 1957. Epstein (1958) worked in the field of life testing with assumption of exponential distribution. After this the studies were carried out to evaluate other measures besides finding reliability of the system.

The concept of availability is widely discussed in literature and the main contributors are Barlow and Hunter (1960), Graver (1963), Sandler (1963), Myers et al. (1964), Barlow and Proschan (1965), Rau (1970). Nakagawa (1976**)** considered the replacement of the unit at a certain level of damage while Arora (1977), Mine and Kaiwal (1979) enhanced the system reliability by assigning priority repair discipline.

Reliability of a system can be improved by using superior components with low failure rates which will require more time and money for development. If the cost of producing highly reliable components is very high, then the system reliability can be improved by introducing the technique of redundancy. Nakagawa (1980) studied an inspection policy for a standby electric generator as an example. Yamashem (1980) worked on a multi state system with several failure modes and cold standby unit. Goel, et al. (1984) studied a two unit cold standby system with two types of repair facilities.

Some other aspects applicable to systems were also taken by authors. Kumar, et al. (1996) carried out a comparative study of the profit of a two server system including patience

time and instructions. Taneja and Gupta (1999) studied a reliability model for two–unit cold standby system with instructions and two types of repair. Tuteja, et.al (2001) studied reliability and profit analysis of two-unit cold standby system with partial failure and two types of repairman. Reliability and profit analysis of some systems with different types of repairman viz. ordinary and expert repairman, regular and visiting repairman are studied by Rizwan and Taneja (2001), Sindhu and Gupta (2002) respectively.

Some authors have carried out analysis of some systems on the basis of collected real data. Tuteja, et al. (2006) discussed reliability and profit of a two-out of-three unit system particularly for the case of ash handling plant consisting of three ash water pumps. Taneja, et al. (2006) presented an economic analysis of a model on the programmable logic controller (PLC) cold standby system. There are many important and versatile systems working under different condition/situations that are still need to be analysed in terms of their reliability and costs.

We now discuss some basic concepts related to the reliability:

RELIABILITY

Reliability of a unit (or a product) is the probability that it will give satisfactory performance for specified period under specified operating conditions.

Quantitatively, reliability of a device in time t is the probability that it will not fail in a given environment before time t. If T is a random variable representing the time till the failure of the device starting with an initial operable condition at t = 0, then reliability R (t) of device is given by

$$R(t) = P[T > t] = 1 - P[T \leq t] = 1 - F(t).$$

Thus, reliability is always a function of time. It also depends on environmental conditions which may or may not vary with time. Following assumptions were made:

(i) $R(0) = 1$ since the device is assumed to be operable at $t = 0$.

(ii) $R(\infty) = 0$ since no device can work forever without failure.

(iii) $R(t)$ is non-increasing function between 0 and 1.

INSTANTANEOUS HAZARD RATE (OR FAILURE RATE)

It is defined as the conditional probability that the system fails during the time interval $(t, t+\Delta t]$ given that it was operating during $(0, t]$. Let $r(t)\Delta t$ be the probability that the device has life time between t and $t+\Delta t$, given that it has functioned up to time t.

$$= Pr[\, t < T \leq t + \Delta t \mid T > t]$$

$$= P[t < T \leq t + \Delta t] / P[T > t]$$

$$= P[T < t + \Delta t] - P[T < t] / P[T > t]$$

$$= \frac{\left[1 - R(t + \Delta t)\right] - \left[1 - R(t)\right]}{R(t)}$$

$$= -\frac{R(t + \Delta t) - R(t)}{R(t)}$$

Now the instantaneous failure rate or hazard rate r(t) at time t is defined by

$$r(t) = \lim_{\Delta \to 0} - \frac{R(t+\Delta t) - R(t)}{R(t)\Delta(t)} = \frac{-R'(t)}{R(t)} = \frac{f(t)}{R(t)}$$

where f(t) is defined as the p.d.f. of the device life time. If F(t) is the c.d.f. of failure times, we have the following relations:

$$\overline{F\ t} = \int_t^{\infty} f\ u\ du = R\ t\ = e^{-\int_0^t r\ u\ du}$$

and

$$f\ t\ = r\ t\ e^{-\int_0^t r\ u\ du}$$

STOCHASTIC PROCESS

In probability theory, a stochastic process, or sometimes random process (widely used) is a collection of random variables, representing the evolution of some system of random values over time. This is the probabilistic counterpart to a deterministic process (or deterministic system). We can also say families of random variables which are functions of say, time are known as stochastic processes. The state space of a stochastic process is a list of possible outcomes at any time point. At every time point, the process is in exactly one of its possible states. Thus a stochastic process has a state space and a time structure.

If X denotes the state of the system at time t_0 and the set E describes certain collection of states in the system, then the probability that the system which is at time 't_0' is in the state X will pass into one of the states of E at time t, is denoted as $P\{t_0; t, E\}$. Then the process of changing from one state to another with time having the probability $P\{t_0; t, E\}$ is known as stochastic process.

MARKOV PROCESS

A stochastic process is said to be Markov process if the future state is completely determined by the present state and is independent of the way in which the present state has developed. In this process, the state at time t_n is only influenced by the state of the process at time t_{n-1}.

Thus a sequence of trials X_1, X_2, X_3, ..., is said to constitute a Markov process if the following properties are satisfy.

(i) Each outcome belongs to a finite set of outcomes $\{E_1, E_2, ..., E_m\}$, known as state space of the system and the outcome E_j at the j^{th} trial is defined as the state E_j of the system at time j.

(ii) The outcome of any trial depend at most upon the outcome of the immediately preceding trial and not upon any previous outcomes, i.e. with each pair of states (E_i, E_j), there correspond a probability p_{ij} that outcome E_j happens immediately after E_i. The values of p_{ij} for different combinations of i, j = 1, 2 ,....., m are known as 'transition probabilities' in a single step. These can be arranged in a matrix form, known as single step transition matrix P.

(iii) The process has a set of initial probabilities.

	Initial state	Final state after one step				
		1	2	3	– –	m
	1	p_{11}	p_{12}	p_{13}	– –	p_{1m}
	2	p_{21}	p_{22}	p_{23}	– –	p_{2m}
P =	3	p_{31}	p_{32}	p_{33}	– –	p_{3m}
	–	–	–	–	– –	–
	m	p_{m1}	p_{m2}	p_{m3}	– –	p_{mm}

RENEWAL PROCESS

Suppose we have a repairable system which starts operation at t = 0. If X_1 denotes the time to first failure and Y_1 denotes the time from first failure to next system operation (after repair) then $t_1 = X_1+Y_1$ denotes the time of first renewal. Similarly, if X_2 denotes the time from first renewal to second failure and Y_2 denotes the time from second failure to second renewal then $t_2 = X_2 +Y_2$ and the time of second renewal is $t_1 + t_2$.

In general $t_i = X_i +Y_i$ (inter arrival time between the $(i-1)^{th}$ and i^{th} renewal) for i = 1, 2, ..., if we define $S_0 = 0$, $S_n = t_1 + t_2 + t_3 + ... + t_n$

= epoch of the nth renewal

and N(t) = number of renewal during (0, t] then the process N(t), $t \geq 0$ is called renewal process.

SEMI-MARKOV PROCESS

In the above, assume that the process is time homogeneous, i.e.

$P\{X_{n+1} = j,\ t_{n-1} - t_n \leq t / X_n = i\} = Q_{ij}$, i, j∈S is independent of n , then there exists limiting transition probabilities

$P_{ij} = \lim_{t\to\infty} Q_{ij}(t) = P\{X_{n+1} = j / X_n = i\}$.

Then $\{X_n, n = 0, 1, 2,.....\}$ constitute a Markov chain with state space E and transition probability matrix (t. p .m.) is given by $P = (p_{ij})$.

The continuous parameter Stochastic process Y(t) with state space E defined by

$Y(t) = X_n,\ t_n < t < t_{n+1}$

is called a semi-Markov process.

In other words, we define the semi-Markov process as a process in which the transition from one state to another is governed by the transition probabilities of a Markov process but the time spent in each state before the transition occurs is a random variable depending upon the last transition made. Thus at transition instants the semi-Markov process behaves just like a Markov process. However the time at which transition occurs are governed by a different probability mechanism.

REGENERATIVE PROCESS

A time point at which the system history prior to the time point is irrelevant to the system conditions is called a regenerative point. Regenerative Stochastic processes were introduced by Smith (1955) and has been crucial in the analysis of complex system. Let X(t) be the state of the system at epochs. If t_1, t_2, ... are the epochs at which the process probabilistically restarts,

then these epochs are called regenerative epochs and the process {X(t), t = t_1, t_2, ... } is called regenerative process.

MEAN SOJOURN TIME

The expected time taken by the system in a particular state before transiting to any other state is known as mean sojourn time or mean survival time in that state. If T_i be the sojourn time in state i, then mean sojourn time in state i is

$$\mu_i = \int_0^\infty P(T_i > t)dt$$

MEAN TIME TO SYSTEM FAILURE

Components and systems (simple or complex) do not operate in the same manner in all condition. They cannot operate for an infinitely long time due to aging of components or some other reasons (initial failure due to manufacturing defects, random failures due to change in working stresses or environment conditions, etc.). To avoid the sudden failure, one must be interested in measure representing the life time of the system. This measure is aptly described as the Mean Time to System Failure (MTSF), as it corresponds to the average duration between successive system failures. This measure is defined as the expected time for which the system is in operation before it completely fails.

Suppose 'T' be life time of the system then the reliability function for the system is given by R(t) = 1- F(t), where F(t) is the failure time distribution function and $f(t) = \frac{dF(t)}{dt}$ is the failure time density function. The mean time to system failure is given by

$$MTSF = \int_0^\infty t f(t)\, dt$$

$$= -\int_0^\infty t\left(\frac{dR(t)}{dt}\right)dt$$

$$= \int_0^\infty R(t)\, dt = \lim_{s\to 0} R^*(s)$$

where $R^*(s)$ is the Laplace transform of the reliability function R(t).

Let $\phi_0(t)$ be the cumulative distribution function of the first passage time from initial state to a failed state, then

$$R^*(s) = \frac{[1-\phi_0^{**}(s)]}{s}$$

here $\phi_0^{**}(s)$ is the Laplace Stieltjes transform of the $\phi_0(t)$.
Thus, we have

$$MTSF = \lim_{s\to 0} \frac{\left[1-\phi_0^{**}(s)\right]}{s}.$$

AVAILABILITY

The availability A(t) is the probability that the system is operating satisfactorily at time t. When a system is often unavailable due to break downs, the concerning department becomes interested to put it back into operation after each break down with proper repairs. In fact, it is concerned with availability equally as it does with reliability because of additional costs and inconvenience incurred when the system is not available. The differences between the measures of reliability and availability are as follows:

(a) The reliability is an interval function while the availability is a point function describing the behaviour of the system at a specified epoch.

(b) The reliability function precludes the failure of the system during the interval under consideration, while availability function does not impose any such restrictions on the behaviour of the system.

Availability may be categorized as:

(i) Instantaneous (Point-wise) Availability

(ii) Average (Interval) Availability

(iii) Steady-State (Limiting Interval) Availability

MAINTAINABILITY

Maintainability is an indices associated with a system under repair. It is the probability that the system will be restored to operational effectiveness within a specified time when the maintenance is taken in accordance with prescribed conditions. Maintenance is one of the effective ways of increasing the reliability of a system. It is considered to be beneficial when the repair cost in terms of time and money spent is considerably low as compared to the cost of the equipment. A low repair time will minimize the ill-effects of the failure. Maintenance of a system is of two types:

(i) Preventive Maintenance (P. M.), and

(ii) Corrective Maintenance (C. M.)

(i) Preventive Maintenance

In this category equipment may be maintained by replacing or repairing components prior to the possible occurrence of a failure. It includes actions such as lubrications, replacement of a nut or a screw or some part of the system, refueling, cleaning, etc.

(ii) Corrective Maintenance

When a system fails to work, repair and adjustment are started immediately to put it in operable conditions. When such a failure will occur we cannot for see exactly and so this is called un-scheduled maintenance. It involves minor repairs that may crop up between inspections.

On failure of a unit, it is sent to a repair facility if available, otherwise it queues up for repair. There may be two types of repair policies as follows:

(a) Repeat Repair Policy

(b) Resume Repair Policy

REPAIRABLE SYSTEMS

If on failure, a unit is replaced by a new one, then the reliability of the system increases. In a good number of cases this will turn out to be expensive and it will be necessary to repair the

failed units. Thus on the failure of a unit, it is sent to a repair facility. If no repair facility is free, then the failed units queue up for repair and the repairs are normally undertaken in First In First Out (FIFO) order. We assume that the life time of an on-line unit, standby and the repair time of a failed unit are all independent random variables and that the distribution functions of these random variables are known and that they admit the probability density functions.

BUSY PERIOD OF REPAIRMAN

Let B(t) be the probability that a repairman is busy with the system in the interval (0,t] Then in the long run total fraction of time for which a repairman is busy is given by

$$B = \lim_{t\to\infty} \frac{B(t)}{t}$$

EXPECTED NUMBER OF REPAIRS

Let $R_r(t)$ be a random variable representing the number of times the units are repaired in (0, t], then the expected number of repairs in (0, t] is E[R(t)] and in the long run the expected number of repairs per unit time is given by

$$R_r = \lim_{t\to\infty} \frac{E[R_r(t)]}{t}$$

EXPECTED NUMBER OF REPLACEMENTS

Let $R_P(t)$ be a random variable representing the number of times the units are replaced in (0, t], then the expected number of replacements in (0, t] is E $[R_P(t)]$ and in the long run the expected number of replacements per unit time is given by

$$R_P = \lim_{t\to\infty} \frac{E[R_p(t)]}{t}$$

EXPECTED NUMBER OF PREVENTIVE/CORRECTIVE MAINTENANCES

Let $E_i(t)$ be a random variable representing the number of times the units are maintained in (0, t], then the expected number of maintenances in (0, t] is $E[E_i(t)]$ and in the long run the expected number of maintenances per unit time is given by

$$E_i = \lim_{t\to\infty} \frac{E[E_i(t)]}{t}$$

PROFIT ANALYSIS

Any manufacturing industry is basically a profit making organisation and no organisation can serve for long without minimum financial returns for its investment. There must be an optimal balance between the reliability aspect of a product and its cost.

Availability of the system leads to the revenue whereas the busy period of the repairman, expected number of replacements, etc. leads to the cost of maintenance and spares. Therefore, profit analysis is an important aspect in the field of reliability and depends upon production cost of maintenance and spares, failure rates, repairman employed, accidents, cost of calling repairman, etc.

The revenue and cost function leads to the profit function of a firm, as the profit is excess of revenue over the cost of production. The profit function takes the form

P(t) = Expected revenue in (0, t] – Expected total cost in (0, t].

Let us consider a system which involves only the following costs:

C_0 = revenue per unit up time of the system

C_1 = revenue per unit down time of the system

C_2 = cost per unit time of inspection

C_3 = cost per unit time of repair

C_4 = cost per unit time of replacement

C = other fixed costs

Here other fixed costs (C) includes cost of installation of the system, wages of the repairman/operator etc. Then the expected profit incurred of the system is

$P = C_0UT_0 - C_1DT_0 - C_2BI_0 - C_3BR_0 - C_4BRP_0 – C$

REFERENCES

[1] Aggarwal, K.K. and Gupta, J.S. (1975): On minimizing the cost of reliable systems, IEEE Transactions on Reliability, Vol.R-24, No.3, pp.205,205.

[2] Arora, J.R. (1976): Reliability of a 2-unit standby redundant system with constrained repair time, IEEE Transactions on Reliability, Vol.R-25, No.3, pp.203-205.

[3] Arora, J.R. (1977): Reliability of several standby priority redundant systems, IEEE Transactions on Reliability, Vol.R-26, No.4, pp.290-293.

[4] Balagurusamy, E. (1984): Reliability Engineering, Tata McGraw Hill Publishing Co. Ltd., New Delhi.

[5] Bai, L. (2005): Circular sequential k-out-of-n congestion system, IEEE Transactions on Reliability, Vol.54, No.3, pp.412-420.

[6] Barlow, R.E. and Hunter, L.C. (1960): Optimum preventive maintenance policies, Operations Research, Vol.8, No.1, pp.90-100.

[7] Barlow, R.E. and Proschan, F. (1965): Mathematical Theory of Reliability, John Wiley, New York.

[8] Bothwell, R., Donthamsetty, R., Kania, Z. and Wesoloski, R. (1996): Reliability evaluation: a field experience from motorola's cellular base transceiver systems, Proceedings of Reliability and Maintainability Symposium, International Symposium on Product Quality and Integrity, pp.348-359.

[9] Chevance, R.J. (2005): Hardware and software solutions for high availability, Server Architectures, pp.609-652.

[10] Dhillon, B.S. (1999): Design Reliability, Fundamentals and Applications, CRC Press, New-York.

[11] Epstein, B. (1958): The exponential distribution and its role in life testing, Industrial Quality Control, Vol.15, pp.2-7.

[12] Epstein, B. and Sobel, M. (1953): Life testing, J. Amer. Statist. Assoc., Vol.48, No.263, pp.486-502.

[13] Feller, J.W. (1968): An Introduction to Probability Theory and its Applications, Vol.1, 3rd edition, Wiley.

[14] Friedman, M.A. and Tran, P. (1992): Reliability techniques for combined hardware-software systems, Proceedings of Annual Reliability and Maintainability Symposium, pp.290-293.

[15] Garg, R.C. and Kumar, A. (1977): A complex system with two types of repair of failure and repair, IEEE Transactions on Reliability, Vol.26, No.4, pp.299-300.

[16] Gaver, D.P. (1963): Time to failure and availability of paralleled systems with repair, IEEE Transactions on Reliability, Vol.R-12, No.2, pp.30-38.

TEST AUTOMATION

Rajiv Sharma

*Computer Science Engineering Department, Shri Baba Mastnath Engineering College, Asthal Bohar, Rohtak, Haryana, India rajivsharma.075@ymail.com

In software testing, test automation is the use of special software (separate from the software being tested) to control the execution of tests and the comparison of actual outcomes with predicted outcomes.[1] Test automation can automate some repetitive but necessary tasks in a formalized testing process already in place, or add additional testing that would be difficult to perform manually.

Some software testing tasks, such as extensive low-level interface regression testing, can be laborious and time consuming to do manually. In addition, a manual approach might not always be effective in finding certain classes of defects. Test automation offers a possibility to perform these types of testing effectively. Once automated tests have been developed, they can be run quickly and repeatedly. Many times, this can be a cost-effective method for regression testing of software products that have a long maintenance life. Even minor patches over the lifetime of the application can cause existing features to break which were working at an earlier point in time.

There are many approaches to test automation, however below are the general approaches used widely:

- Code-driven testing. The public (usually) interfaces to classes, modules or libraries are tested with a variety of input arguments to validate that the results that are returned are correct.
- Graphical user interface testing. A testing framework generates user interface events such as keystrokes and mouse clicks, and observes the changes that result in the user interface, to validate that the observable behavior of the program is correct.
- API driven testing. A testing framework that uses a programming interface to the application to validate the behaviour under test. Typically API driven testing bypasses application user interface altogether.

Test automation tools can be expensive, and are usually employed in combination with manual testing. Test automation can be made cost-effective in the long term, especially when used repeatedly in regression testing.

In automated testing the Test Engineer or Software quality assurance person must have software coding ability, since the test cases are written in the form of source code which, when run, produce output according to the assertions that are a part of it.

One way to generate test cases automatically is model-based testing through use of a model of the system for test case generation, but research continues into a variety of alternative methodologies for doing so. In some cases, the model-based approach enables non-technical users to create automated business test cases in plain English so that no programming of any kind is needed in order to configure them for multiple operating systems, browsers, and smart devices. What to automate, when to automate, or even whether one really needs automation are crucial decisions which the testing (or development) team must make. Selecting the correct features of the product for automation largely determines the success of the automation. Automating unstable features or features that are undergoing changes should be avoided.

CODE DRIVEN TESTING

A growing trend in software development is the use of testing frameworks such as the xUnit frameworks (for example, JUnit and NUnit) that allow the execution of unit tests to determine whether various sections of the code are acting as expected under various circumstances. Test cases describe tests that need to be run on the program to verify that the program runs as expected. Code driven test automation is a key feature of agile software development, where it is known as test-driven development (TDD). Unit tests are written to define the functionality*before* the code is written. However, these unit tests evolve and are extended as coding progresses, issues are discovered and the code is subjected to refactoring .[4] Only when all the tests for all the demanded features pass is the code considered complete. Proponents argue that it produces software that is both more reliable and less costly than code that is tested by manual exploration.It is considered more reliable because the code coverage is better, and because it is run constantly during development rather than once at the end of a waterfall development cycle. The developer discovers defects immediately upon making a change, when it is least expensive to fix. Finally, code refactoring is safer; transforming the code into a simpler form with less code duplication, but equivalent behavior, is much less likely to introduce new defects.

GRAPHICAL USER INTERFACE (GUI) TESTING

Many test automation tools provide record and playback features that allow users to interactively record user actions and replay them back any number of times, comparing actual results to those expected. The advantage of this approach is that it requires little or no software development. This approach can be applied to any application that has agraphical user interface. However, reliance on these features poses major reliability and maintainability problems. Relabelling a button or moving it to another part of the window may require the test to be re-recorded. Record and playback also often adds irrelevant activities or incorrectly records some activities.[

A variation on this type of tool is for testing of web sites. Here, the "interface" is the web page. This type of tool also requires little or no software development. However, such a framework utilizes entirely different techniques because it is reading HTML instead of observing window events

Another variation is scriptless test automation that does not use record and playback, but instead builds a model of the Application Under Test (AUT) and then enables the tester to create test cases by simply editing in test parameters and conditions. This requires no scripting skills, but has all the power and flexibility of a scripted approach. Test-case maintenance seems to be easy, as there is no code to maintain and as the AUT changes the software objects can simply be re-learned or added. It can be applied to any GUI-based software application. The problem is the model of the AUT is actually implemented using test scripts, which have to be constantly maintained whenever there's change to the AUT.

FRAMEWORK APPROACH IN AUTOMATION

A test automation framework is an integrated system that sets the rules of automation of a specific product. This system integrates the function libraries, test data sources, object details and various reusable modules. These components act as small building blocks which need to be assembled to represent a business process. The framework provides the basis of test automation and simplifies the automation effort.

The main advantage of a framework of assumptions, concepts and tools that provide support for automated software testing is the low cost for maintenance. If there is change to

any test case then only the test case file needs to be updated and the driver Script and startup script will remain the same. Ideally, there is no need to update the scripts in case of changes to the application.

Choosing the right framework/scripting technique helps in maintaining lower costs. The costs associated with test scripting are due to development and maintenance efforts. The approach of scripting used during test automation has effect on costs.

Various framework/scripting techniques are generally used:

1. Linear (procedural code, possibly generated by tools like those that use record and playback)
2. Structured (uses control structures - typically 'if-else', 'switch', 'for', 'while' conditions/ statements)
3. Data-driven (data is persisted outside of tests in a database, spreadsheet, or other mechanism)
4. Keyword-driven
5. Hybrid (two or more of the patterns above are used)
6. Agile automation framework

The testing framework is responsible for :

1. Defining the format in which to express expectations
2. Creating a mechanism to hook into or drive the application under test
3. Executing the tests
4. Reporting results

HADOOP TECHNOLOGIES

Kavita

Department of computer science, S.B.M.N Engg. College, Asthal Bohar,Rohtak-124021(Haryana) INDIA
kavita.kathuria1990@gmail.com

Nowadays, Companies need to process Multi Petabyte Datasets efficiently. The Data may not have strict schema for the large system. It has become Expensive to build reliability in each Application for processing petabytes of datasets. If there is a problem of Nodes fail every day, some of the causes of failure may be. Failure is expected, rather than exceptional. The number of nodes in a cluster is not constant. So there is a Need for common infrastructure to have Efficient, reliable, Open Source Apache License.

The Hadoop platform was designed to solve problems where you have a lot of data perhaps a mixture of complex and structured data and it doesn't fit nicely into tables. It's for situations where you want to run analytics that are deep and computationally extensive, like clustering and targeting. That's exactly what Google was doing when it was indexing the web and examining user behavior to improve performance algorithms. This article has made an attempt to study its need, uses and application, thereby brought to the notice of the readers.

Hadoop is an open-source software framework written in Java for distributed storage and distributed processing of very large data sets on computer clusters built from commodity hardware. All the modules in Hadoop are designed with a fundamental assumption that hardware failures (of individual machines, or racks of machines) are commonplace and thus should be automatically handled in software by the framework.

Fundamental concept: Rather than banging away at one, huge block of data with a single machine, Hadoop breaks up Big Data into multiple parts so each part can be processed and analyzed at the same time. Why Hadoop used for searching, log processing, recommendation systems, analytics, video and image analysis, data retention? It is used by the top level apache foundation project, large active user base, mailing lists, users groups, very active development, and strong development teams.

HADOOP'S COMPONENTS

Hadoop Distributed File System: HDFS, the storage layer of Hadoop, is a distributed, scalable, Java-based file system adept at storing large volumes of unstructured data. MapReduce is a software framework that serves as the compute layer of Hadoop. MapReduce jobs are divided into two (obviously named) parts. The "Map" function divides a query into multiple parts and processes data at the node level. The "Reduce" function aggregates the results of the "Map" function to determine the "answer" to the query. Hive is a Hadoop-based data warehousing-like framework originally developed by Facebook. It allows users to write queries in a SQL-like language called HiveQL, which are then converted to Map Reduce. This allows SQL programmers with no Map Reduce experience to use the warehouse and makes it easier to

integrate with business intelligence and visualization tools such as Microstrategy, Tableau, Revolutions Analytics, etc. Pig Latin is a Hadoop-based language developed by Yahoo. It is relatively easy to learn and is adept at very deep, very long data pipelines (a limitation of SQL.)

Working process of Hadoop Architecture

Hadoop is designed to run on a large number of machines that don't share any memory or disks. That means you can buy a whole bunch of commodity servers, slap them in a rack, and run the Hadoop software on each one. When you want to load all of your organization's data into Hadoop, what the software does is bust that data into pieces that it then spreads across your different servers. There's no one place where you go to talk to all of your data; Hadoop keeps track of where the data resides. And because there are multiple copy stores, data stored on a server that goes offline or dies can be automatically replicated from a known good copy.

In a centralized database system, you've got one big disk connected to four or eight or 16 big processors. But that is as much horsepower as you can bring to bear. In a Hadoop cluster, every one of those servers has two or four or eight CPUs. You can run your indexing job by sending your code to each of the dozens of servers in your cluster, and each server operates on its own little piece of the data. Results are then delivered back to you in a unified whole. That's MapReduce you map the operation out to all of those servers and then you reduce the results back into a single result set.

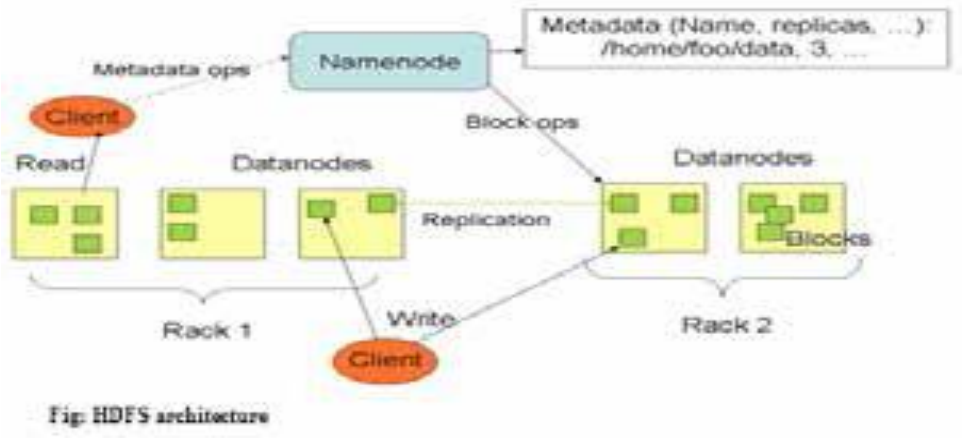


Figure-1

Architecturally, the reason you're able to deal with lots of data is because Hadoop spreads it out. And the reason you're able to ask complicated computational questions is because you've got all of these processors, working in parallel, harnessed together. Hadoop implements a computational paradigm named Map/Reduce, where the application is divided into many small fragments of work, each of which may be executed or re-executed on any node in the cluster.

In addition, it provides a distributed file system (HDFS) that stores data on the compute nodes, providing very high aggregate bandwidth across the cluster. Both Map/Reduce and the distributed file system are designed so that node failures are automatically handled by the framework. Hadoop Common is a set of utilities that support the other Hadoop subprojects. Hadoop Common includes File System, RPC, and serialization libraries.

CONCLUSION

The Hadoop Distributed File System (HDFS) is a distributed file system designed to run on commodity hardware. Hadoop is designed to run on cheap commodity hardware, It automatically handles data replication and node failure, It does the hard work – you can focus on processing data, Cost Saving and efficient and reliable data processing. It has many similarities with existing distributed file systems.

However, the differences from other distributed file systems are significant. HDFS is highly fault-tolerant and is designed to be deployed on low-cost hardware. HDFS provides high throughput access to application data and is suitable for applications that have large data sets. HDFS relaxes a few POSIX requirements to enable streaming access to file system data. HDFS was originally built as infrastructure for the Apache Nutch web search engine project. HDFS is part of the Apache Hadoop Core project.

EMOTIONAL INTELLIGENCE AND ITS IMPACT ON ORGNISATIONS

RITIKA KHURANA

Assistant Professor S.B.M.N Engg. College, Rohtak

INTRODUCTION:

Now a days organizations are facing cut throat competition in, Changing working environment and Culture, Encouraging Cosmopolitism (employing different regions people), Upgrading new technology, Increasing work pressure, Performance based pay, 360 performance, etc.. In these conditions to navigate our lives, it is our fears and envies, our rages and depressions, our worries and anxieties that steer us day to day more stressful declining our performance levels reducing productivity of firms.

HR managers are having a tough time in selecting a right person at right place. As we all know that whole recruitment process is an expensive process in any organization. In spite of India called as "Young Country" with skilled man power and large pool of educated unemployed students. Organizations are facing crunch of apt student who suits their requirement with adequate skill and expertise.

"Howard Gardner" a psychologist and author of "Frames Of Mind" said, to work in an organization. The pencil and paper work alone will not help. There are many tasks to perform like…..

- PARTICIPATIVE MANAGEMENT
- DECISION MAKING/PROBLEM SOLVING
- BENDING & MENDING RELATIONSHIPS -
- WORKING IN GROUP/TEAMS EFFECTIVELY
- PUTTING PEOPLE AT EASE
- LEADERSHIP SKILLS (INTIATIVE/RESPONSIBLE)

But for decades a lot of emphasis has been put on certain aspects of intelligence such as Logical reasoning, Mathematical skills, Quantitative aptitude skills, Understanding analogies, Verbal skills etc…in recruiting employees and admission tests in to premiere educational institutes.

Even the most academically brilliant among us are vulnerable to undone by unruly emotions. The price we pay for emotional literacy in professional life is Conflict with peers, Poor decision making, Less efficient in team etc.

And in personal life failed marriages, frequent clashes with parents & siblings deteriorating health and mental peace.

In US Emotional Intelligence is a concept evolved from psychology early 1950-1969.In 1970-1989, where psychologist focused on how emotions and thoughts influenced one another. Mayer and Salovey were the pioneers in it, In 1995 Daniel Goleman wrote a book on EI, it was hugely sold, many business leaders preferred to implement in their organizations.

According to Goleman, Emotional Intelligence encompasses the following five characteristics and abilities.

- **SELF AWARENESS**: knowing your emotions, recognizing feelings as they occur discriminating between them. The key to success is knowing oneself i.e. self awareness is knowing one's internal states, preferences, resources, solutions etc..It indicates the ability

to recognize, understood & accept one's own moods, emotions, drives, strengths and short comings as well as to see now these affect other people.

- **MOOD MANAGEMENT**: Handling feelings so they are relevant to the current situation and you react appropriately. It refers to managing & handling impulses, distressing feelings & upsets rather than denying or representing these feelings. Self control helps in staying compose, focused, calm and helps think clearly even under pressure.
- **SELF MOTIVATION**: "Gathering Up" your feelings, directing yourself towards a goal, despite self-doubt inertia and impulsiveness. It refers to striving to impose to meet a standard of excellence. They are result oriented. To take calculated risks & readily face any type of challenges.
- **EMPATHY:** Recognizing feelings in others & tuning in to their verbal & non-verbal cues. Sensing other feelings & perspective ness, and taking an active interest in their concerns, it is the ability to put oneself in to another's shoes and look at things or think from his point of view.
- **MANAGING RELATIONSHIPS**: Handling inter personal interaction, conflict resolution negotiation. Adeptness at inducing desirable responses in others. Developing others, sensing others development needs and bolstering their activities. They recognize other's strengths and accomplishments and help them in developing their personality. They provide useful feed back give timely coaching, offer challenging assignments.

Millennium manager's roles are diversified from regular managerial functions to others things like.

- The ability to accurately perceive emotions,
- To access and generate emotions so as to assist thought process,
- To understand the information of those emotions and manage them
- And solve their problems on the basis of them.

Emotions' and 'Emotional Needs' are given main importance for better relationships of internal and external customers.

Relationship in Business: Emotions like sad, negative, irritation, frustration, aggressiveness, short temper, ego, envy etc…, play a negative role in work place.

Emotions are contagious; a single person can influence the emotions of others in a team. Some how through facial expressions such as happiness, sadness, anger and fear were universally recognizable in human beings. The mental ability of EI can be evaluated by two tests.

1. Specific ability test.
2. General integrative test.

Most of the big organizations like XEROX, IBM, GE, TCS have their own measure of evaluating EQ (Emotional Quotient), they relate it that higher levels of EI are associated with better performance in the following areas.

1. Participative Management.
2. Pulling people at Ease.
3. Balance between personal life and work.
4. Straight Forwardness & Composure.
5. Decisiveness
6. Doing whatever it takes
7. Adaptability.
8. Confronting Problem Employees

Organizations believe that EI be useful as a mental ability test, which means it is part of a broader class of mental capacities that also include creativity, verbal fluency, possibility thinking, mental absorption etc…

EI has proven better predictor of success. The people who manage their own feelings well can deal effectively with others who are more likely to live content lives.

EI influence organizational effectiveness in:

- Employee Recruitment & Retention,
- Development of Talent,
- Team Work,
- Innovation,
- Customer Loyalty,
- Productivity,
- Efficiency,
- Sales / Revenues,
- Quality of Service.

MSCEIT (Mayor Salovey Caruso EI Test), used to performance measure of EI which helps in:

- Accurately perceiving a persons emotions facilitates the prediction, understanding of that subsequent actions and

- Understanding the emotional states of others makes to know their attention, sincerity, seriousness, decision making and behavioral responses.

CONCLUSION:

90% of top performers are high in EQ, EQ alone explains 58% of a leaders job performance.

Now-a-days, a person is exposed to many cultures and influenced by many things. Organizations earlier used to give tangible benefits to the internal / external customers, these days customers are looking for fulfillment of their emotional needs. Organizations in along run to maintain customer loyalty they need to take care of employees emotional needs and behave empathetically.

In real life, the companies who had experienced the importance of EI are

US Air Force reduced recruiter turnover from 35%annually to 5% annually by selecting candidates high in EI.

Outcomes of improving emotional intelligence.

- Have better social support,
- More successful and avoid inter arguments & flights,
- Lower level of aggression and less conflicts,
- Involved in more social networks,
- Less likely to addict drugs & alcohol,
- To motivate oneself,
- To control impulses,
- To persist in the face of frustrations,
- To regulate moods.

References:

Daniel Goleman.
"emotional intelligence"
www.google/hr/ei.com

STRESS-STRAIN DIAGRAM

Sandeep Kumar Jangra

*Mechanical Engineering Department, Shri Baba Mastnath Engineering College, Asthal Bohar, Rohtak, Haryana, India sandeepjangra095@ymail.com

STRESS-STRAIN BEHAVIOUR OF MATERIAL

All engineering materials do not show same sort of behaviour when subjected to tension as well ascompression. There exist some materials like metals, alloys etc., which are more or less equally strong in both tension and compression. And these materials are generally tested in tension again concrete, stones, bricks etc., are such type of materials which are weaker in tension and stronger in compression. Hence, these materials are tested in compression. Now the stress-strain characteristics of mild steel are of specific importance to the community dealing with basic engineering science.

STRESS-STRAIN CHARACTERISTICS OF MILD STEEL (M.S)

In order to obtain stress-strain behaviour of M.S, a specimen of uniform circular cross-section is prepared following the specification laid in IS 1608:2005 identical to ISO 6892:1998. A specific length of maximum 4 inch or 100mm is generally selected in the well-middle part of the specimen and this length is designated as gauge length, over which the amount of elongation is studied.

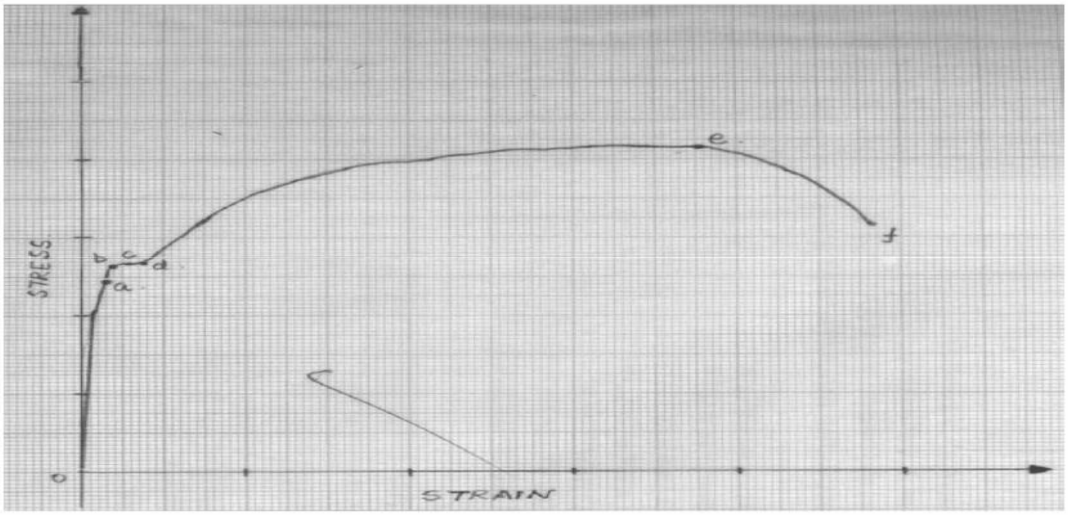


Fig. 1

Now the specimen, suitably fitted in extensometer, mounted on the machine where loading is started gradually from zero till failure. Following is a stress-strain curve of M.S specimen having gauge length 100mm, tested in Amsler Universal Testing Machine of capacity 20T . Various points on stress-strain curve are marked in Figure.

PROPORTIONAL LIMIT

It is the point on the stress-strain curve, up to which the plot is a straight line and stress is proportional to strain. Up to proportional limit, the material remains elastic and strictly follows Hooke's Law.

ELASTIC LIMIT

In the stress-strain curve, it is the point just beyond proportional limit. From proportional limit to elastic limit, the material remains elastic but does not follow Hooke's Law and so, stress and strain are not proportional.

YIELD POINT

When the specimen is loaded beyond elastic limit, it enters into elasto-plastic zone. In this region, elongation of specimen occurs by considerable amount without any perceivable amount of increase in load. Sometimes this yielding is accompanied by an abrupt reduction of load and thereby stress. In this case the upper and lower limits of stress are called upper yield point or stress and lower yield point or stress, respectively. Lower yield stress is normally considered as yield stress σy of material, because upper yield stress is affected by speed of testing, form of specimen and shape of cross-section.

PROOF STRESS

Some materials like High Strength Deformed (HSD) steel, brass, duralumin etc., do not show any well defined yield point. For these materials, proof stress serves as analogous to yield stress.

ULTIMATE STRESS

Yield point serves as the gateway to plastic zone. Beyond yield point, due to sudden decrease in load, material begins to strain-harden and recover some of the elastic property. And by virtue of that, gradual uprise of stress-strain curve occurs and terminates at a point, called ultimate stress. This is the maximum stress, the specimen can withstand, without any appreciable damage or permanent deformation.

BREAKING STRESS

While ultimate stress is the maximum stress with standing capacity prior to failure, further increase of ultimate stress leads to failure of the specimen and this occurs at breaking stress. Here the value of breaking stress lower than ultimate stress, as appearing in the stress-strain diagram obtained during experiment, of ductile material, is somehow misleading. What happens in reality is that, beyond ultimate stress, there occurs a reduction in area of cross-section near at the middle of gauge length. This phenomenon is called formation of neck or formation of waist. As the grips of extensometer are attached at the end of gauge length, the effect of neck formation thereby the reduction in diameter of the specimen cannot be taken into account. By reason of which breaking stress exhibits value lower than ultimate stress. And this breaking stress is called Nominal Breaking Stress. When the reduced cross-sectional area at neck is considered to compute actual stress, it is found that breaking stress is pretty higher than ultimate stress. And this is called True Breaking Stress. In case of brittle material, ultimate stress is same as breaking stress.

WORKING STRESS AND FACTOR OF SAFETY

In practical design of structures, some uncertainties may be associated in terms of loading, material properties etc. Not only that, in some materials, like concrete, non-ferrous alloys etc., Hooke's Law does not hold good. To encompass all these aspects, it is essential to limit actual stress generated to a value comparatively lower than yield stress of the material. And this stress is considered as a safe one. This safe stress is designated at Working Stress (σw). A pure number, higher than 1 (whole or fraction) that divides the yield stress to obtain working stress is called Factor of Safety.

SWITCHING

Seema

Department of computer science, S.B.M.N Engg. College, Asthal Bohar,Rohtak-124021(Haryana) INDIA
E-mail-seemakoshik16@yahoo.co.in

A switch is a device used on a computer network to physically connect devices together. Multiple cables can be connected to a switch to enable networked devices to communicate with each other. Switches manage the flow of data across a network by only transmitting a received message to the device for which the message was intended. Each networked device connected to a switch can be identified using a MAC address, allowing the switch to regulate the flow of traffic.

Because of these features, a switch is often considered more "intelligent" than a network hub. Hubs neither provide security, or identification of connected devices. This means that messages have to be transmitted out of every port of the hub, greatly degrading the efficiency of the network.

SWITCHING

A switched network consists of a series of Interlinked node is called switches. Switches are devices capable of creating temporary connections between two or more devices linked to the switch. In a switched network, some of these
Nodes are connected to the end systems. Others are used only for routing.

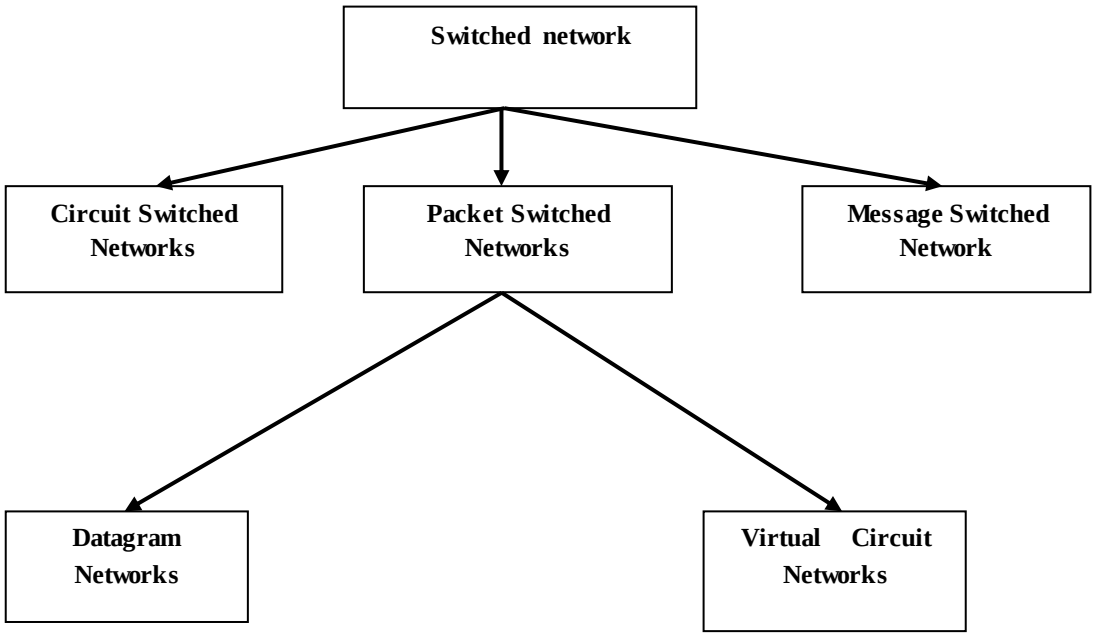


Figure 1

TYPE OF SWITCHING

There are three types.

- Circuit Switched Networks
- Packet Switched Networks
- Message Switched Networks

Circuit Switching

When two nodes communicate with each other over a dedicated communication path, it is called circuit switching. There is a need of pre-specified route from which data will travels and no other data is permitted. In circuit switching, to transfer the data, circuit must be established so that the data transfer can take place.

Circuits can be permanent or temporary. Applications which use circuit switching may have to go through three phases:

- Establish a circuit
- Transfer the data
- Disconnect the circuit

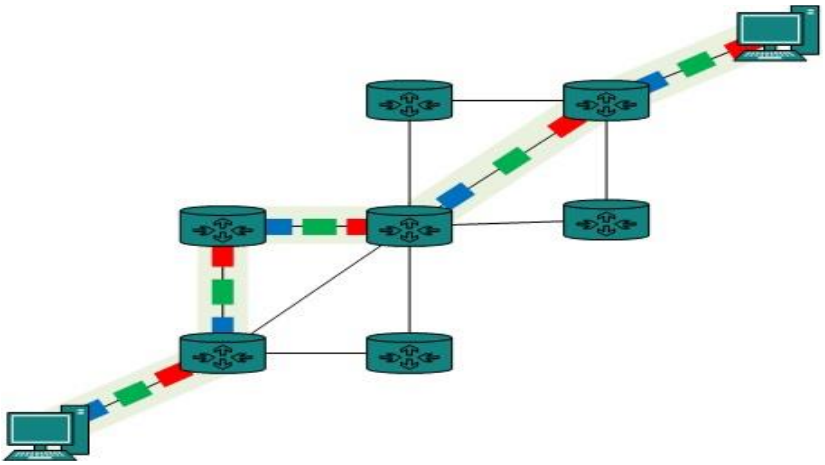


Figure 2

Circuit switching was designed for voice applications. Telephone is the best suitable example of circuit switching. Before a user can make a call, a virtual path between caller and caller is established over the network.

Message Switching

This technique was somewhere in middle of circuit switching and packet switching. In message switching, the whole message is treated as a data unit and is switching / transferred in its entirety. A switch working on message switching, first receives the whole message and buffers it until there are resources available to transfer it to the next hop. If the next hop is not having enough resource to accommodate large size message, the message is stored and switch waits.

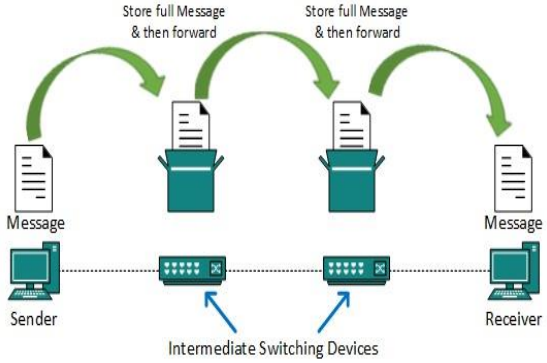


Figure 3

This technique was considered substitute to circuit switching. As in circuit switching the whole path is blocked for two entities only.

Packet Switching

Shortcomings of message switching gave birth to an idea of packet switching. The entire message is broken down into smaller chunks called packets. The switching information is added in the header of each packet and transmitted independently. It is easier for intermediate networking devices to store small size packets and they do not take much resources either on carrier path or in the internal memory of switches.

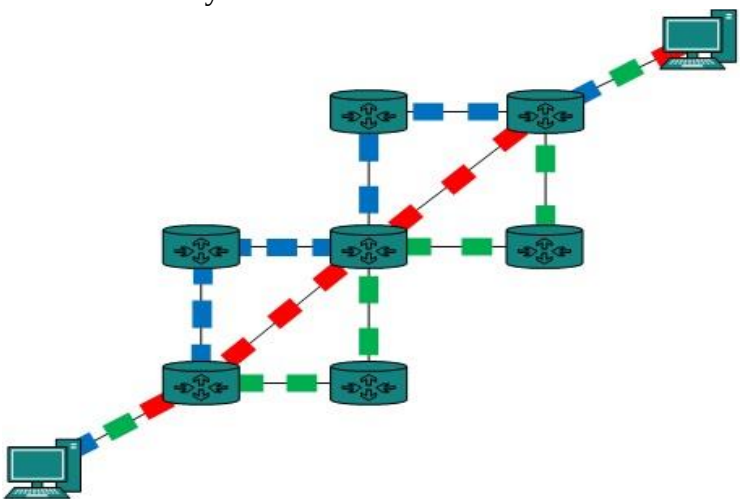


Figure 4

Packet switching enhances line efficiency as packets from multiple applications can be multiplexed over the carrier. The internet uses packet switching technique. Packet switching enables the user to differentiate data streams based on priorities. Packets are stored and forwarded according to their priority to provide quality of service.

REFERENCES

[1] http://en.wikipedia.org/wiki/Network_switch

[2] http://en.wikipedia.org/wiki/Packet_switching

[3] Data communication & networking B.Forouzan fourth edition

REVIEW ON SIGHT DISTANCE CONSIDERATIONS

Sonu Panchal[1],Vikram[2]

Student, Department of Civil Engineering, SBMNEC, Asthal Bohar, Rohtak, Haryana, India;Email: sonu.007panchal@yahoo.com; Email: vikrampanu25@gmail.com

1. ABSTRACT

The safe and efficient operation of vehicles on the road depends very much on the visibility of the road ahead of the driver. Thus the geometric design of the road should be done such that any obstruction on the road length could be visible to the driver from some distance ahead . This distance is said to be the sight distance. In other words, sight distance is the length of the length of the road visible ahead to the driver at any instance.

1.1 TYPES OF SIGHT DISTANCE

Sight distance available from a point is the actual distance along the road surface, over which a driver from a specified height above the carriage way has visibility of stationary or moving objects. Three sight distance situations are considered for design:

- Stopping sight distance (SSD) or the absolute minimum sight distance
- Overtaking sight distance (OSD) for safe overtaking operation
- Safe sight distance to enter into an intersection

Apart from three situations, the following are considered by IRC:-

- Intermediate sight distance (ISD) is the defined as twice SSD
- Head light sight distance is the distance visible to a driver during night driving under the illumination of head light.

The standards for sight distance should satisfy the following conditions:-

- Driver travelling at the design speed of the highway must have sufficient carriageway distance within his line of vision to allow him to stop his vehicle before colliding with a slowly moving or stationary object appearing suddenly in his own traffic lane.
- Driver travelling at the design speed should be able to safely overtake slow moving vehicle at suitable intervals.
- Driver entering an uncontrolled intersection has sufficient visibility to enable hom to take control of his vehicle and to avoid collision.

THE COMPUTATION OF SIGHT DISTANCE DEPENDS ON:

- **REACTION TIME OF THE DRIVER**

Reaction time of a driver is the time taken from the instant the object is visible to the driver to the instant when the brakes are applied. The total reaction time may be split up into four components based on PIEV theory. In practice, all these times are usually combined into a total reaction time suitable for design purposes as well as for easy measurement. Many of

the studies shows that drivers require about 1.5 to 2secs under normal conditions. However taking into consideration the variability of driver characteristics, a higher value is normally used in design. For example, IRC suggests a reaction time of 2.5secs.

- **SPEED OF THE VEHICLE**

The speed of the vehicle very much affects the sight distance. Higher the speed, more time will be required to stop the vehicle. Hence it is evident that, as the speed increases, sight distance also increases.

- **EFFICIENCY OF BRAKES**

The efficiency of the brakes depends upon the age of the vehicle, vehicle characteristics etc. If the brake efficiency is 100%, the vehicle will stop the moment the brakes are applied. But practically, it is not possible to achieve 100% brake efficiency. Therefore it could be understood that sight distance required will be more when the efficiency of brakes are less.

- **FRICTIONAL RESISTANCE**

The frictional resistance between the tyre and road plays an important role to bring the vehicle to stop. When the frictional resistance is more, the vehicles stop immediately. Thus sight required will be less. No separate provision for brake efficiency is provided while computing the sight distance. This is taken into account along with the factor of longitudinal friction. IRC has specified the value of longitudinal friction in between 0.35 to 0.4.

- **GRADIENT OF ROAD ,IF ANY**

Gradient of the road also affects the sight distance. While climbing up a gradient, the vehicle can stop immediately. Therefore sight distance required is less. While descending a gradient, gravity also comes into action and more time will be required to stop the vehicle. Sight distance required will be more in that case.

1.2 STOPPING SIGHT DISTANCE

SSD is the minimum sight distance available on a highway at any spot having sufficient length to enable the driver to stop a vehicle travelling at design speed, safely without collision with any other obstruction.

There is a term called safe stopping distance and is one of the important measures in traffic engineering. It is the distance a vehicle travels from the point at which a situation is first perceived to the time the deceleration is complete. Drivers must have adequate time if they are to suddenly respond to a situation. Thus in a highway design, a sight distance at least equal to the safe stopping distance should be provided.

ANALYSIS OF SSD:-

The stopping sight distance is the sum of lag distance and the braking distance.

LAG DISTANCE:-

Lag distance is the distance the vehicle travelled during the reaction time t and is given by **vt**, where v is the velocity in $m/sec2$.

BRAKING DISTANCE:-

Braking distance is the distance travelled by the vehicle during braking operation. For a level road this is obtained by equating the work done in stopping the vehicle and the kinetic energy of the vehicle. If F is the maximum frictional force developed and the braking distance is l, then work done against friction in stopping the vehicle is Fl = fWl where W is the total weight of the vehicle. The kinetic energy at the design speed is:-

$\frac{1}{2}mv^2 = \frac{1}{2}\frac{wv^2}{g}$ $\qquad$ $fWl = \frac{wv^2}{2g}$

$l = \frac{v^2}{2gf}$

Therefore, the SSD = lag distance + braking distance and given by:

$$SSD = vt + \frac{v^2}{2gf}$$

Where, l is braking distance in m
v is the design speed in m/sec,
t is the reaction time in sec,
g is the acceleration due to gravity and
f is the coefficient of friction.

The coefficient of friction f is given below for various design speed.

Speed, kmph	<30	40	50	60	>80
F	0.40	0.38	0.37	0.36	0.35

Stopping distance at slopes:-

The general equation is given by Equation

$$SSD = vt + \frac{v^2}{2g(f \pm 0.01n)}$$

Where n is gradient + in ascending case and – in descending case

1.3 OVERTAKING SIGHT DISTANCE

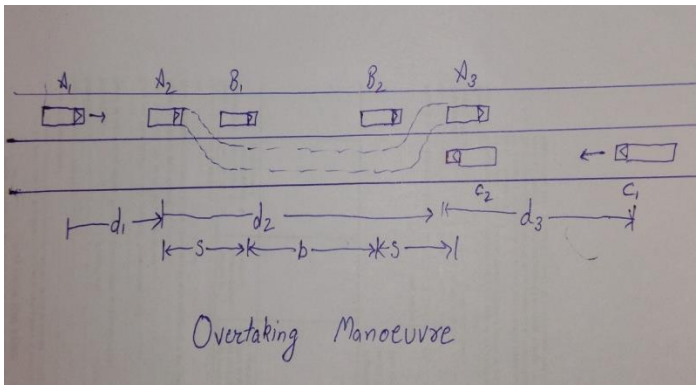
The overtaking sight distance is the minimum distance open to the vision of the driver of a vehicle intending to overtake the slow vehicle ahead with safety against the traffic in the opposite direction .

The overtaking sight distance or passing sight distance is measured along the center line of the road over which a driver with his eye level 1.2 m above the road surface can see the top of an object 1.2 m above the road surface.

The factors that affect the OSD are:

- Speed of the overtaking vehicle, overtaken vehicle and of the vehicle coming in the opposite direction.
- Distance between the overtaking and overtaken vehicles ,spacing
- Skill and reaction time of the driver
- Rate of acceleration of overtaking vehicle
- Gradient of the road

Analysis of OSD:-



The overtaking sight distance consists of three parts.

- d_1 the distance travelled by overtaking vehicle A during the reaction time
- d_2 the distance travelled by the vehicle during the actual overtaking operation T
- d_3 is the distance travelled by on-coming vehicle C during the overtaking operation (T).

Therefore:

$$OSD = d_1 + d_2 + d_3$$

It is assumed that the vehicle A is forced to reduce its speed to vb, the speed of the slow moving vehicle B and travels behind it during the reaction time t of the driver. So d1 is given by:

$$d_1 = v_b \text{ t}$$

Then the vehicle A starts to accelerate, shifts the lane, overtake and shift back to the original lane. The vehicle A maintains the spacing s before and after overtaking. The spacing s in m is given by:

$$\text{s} = 0.7v_b + 6$$

Let T be the duration of actual overtaking. The distance travelled by B during the overtaking operation is 2s + v_bT. Also, during this time, vehicle A accelerated from initial velocity vb and overtaking is completed while reaching nal velocity v. Hence the distance travelled is given by:

$$\text{d}_2 = (\text{b} + 2\text{s}) = (v_b + \frac{\text{aT}^2}{2})$$

therefore

$$2s = \frac{\text{aT}^2}{2}$$

$$T = \sqrt{\frac{4s}{a}}$$

$$\text{d}_2 = v_b + 2\text{s}$$

The distance travelled by the vehicle C moving at design speed v m=sec during overtaking operation is given by:

$$d_3 = \text{vT}$$

Now,

$$\text{OSD} = d_1 + d_2 + d_3$$

$$\textbf{OSD} = \boldsymbol{v_b}\ \textbf{t} + \boldsymbol{v_b}\ \textbf{T+2s + vT}$$

where

$\boldsymbol{v_b}$ is the velocity of the slow moving vehicle in m/sec,

t the reaction time of the driver in sec,

s is the spacing between the two vehicle in m

a is the overtaking vehicles acceleration in m/sec^2

The acceleration values of the fast vehicle depends on its speed and given in Table below

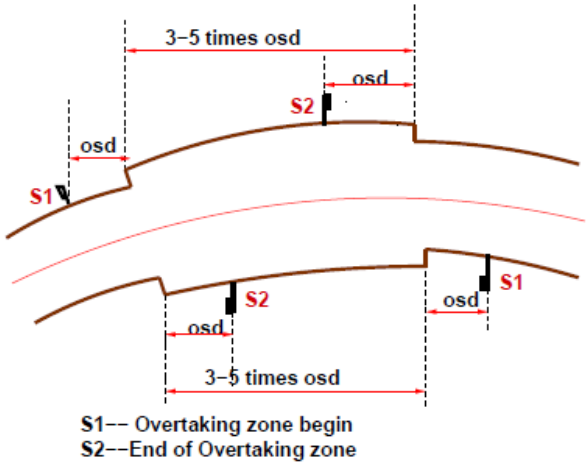
Speed (kmph)	Maximum overtaking acceleration (m/sec^2)
25	1.41
30	1.30
40	1.24
50	1.11
65	0.92

80	0.72
100	0.53

Note that: On divided highways, d3 need not be considered
On divided highways with four or more lanes, IRC suggests that it is not necessary to provide the OSD, but only SSD is sufficient.

Overtaking Zones

Overtaking zones are provided when OSD cannot be provided throughout the length of the highway. These are zones dedicated for overtaking operation, marked with wide roads. The desirable length of overtaking zones is 5 times OSD and the minimum is 3 times OSD.



1.4 SIGHT DISTANCE AT INTERSECTIONS

At intersections where two or more roads meet, visibility should be provided for the drivers approaching the intersection from either sides. They should be able to perceive a hazard and stop the vehicle if required. Stopping sight distance for each road can be computed from the design speed. The sight distance should be provided such that the drivers on either side should be able to see each other. This is illustrated in the figure. Design of sight distance at intersections may be used on three possible conditions:

- Enabling approaching vehicle to change the speed
- Enabling approaching vehicle to stop
- Enabling stopped vehicle to cross a main road

1.5 SUMMARY

One of the key factors for the safe and efficient operation of vehicles on the road is sight distance. Sight distances ensure overtaking and stopping operations at the right time. Different types of sight distances and the equations to find each of these had been discussed here.

RELAY EDITION TECHNOLOGY IN TERMS OF POSITION FOR NEXT GENERATION SYSTEM

Niranjan Yadav[1], Dr. Subham Gandhi[2], Dr. V R Sing[3]
Assistant Professor, PDMCTM Bahadurgarh & PhD Scholar Baba Masthnath University Rohtak[1] Associate Professor, ECE department, BMU Rohtak [2]
Director, PDMCE Bahadurgarh[3]
niranjanyadav97@gmail.com[1], subhamsavailable@gmail.com[2], vrsigh@ieee.org[3]

ABSTRACT

we study the impacts of the deployment of a new type of relay node (RN), i.e., moving RN (MRN) on current cellular systems. For efficient heterogeneous network planning, 3GPP LTE-Advanced has introduced concept of Relay Nodes (RNs). The Relay Nodes are low power eNodeBs that provide enhanced coverage and capacity at cell edges. One of the main benefits of relaying is to provide extended LTE coverage in targeted areas at low cost. There are several benefits as well as challenges of using MRNs to serve users inside public transportation vehicles. In a noise limited single cell system, the deployment of MRN can significantly lower the end-to-end outage probability (OP) at the user equipment (UE) on board compared to serving the UE by direct or fixed RN (FRN) assisted transmission. We focus on the application of MRNs in different scenarios. In order to understand whether there is a potential benefit to add a new type of node to the existing system, transmission assisted by FRN is employed as a reference scheme and conventional BS to UE direct transmission is used as the base line. In all the studies, in order achieve a fair comparison, we employed different ways, i.e., by analytical analysis, numerical methods or by simulation, to optimize the FRN position which minimize the overall OP at the UE on average. Furthermore, MRNs can also help to lower the UE transmit power, which may potentially save the energy used to operate the communication network.

The purpose of this work is to show initial understandings of deploying a new type of node, i.e., the MRN, to the network. The analysis begins with simple setups and both FRN assisted transmission and direct transmission are used as references

Index Terms—LTE-Advanced, RN, MRN, FRN,

1.INTRODUCTION

The Third Generation Partnership Project's Long Term Evolution-Advanced is considering the use of relaying for cost-effective throughput enhancement and coverage extension. While analog repeaters have been used to enhance coverage in commercial cellular networks, the use of more sophisticated fixed relays is relatively new. The main challenge faced by relay deployments in cellular systems is overcoming the extra interference added by the presence of relays. Because most prior work on relaying does not consider interference, good relay strategies for cellular networks are not widely known. LTE offers a very high peak data rate in ideal conditions. However, the capacities of the LTE network is not evenly distributed, i.e., the cell edge users have much worse throughput than cell center users. The successor of LTE, the LTE-Advanced, aims at both further improve the system capacity and improve the cell edge user experiences. In order to extend the coverage for heavily shadowed or remote areas, and guarantee good user experiences at certain capacity demanding hotspot areas, a heterogeneous and small cell networks (HetSNets) design paradigm has been introduced in LTEAdvanced systems.

This paper analyzes the performance of several emerging half-duplex relay strategies in the context of interference- limited cellular systems: one-way relays, two-way relays, and shared relays. The performance benefits of each relay strategy as a function of location, sectoring, and

frequency reuse is compared with local base station coordination. One-way relaying is shown to provide modest gains over single-hop cellular networks in some regimes. Shared relaying is shown to approach many of the gains of local base station coordination at reduced complexity, while two-way relaying further reduces complexity but only works well when the relay is very close to the mobile device. Frequency reuse of one, where each sector and cell reuses the same spectrum, is shown to have the highest network throughput. Simulations with realistic channel models provide performance comparisons that reveal the importance of interference mitigation in multihop cellular networks. The studies have been extended to more practical setups when considering the impact of co-channel interference. In such scenarios, MRN assisted transmission still greatly outperforms direct transmission and FRN assisted transmission in terms of end-to-end OP, when the vehicular penetration loss (VPL) is moderate to high. Moreover, due to the low transmit power nature of an MRN, it generates much less interference to the UEs outside the vehicle, which is very appreciated in a densely deployed urban scenario, since link availabilities are usually dependent on interference rather than on coverage. Hence using MRNs seems very promising for improving the quality-of-service for vehicular user's in future mobile communication systems.

GPP Release 11 is working on enhancing the relay functionality with moving relays. An example use case would be trains. High speed trains are currently deployed worldwide and providing the required services to the massive number of UEs that are moving with the train can be challenging. The carriages are well shielded with coated windows and therefore introduce a rather high penetration loss in the range of 20–30 dB. The high speed generates a Doppler frequency shift that causes frequency and channel estimation errors. Furthermore, the handover success rate of the connected UEs is lower compared to UEs moving with lower speed because of inaccurate or late neighbor cell measurements and excessive signaling; furthermore, frequent cell reselections drain UE batteries.

Relay nodes are targeted to improve user data rates for cell edge users and other users with poor radio coverage, including indoor users. The relay node supports full eNodeB functionality including encoding, decoding and packet scheduling. This chapter presents the general relay overview, relay physical layer, relay architecture and protocols and radio resource management. It discusses the coverage and capacity gains in simulations and future relay enhancements. The simulations show that relay nodes bring a substantial improvement in the cell edge user data rates. The benefit is highest when the relay node can be placed close to the user thus providing good signal-to-noise ratio. The relay node brings the largest benefit for the operator if the backhaul cost of the traditional eNodeB is high, site cost is low and the network is coverage limited.

My research in one year has been focusing on the application of moving relay nodes (MRNs), which in our point of view may bring significant benefits to the existing wireless systems and next generation wireless system

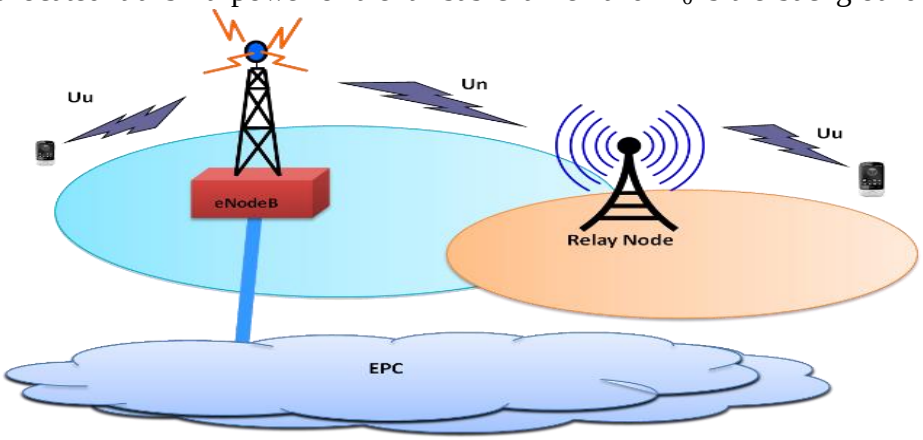
Heterogeneous and Small Cell Networks

The throughput demands and evenly distributes the capacity of a wireless network, a new design paradigm, i.e., the Heterogeneous and Small Cell Networks (HetSNets), was introduced in LTE release 10. The idea of HetSNets is to deploy several low power nodes within the coverage of macro

BSs to either extend the coverage or boost the local capacity in certain hotspot areas. The HetSNets has many advantages. Let us have a look at a simple example as follows. We use the Shannon capacity to approximate the capacity as

$$C = \sum_{i=1}^{k} B_i \log_2 \left(1 + \frac{P_i}{N_0}\right),$$

Where i is the number of sub channels, B_i is the bandwidth of the ith sub channel, P_i is the allocated transmit power of the ith sub channel and N_0 is the background noise power.



We would like to increase the capacity of a network; we can either increase the system bandwidth or increase the received signal-to-noise ratio (SNR). Bandwidth is a scarce resource and the available bandwidth of a mobile communication system is always limited. Hence, increasing the received SNR is one of the most practical ways to increase the network capacity. There have been various methods applied in wireless systems to increase the received SNR, ranging from interference mitigation, interference alignment, to sophisticated multi-point coordination schemes. path loss, the basic cause of the power loss during transmission, cannot be easily alleviated. The fundamental idea of HetSNets is to increase the network capacity by using node densifications, which is an effective way to combat the power loss caused by the path loss. In a HetSNets scenario, as the cell radius becomes smaller, BSs are much closer to the users and the transmit power of BSs can be significantly lowered to achieve the required coverage. RNs are playing indispensable roles in modern wireless communication systems, ranging from sophisticated space exploration to daily mobile communication systems

Amplify and forward RN

This type of RN simply amplifies the signal, and then forwards it to the destination. This type of RN is easy to implement and has been widely used in the current wireless systems, e.g., repeaters in the GSM network. The signals received at the RN and the destination from the source can be expressed as

$$y_{\mathrm{R}}^{(\mathrm{S})} = h_{\mathrm{R}}^{(\mathrm{S})} x + n_{\mathrm{S}},$$
$$y_{\mathrm{D}}^{(\mathrm{S})} = h_{\mathrm{D}}^{(\mathrm{S})} x + n_{\mathrm{D}},$$

Where x represents the signal sent from the source; $h^{(S)}R$ and $h^{(S)}D$ are channel gain between the source and the RN, the source and the destination, respectively; n_S and n_D represent the noise affecting the communication. After receiving the signal from the source, the RN amplifies the received signal and forwards it to the destination. If we denote the gain of the RN as G, and a channel gain of $h^{(R)}$ D is assumed between the RN and the destination, then the signal received at the destination from the RN is given as

$$y_{\mathrm{D}}^{(\mathrm{R})} = G\, h_{\mathrm{D}}^{(\mathrm{R})}\, y_{\mathrm{R}}^{(\mathrm{S})} + n_{\mathrm{D}}.$$

If there is no direct link between source and destination we can see that a drawback of amplify and forward (AF) RN is that it both amplifies the signal and the noise. In the presence of interference, the AF RN also amplifies the interference. Thus, the deployment of AF RN must be at a place with SNR or SINR advantages; otherwise little gains can be expected from such RNs.

Decode and forward RN

In a decode and forward (DF) RN assisted transmission, in the first step the RN decodes part or the entire received signal. In the second step, the RN re-encodes the decoded message and forwards it to the destination. DF RNs do not suffer from the problem of noise or interference amplification as AF RNs, but it may have the risk of error propagation if the RN fails to decode the message from the source correctly. Thus, in practical systems, e.g., the fixed RN (FRN) in the LTE-Advanced system, a cyclic redundancy check (CRC) may be used to spot the wrongly decoded message, and if an error is detected, the RN stops to forward the message and requests a re-transmit of the message from the source.

The end-to-end SNR or SINR expression of the DF RN is generally unknown. Instead, the outage probability (OP) is used as a measure of its performance. Since the DF RN has to decode the message from the source first and then
forwards the message to the destination, as long as one of the hops is in outage then an outage can be declared for the end-to-end communication. the QoS requirements can be mapped to a minimum SNR or SINR requirement. Thus, if either of the two links falls below a certain SNR or SINR threshold, an outage occurs. Defining the minimum required SNR or SINR threshold as γth_R, the end-to-end OP for DF RN can be expressed as

$$P_{\mathrm{out_R}}\left(\gamma_{\mathrm{th_R}}\right) = \Pr\left(\min\left(\gamma_{\mathrm{R}}, \gamma_{\mathrm{D}}\right) < \gamma_{\mathrm{th_R}}\right),$$

Where γ_R and γ_D are the SNR or SINR at the RN and the destination respectively.

Other Relay Nodes

There are lots of other types of RNs being widely studied both academically and during industrial standardization processes, and they all show their advantages in certain scenarios. Those types of RNs include but not limited to compress and- forward (CF), estimate-and-forward (EF) and hybrid AF/DF RNs.

II.RN techniques in 3GPP LTE systems

The application of RNs in practical cellular systems has a long history. Repeaters have been used in the age of the GSM system and more advanced FRNs are standardized in LTE release 10. During the initial study of 3GPP LTE systems, two types of RNs have been defined, type-1 RN and type-2 RN. Type-1 RNs are non-transparent, both to the BS and the UE. From the UE viewpoint, type-1 RN appears as a regular BS, and terminates all the layer-2 and layer-3 communication protocols. It has its own cell ID, control channel, reference and Synchronization signals as well as support re-transmission processes. Thus, type-1 RNs can be used for both capacity boosting and coverage extension. The FRN in LTE re-use most of the radio interface standardized for the UEs for the initial attachment to the serving BS (also known as Donor eNB or DeNB in LTE3), and then identifies itself as an RN. Three classes of type-1 RNs were discussed during the studies of 3GPP, i.e., in-band half duplex, out band full duplex and in-band

full duplex RNs but only the in-band half duplex type-1 RNs are standardized in 3GPP LTE release 10. In contrast, type-2 RNs have not been standardized yet and only received a functional description from 3GPP. Type-2 RNs do not have their own cell ID, and thereby are transparent to UEs. Rather to be used for coverage extension, this type of RNs is designed to cooperate with BSs to improve the capacity for certain hotspot areas. Type-2 RNs are not standardized yet but it opens the window for technologies such as CF, EF and hybrid AF/DF RNs in future release of the LTE standards. As type 2 RNs require cooperation with BSs, new air interfaces need to be defined between BSs and RNs. Hence, type-2 RNs may bring much bigger impacts to the existing system than type-1 RNs.

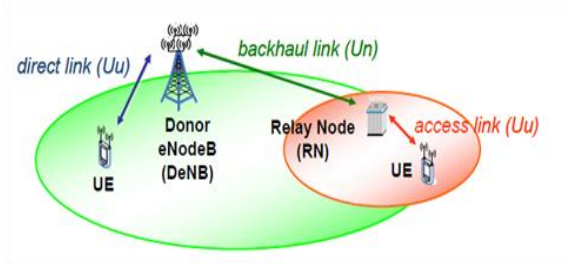


Figure 11: Rel-10 LTE Relaying Architecture

A specific vocabulary has been associated to this LTE feature:

- Main BS is the Donor eNodeB (DeNB)
- Relay Station is the Relay Node (RN)
- Mobile station is the User Equipment (UE)
- BS-RN link is the Backhaul Link
- RN-MS link is the Access Link
- BS-MS link is the Direct Link

During the standardization of relaying node at 3GPP level, only Layer 3 RN have been proposed and two types of them have been clearly defined: type-1 RN and type-2 RN. One major reason on this specific focus was that the relay should be transparent to the UE. Being not aware of the presence of the relay, existing R8/R9 UE can be served by relays defined in R10.

Type-1 RNs are appearing as a regular BS to connected UE and terminates all the layer-2 and layer-3 communication protocols. It has its own cell ID, control channel, reference and synchronization signals as well as support re-transmission processes. Thus, type-1 RNs can be used for both capacity boosting and coverage extension. The relay node acts first as a classical UE for the Donor eNodeB on which an initial attachment is performed. During this attachment phase, the UE indicates explicitly to the networks its RN status and deals with the BS the radio resources to open the backhaul link. Three classes of type-1 RNs were discussed during the studies of 3GPP: inband half duplex (type-1), out band full duplex (type-1a) and inband full duplex (type-1a) RNs. Only the in-band half duplex type-1 RNs are standardized in 3GPP LTE release 10.

The opposite, type-2 RNs have not been standardized in R10 and only received a functional description from 3GPP. Type-2 RNs do not have their own cell ID and look just like the main cell. Any UE is not able to distinguish a relay from the main eNB within the cell. Control information can be transmitted from the eNB and user data from the LTE relay. Instead of using it for coverage extension, this type of RNs is suitable for a global cooperation scheme manage in the cell to improve the capacity for hotspot areas. Under the COMP (Coordination

Multipoint) 3GPP framework some type-2-RN schemes have been investigated and proposed.

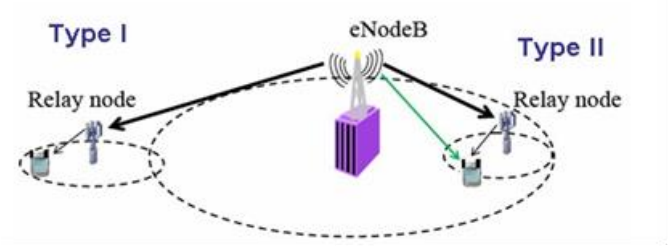
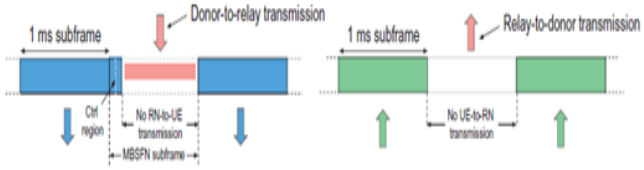


Figure 12: Type I & Type II LTE Relay Nodes

When in band relaying mode is selected, the backhaul and access links operate in the frequency. When it is not possible to isolate both links through HW-based isolation mechanisms (using specific antenna configurations for example) time domain mechanisms must be proposed to eliminate the avoid loop-back interference effect.



Such mechanism must guarantee that the relay is not transmitting on the access link at the same time as it is receiving on the backhaul link (and vice versa). The usage of certain LTE sub frames has been revisited to match this basic behavior.

III.Implementation and scenario

Increasing capacity and reducing capital and operational costs are major goals of every mobile operator throughout the world. With the introduction of Long Term Evolution – Advanced, a 3GPP standard, a number of techniques to improve capacity and coverage while reducing the cost to the operator have been introduced. Amongst these techniques is the introduction of advanced relays.

A relay is similar to a repeater which is currently in widespread use for 2G/3G technologies. Relay nodes are basically low power base stations which enhance coverage in areas where coverage is poor or to provide coverage in rural areas without the need for a wired backhaul connection. The cost to purchase and implement a relay station is significantly less than that of an eNodeB. The backhaul between the relay and the eNodeB is wireless which reduces costs in the implementation of infrastructure. A repeater (also known as a Layer 1 relay) extends the coverage of an existing base station to areas where the base station cannot reach or there is a high SINR such as at the edge of the cell, areas where shadowing occurs or to provide coverage indoors. A Layer 1 relay amplifies and forwards downlink and uplink signals between the UE and the base station. A major disadvantage of a Layer 1 relay is that not only is the desired signal amplified and forwarded but also unwanted interference such as inter-cell interference and noise. A Layer 2 Relay acts by demodulating and decoding the incoming signal and re-modulating and re-encoding the signal before the amplified version is transmitted. This process overcomes the problems associated with the Layer 1 relay as no inter-cell interference or noise is being amplified and re-transmitted. A Layer 3 relay operates in a similar manner to a Layer 2 relay with additions such as having a unique Physical Cell ID to ensure the UE knows it is connected to a relay node and treats the relay node as a base station. Layer 3 relays have been standardized by 3GPP for release with LTE Advanced (Rel. 10).

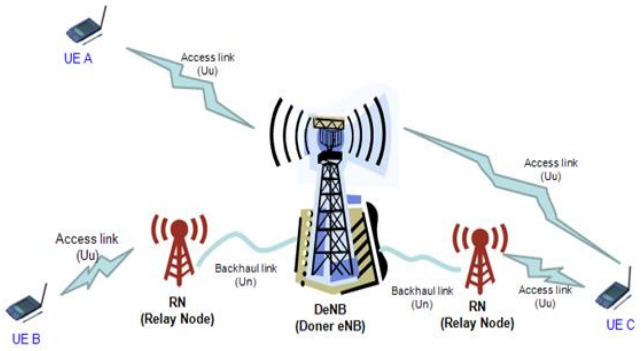


Figure 1: Relay Nodes

The figure above shows a scenario where relaying is utilized. In scenario (a) the UE communicates directly with the eNodeB. In scenario (b) a relay is utilized to extend the coverage area of the eNodeB and the UE communicates with the eNodeB via a relay node. In scenario (c) a relay node is utilized to overcome excessive shadowing, the signal from the eNodeB would not be able to reach the desired location behind tall buildings if it was not for the use of this relay node. In scenario (d) the area is considered to be a "hotspot" where there are a high number of users and the relay node is utilized to increase the available throughput to these users.

The installation of new eNodeBs in a network can be very expensive both in the cost of installing and the cost of operation. The installation of relays in a network is a cheaper alternative to improving network performance while reducing costs and energy consumption. Relays do not need to be as high as a standard eNodeB and they also connect back to the eNodeB wirelessly which reduces the need for a complicated wired interface. They are also not as large as eNodeBs and do not require air conditioned rooms or cooling units

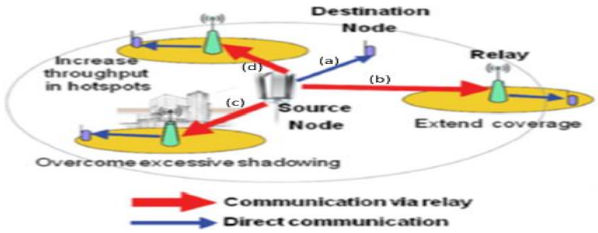
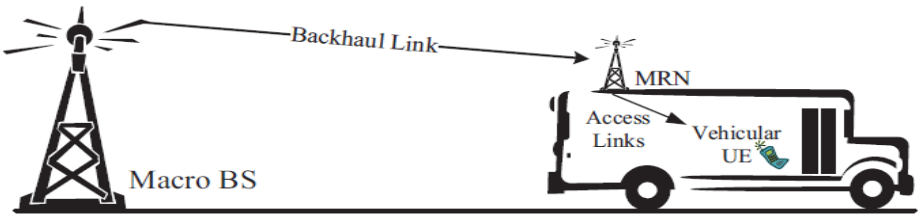


Figure 2: Implementation Scenarios

IV.Dedicated MRN deployment

A new solution that is under investigation is to deploy dedicated MRNs on top of public transportation vehicles to serve the UEs on board, as shown on Fig. 3.1. MRNs are low power BSs mounted on the vehicles where MRNs connect to macro BSs via radio interface. The advantages of deploying a dedicated MRN are not only due to its ability to eliminate the VPL. As MRNs are not limited by size and power as regular UEs, it can better exploit various smart antenna techniques as well as more advanced signal processing schemes to further boost the performance. For example, in a train, several backhaul antennas can be interconnected and form a cooperative and coordinated relay system (CCRS), which strengthens the backhaul link by using antenna selection techniques. In a bus or tram, prediction antennas can be exploited. Thus more reliable CSI can be obtained; hence, facilitates the channel dependent scheduling of the

backhaul link.



FRN physical layer interfaces standardized in 3GPP LTE release 10 can be reused by MRNs. An MRN can create its own cell and terminates all the layer-2 and layer-3 communication protocols, and thereby serves the UEs on board as a regular BS. MRN may also have the potential to support multi-RAT functionalists,

V.Simulation Results

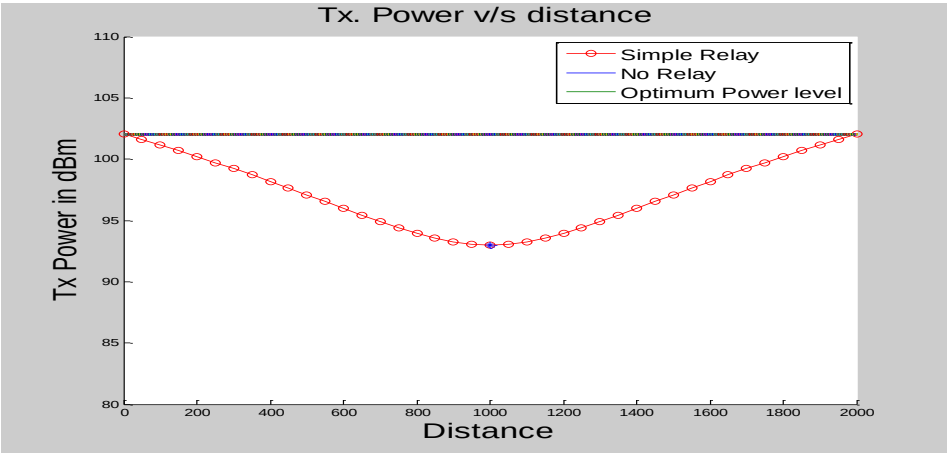
A matlab simulation was performed to demonstrate the reduction in transmission level in simple two-hop relay compared to conventional line of sight transmission.

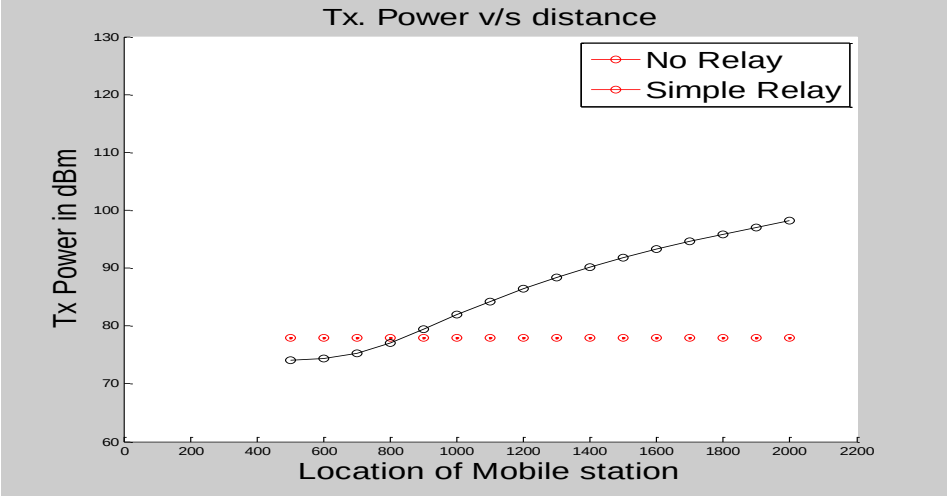
Case1: calculation for transmission power level over a distance calculation of transmission power level over a fixed between BS and MS, while RS is placed at variable distance.

The calculations were performed for receiving power level of -30 dBm with path loss exponent4. The values were taken over a distance 0 to 5000 meters. The result in figure shows that using simple relay reduces the transmission power level between BS and MS. The minimum power level in this ideal case is obtained when the RS is placed at halfway distance between BS & MS

Case 2: demonstrate the reduction in transmission power level by using the simple relay compared to direct line of sight transmission, when the MS is in motion away from BS.

The calculation with the same parameter as case 1 to receive power level of -30 dBm as MS and the Rs was placed at 400 meter from BS. The location of MS is changed further away from the each calculation.





VI.Conclusion & future work

It is shown that relay technologies can effectively improve service coverage and system throughput, especially when multiple RSs are deployed. With additional complexity and processing delay, a selective DCF scheme can achieve better performance than AF and DMF relay schemes under different radio channel conditions. In order to serve as many UE units as possible in a realistic multiple-RS-multiple-UE scenario, we have proposed and evaluated both centralized and distributed pairing schemes, which can achieve maximal numbers of served UE units and much higher cell throughput performance than random and opportunistic pairing schemes.

The relay technology to be used in the network provides better performance compared to conventional method, in terms of increase in coverage capacity, achievable peak data rate and reduction in power level. Thus, the cost effective relay technology is better alternate for signal transmission to meet the requirement of high data rate and QOS in advanced mobile system. for future we would like to demonstrate the power consumption level in DCF and simulate effects of noise in current simulation result. The result is based on idea noise free environment.

Our studies were only limited to simple scenarios where analytical analysis can be performed. We gained quite a few understandings of the benefits of employing MRNs as well as the challenges. In order to gain more comprehensive understandings of how this type of new nodes will impact on the mobile networks, system level evaluations need to be conducted; however, this needs to be done by simulations. With the help of more advanced simulations, more practical scenarios, e.g., multiple vehicles moving through multiple cells, can be studied

ACKNOWLEDGMENT

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REFERENCES

[1] Niranjan Yadav, Dr. Subham Gandhi, Prof. (Dr.) V R Singh, **"**Relay Edition Technology for next Generation Cellular System" 2nd International Conference on Electrical, Electronics, Engineering Trends, Communication, Optimization and Sciences (E^3COS) , 28th - 30th March-2015, page 779-805

[2] L. M. Correia, D. Zeller, O. Blume, D. Ferlin *et al*., "Challenges and enabling technologies for energy aware mobile radio networks," *IEEE Commun*

[3] Yang Yang, Honglin Hu, Guoqiang Mao Relay Technologies for WiMAX and LTE-Advanced Mobile Systems

[4] S. Stavrou and S. Saunder, "Factors influencing outdoor to indoor radio wave propagation," in Proc. 12th ICAP, 2003.
[5] C.Mehlfhrer, M.Wrulich, J. C. Ikuno, D. Bosanska and M. Rupp, Simulating the Long Term Evolution Physical Layer; in Proc. of the 17th European Signal Processing
[6] Abdallah Bou Saleh, Simone Redana, Jyri Hamalainen, and Bernhard Raaf, On the Coverage Extension and Capacity Enhancement of Inband Relay Deployments in LTE-Advanced Networks, in Hindawi Publishing Corporation Journal of Electrical and Computer Engineering Volume 2010, Article ID 894846
[7] Tommaso Beniero, Simone Redana, Jyri Hämäläinen, Bernhard Raaf; Effect of Relaying on Coverage in 3GPP LTE-Advanced; IEEE 69th Vehicular Technology Conference, VTC Spring 2009, Barcelona.
[8] Cyril Leung, Raymond Kwan, Seppo Hamalainen, and WenboWang, LTE/LTE-Advanced Cellular Communication Networks, in Journal of Electrical and Computer Engineering, volume 2010.
[9] ITU-R Rep. M.2134, "Requirements Related to Technical Performance for IMT-Advanced Radio Interface(s)," 2008.
[10] A. F. Molisch, *Wireless Communications*, 2nd ed. West Sussex, UK: John Wiley & Sons Ltd., 2010
[11] A. Papadogiannis, M. Farber, A. Saadani, M. D. Nisar *et al*., "Advanced relaying concepts for future wireless networks," in *Proc. Future Network Mobile Summit*, 2012
[12] Y. Sui, A. Papadogiannis, and T. Svensson, "The potential of moving relays—a performance analysis," in *Proc. IEEE VTC 2012-Spring*.
[13] A. Papadogiannis, M. Farber, A. Saadani, M. D. Nisar et al., "Advanced relaying concepts for future wireless networks," in Proc. Future Network Mobile Summit, 2012

TO STUDY THE STOCK STRIP LAYOUT FOR BLANKING OPERATION IN SHEET METAL

Yati Parjesh[1] Sumit[2]
[1,2] Department of Mechanical Engineering, Shri Baba Mastnath Engineering College, Asthal Bohar, Rohtak, Maharshi Dayanand University, Rohtak, Haryana, India.

Abstract
Sheet Metal Operations is one of the basic mechanical operations of forming metal into thin, flat uniform size & required shape part. To name a few sheet metal operations—shearing, blanking, punching, bending, embossing, trimming etc are used to produce a variety of shapes. In this paper, designing of strip layout for blanking operation has been discussed about. A comparison of different layouts for single component has been made to show the relative yield percentage.
Keywords: - Sheet Metal, Strip Layout, Blanking, Scrap Web, Strip Width, Advance Distance, Yield

I. INTRODUCTION

A variety of parts are manufactured using sheet metal operations. The first basic operation after shearing of the metal sheet is blanking, wherein the component is blanked out of a stock strip of metal. So, it is very necessary to design the layout of parts for blanking in such a way that maximum possible stock utilization or yield is obtained. A proper nesting of the parts for blanking operation is a must to produce maximum yield. Even though maximum utilization of the stock is necessary, but the shape & size of the part is of the component is of prime importance. So, the strip layout must be designed in such a way that the part obtained is given priority to the yield percentage of the stock.

The various terminologies associated with stock & scrap strip layout as shown in figure 1 are as below—

1. Feed Direction: It is the direction in which stock strip is fed in to the die.
2. Lead End: It is the end of the stock strip heading towards the die.
3. Tail End: It is the end of the stock strip opposite to the lead end.
4. Advance: It is the distance moved by the stock strip in blanking operation between two consecutive strokes of the machine.
5. Scrap Bridge: Scrap Bridge is the distance between the peripheries of two consecutive blanks. It is very necessary as it joins & separates the boundaries of two consecutive parts.
6. Stock Width: It is the width of the strip in vertical direction from which blanks are blanked out.
7. Front & Back Scrap: It is the scrap width in front & back of the blank.

The scrap or stock web to be left on the scrap strip for effective blanking operation depends on the type of blanking, thickness of the sheet, width of the strip, dimension or size of the component & contour of the blanked shape. The table below shows the values of scrap web allowance for various sizes of the parallel edged component.

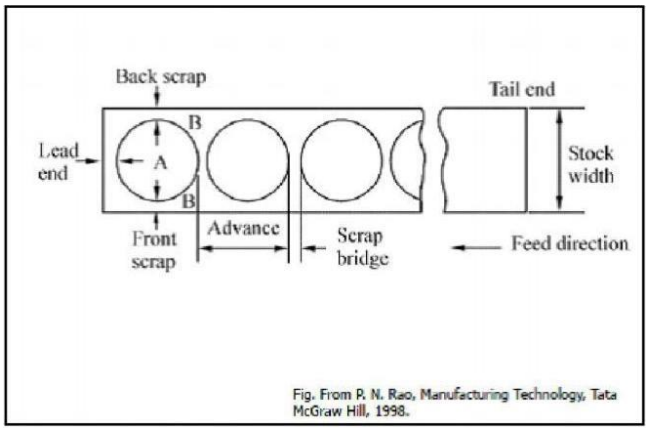


Fig.1: Scrap Strip Terminology

TABLE ISCRAP VALUES FOR VARIOUS MAXIMUM DIMENSIONS

	Dimension A		Dimension B	
Maximum Dimension (mm)	**General**	**Minimum (mm)**	**General**	**Minimum (mm)**
Up to 25 mm	1.25T	1.50	1.50T	1.50
26 to 75	1.25T	1.50	1.50T	1.50
76 to 150	1.50T	2.00	1.50T	2.40
151 to 250	1.75T	2.40	1.75T	3.00
251 to 400	2.00T	3.00	2.00T	4.75

Figure 2 shows the dimensions of the object as below-

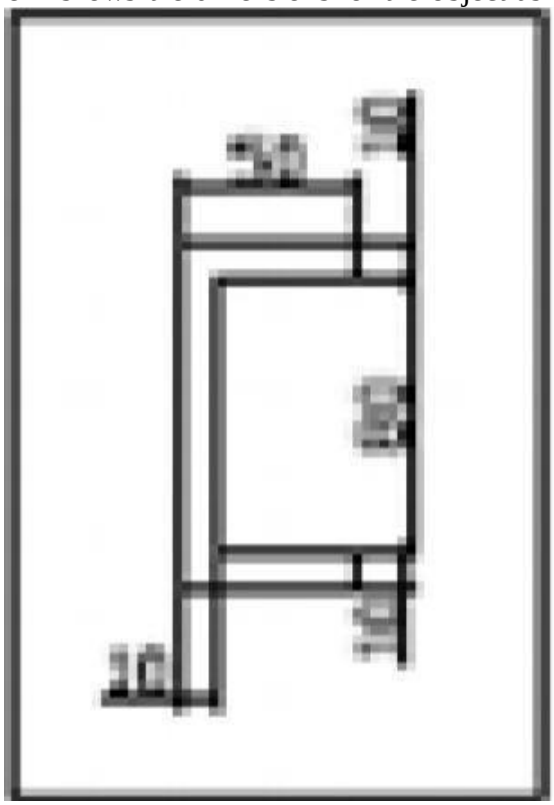
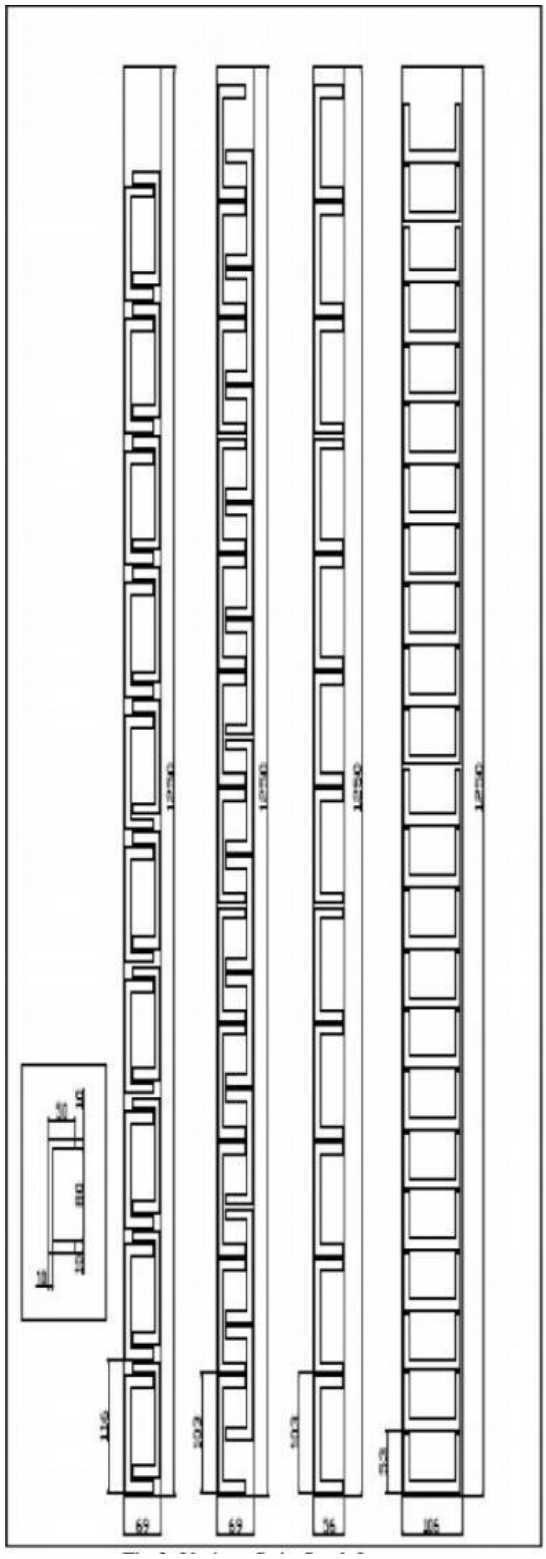


Fig.2: Component Dimension

Figure 3 shows four types of strip layout for blanking operation for the same component. Sheet Metal stock is obtained usually in the form of rolls but when in cut sheet, the usual size is

1250X2500sq.mm. So, we too have taken the cut sheet for our calculation. Before discussing on each of the options shown in figure 3 above, the common aspects for each option are— 1) Area of the Cross-section of 1 Component = (50*10) + (80*10) + (50*10) = 1800 sq. mm 2) Area of 1 sheet = 1250*2500 = 3125000 sq. mm 3) Thickness of Sheet = 2mm The various options are worked as below to find to find the yield percentage of each of the method—



A. Option 1

Maximum Dimension of the component = 100mm
From Table, for maximum dimension of 100mm,
Scrap Web = 1.5(T) = 1.5*2 = 3mm
Therefore, maximum dimension with 2 components
= 100+3+10 = 113mm
Advance Length = 113+3 = 116mm
No. of blanks per strip = 1250/116 = 10.77566 ~~ 10 blanks
But, 2 in number, therefore no. of blanks = 2*10 = 20
Strip Width = 50 + 1.5(T) + 10 + 1.5(T) + 1.5(T) = 69mm
No. of strips per sheet = (2500/ Strip Width) = 2500/69
= 36.23188 ~~ 36 strips
Thus, total number of blanks per sheet = 20*36 = 720
Area of total no. of blanks = 720 * 1800 = 1296000 sq. mm
Therefore, % utilization of stock or Yield %
= 1296000/ 3125000 = 41.47%

B. Option 2

Maximum Dimension of the component = 100mm
From Table, for maximum dimension of 100mm,
Scrap Web = 1.5(T) = 1.5*2 = 3mm
Advance Length = 100+3 = 103mm
No. of blanks per strip = (1250/103) + (1250/103) – 1
~~ 12+12-1 = 23 blanks
Strip Width = 50 + 1.5(T) + 10 + 1.5(T) + 1.5(T) = 69mm
No. of strips per sheet = (2500/ Strip Width) = 2500/69 = 36.23188 ~~ 36 strips
Thus, total number of blanks per sheet = 23*36 = 828
Area of total no. of blanks = 828 * 1800 = 1490400

sq. mm
Therefore, % utilization of stock or Yield %
= 1490400/ 3125000 = 47.69%

C. Option 3

Maximum Dimension of the component = 100mm
From Table, for maximum dimension of 100mm, Scrap Web = 1.5(T) = 1.5*2 = 3mm
Advance Length = 100+3 = 103mm
No. of blanks per strip = (1250/103) = 12.13592
~~ 12 blanks
Strip Width = 50 + 1.5(T) + 1.5(T) = 56mm
No. of strips per sheet = (2500/ Strip Width)
= 2500/56 = 44.64286 ~~ 44 strips
Thus, total number of blanks per sheet = 12*44
= 528
Area of total no. of blanks = 528 * 1800 = 950400 sq. mm
Therefore, % utilization of stock or Yield %
= 950400/ 3125000 = 30.41%

D. Option 4

Maximum Dimension of the component = 50mm
From Table, for maximum dimension of 100mm, Scrap Web = 1.25(T) = 1.25*2 = 2.5 ~~ 3mm
Advance Length = 50+3 = 53mm
No. of blanks per strip = (1250/53) = 23.58491
~~ 23 blanks
Strip Width = 100 + 1.25(T) + 1.25(T) = 106mm
No. of strips per sheet = (2500/ Strip Width)
= 2500/106 = 23.58491 ~~ 23 strips
Thus, total number of blanks per sheet = 23*23
= 529
Area of total no. of blanks = 529 * 1800 = 952200 sq. mm
Therefore, % utilization of stock or Yield %
= 952200/ 3125000 = 30.47% Thus, seeing all the 4 options, it can easily be said that Option 2 is the best as the maximum stock utilization of about 47.69% is obtained with it. Therefore, option 2 is the most economical method as far as maximum stock utilization is concerned.

II. CONCLUSIONS

Stock Strip Layout is a very important step in designing blanking process in sheet metal operations as the rest of the operations are performed after the part is blanked out of the stock strip. So, proper nesting of the strip must be done so as to obtain maximum yield & less scrap out of the sheet stock. As is seen here, option 2 was nested by making a third blank in between 2 blanks. So, maximum yield has been obtained in it. Here, neither the number of passes to cut the blanks nor the cost of the die has been discussed, which also are a significant factor in designing the stock strip layout.

ACKNOWLEDGMENT

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REFERENCES

[1] P N Rao, Manufacturing Technology: Foundry, Forming and Welding, 2010, vol. 1.
[2] D. Eugene Ostergaard, Basic Diemaking: Committee of the National Tooling & Machining Association, 1963.
[3] D. Eugene Ostergaard, Advanced Diemaking: Resources and Materials of the National Tooling & Machining Association, 1967.
[4] J.R. Paquin& R.E. Crowley, Die Design Fundamentals2nd ed., 1967.
[5] S. Kumar and R. Singh, --Automation of strip-layout design for sheet metal work on progressive die, Journal of Materials Processing Technology, 195:94-100, 2008.
[6] PRESS TOOLS- Striplayout –Oocities. Available:
http://www.oocities.org/toolanddiemaking/presstools/striplayout.html
[7] The Tooling University Website. [Online]. Available: http://www.toolingu.com/definition-400130-15721-blanking.html
[8] Sheet Metal- Wikipedia, The free encyclopedia. Available: http://en.wikipedia.org/wiki/Sheet_metal

GREEN CHEMISTRY INITIATIVES IN INDIA FOR PRESENT AND FUTURE

Sharwan K Dewan*,Vinay Batra[1]

* Formerly Prof.of Organic Chemistry , Deptt. Of Chemistry ,M.D.University,Rohtak.
1. Assistant Prof. S.B.M.N.Engg.College,Asthal Bohar,Rohtak. Email id: sharwankumardewan@yahoo.com; vb30393@gmail.com

ABSTRACT

Our environment, which is endowed by nature, needs to be protected from ever increasing chemical pollution. The challenge for the institution and industries is to come together and pursue development in the field of Green chemistry by reducing or eliminating the use and generation of hazardous substances. Delhi University has organized many national and international symposiums and workshops for promoting Green chemistry in India, which has provided the platform for interaction of concepts among the leading researchers. The main idea behind this to activate work toward Green chemistry for which involvement of academic, industrial, and governmental and non-governmental bodies is needed collectively, which will help in designing and development of environment-friendly chemistry practices in India.

INTRODUCTION

India, the second largest producer of pesticides and ranked 12th in pharmaceutical production, is fast emerging among the top 5 players in selected petrochemicals. These facts, in turn, have led to an increased stress on our delicate environmental balance, thus India needs to pursue Green chemistry along with progressive chemistry more exhaustively and extensively. Due to large-scale production of pesticides, pharmaceuticals, petrochemicals, and other consumer durables, there is a great potential for Green chemistry research in India to refine the existing technologies and also to find more environmentally benign alternatives. To increase the research in this field, we need to publicize the needs, effects, and practice of Green chemistry. Recently, Green chemistry research in India is confined mainly to areas of Greener synthetic strategies, catalyst development, usage of biocatalysis, usage of nonconventional technologies, and analytical techniques.

BASIC PRINCIPLES OF GREEN CHEMISTRY

Green chemistry is defined as environment benign chemical synthesis. Any synthesis, whether performed in teaching laboratories or industries should create none or minimum by-products which pollute the atmosphere. According to the research carried out by Paul T Anastas (father of Green chemistry), the following basic principles of Green chemistry have been formulated.

1 (Prevention of waste/ by-products) It is better to prevent waste than to clean up waste after it is formed.

2 (Maximum incorporation of the reactants, starting materials and reagents, into the final product) Synthetic method should be designed to maximize the incorporation of all materials used in the generation process into the final product.

3 (Prevention or minimization of hazardous products) Whenever practicable, synthetic methodologies should be designed to use and generate substances that possess little or no toxicity to human health and the environment.

4 (Designing of safer chemicals) Chemical products should be designed to preserve efficacy of function while reducing toxicity.

5 (Energy requirement for any synthesis should be minimum) Energy requirement should be recognized for their environmental and economic impacts, and should be minimized. Synthetic methods should be conducted at ambient temperature and pressure.

6 (Selecting the most appropriate solvents)) The use of solvents, separation agents, etc (auxiliary substances) should be made unnecessary whenever possible and innocuous when used.

7 (Selecting the most appropriate starting materials) A raw material of feedstock should be renewable rather than depleting wherever technically and economically practicable.

8 (Use of protecting group should be avoided whenever possible) Unnecessary preparation of derivatives, by blocking groups, temporary modification of physical/chemical processes etc, should be avoided wherever possible.

9 (Use of catalysts should be preferred wherever possible) Catalytic reagents, as selective as possible, should be considered superior to stoichiometric reagents.

10 (Products obtained should be biodegradable) Chemical products should be designed so that at the end of their function they do not persist in the environment and break down into innocuous degradation products.

11 (The manufacturing plants should be so designed as to eliminate the possibility of accidents during operations) Substances and form a substance used in chemical processes should be chosen so as to minimize the potential for chemical accidents, including releases, explosions and fire. 12(Strengthening of analytical techniques to control hazardous compounds) Analytical methodologies need to be developed to allow for real time, in-process monitoring and control prior to the formation to hazardous substances.

GREEN STRATEGIES

In developing Green synthetic strategies, Indian researchers are mainly concentrating on avoiding environmentally noncompatible reagents, solid-phase syntheses, modification of synthetic routes to decrease the number of steps and increase overall yield, usage of newer catalysts and simplification of classical procedures of reaction. However, what is required is a combined approach for a Greener synthesis. Catalyst and reagent chemistry is one of the most important aspects of eco-friendly chemistry. Reagent chemists in India are working toward development of more benign and selective reagents that require ambient conditions. The elimination of hazardous solvents is one of the prime concerns among them. Enzymes have emerged as biotechnological tools, which can offer solutions to the major problems of the chemical industry in India. Over the years, chemists in India are engaged in enhancement of an application base of enzymes to develop new alternative sweeteners such as high fructose corn syrup (HFCS), synthetic honey, and other food products such as polysaccharide gums, thickeners, and flavor enhancers. There is a great need to develop newer enzymes that can work at ambient conditions and to determine their optimum activity by in-depth study. An interdisciplinary approach and healthy partnership between research institutions and industry can very effectively evolve solutions to problems faced like the increase in the cost of chemical fertilizers and consequent risk of degradation of soil fertility by excessive use of chemical fertilizers, the role of biofertilizers is becoming significant. India has been using rhizobium for leguminous crops and blue-Green algae for rice cultivation, but the consumption levels have been low. Keeping in view the vast Indian biodiversity, there is need to explore the same without damaging the fragile ecological balance. In India, although there is growing awareness about the ill effects of pollution, promotion of continual introduction of environmentally friendly products and methodologies in the chemical industry needs to be furthered. Usage of nonconventional technologies is highly popular in India. First in this list is the usage of microwaves, which is also the field of my research work. Further, the microwave chemists are turning their attention toward microwave-assisted dry-media reactions in order to minimize solvent usage, an added advantage to already established microwave chemistry.In addition to microwave-assisted reactions, ultrasonic and photochemical reactions are also used as nonconventional reaction technology. Analytical chemistry has been at the center of the Green chemistry movement. Advances in analytical chemistry are keys to environmental protection. In India, the focus for analytical chemistry is mainly on extraction technologies such as solid phase, ultrasound and microwave, supercritical fluid extraction, and automated soxhlet extraction. Monitoring and analysis of heavy metals and pesticides is very important for an agroeconomy-based country like India, and chief governmental institutes like the Indian Agricultural Research Institute (IARI) and the Defense Research and Development Organisation (DRDO) are working extensively in this field. Further removing of these elements from industrial and agrochemical usage is of prime importance for these institutes.

NONACADEMIC INITIATIVES

Industry in India still needs to make significant improvement from the environmental point of view. Most of the industrial R&D is mainly concerned with cost effectiveness rather than eco effectivemethods. Although there has been some collaborative work between academia and industries, still there is ample

opportunity for increased collaboration. There is immediate need for technology transfer from academic laboratories to industrial plants for meaningful application of Green research. The best examples are the applications of enzymes in various industries ranging from drugs to leather. The textile industry is one of the highly revenue generating industries in India, and they are now switching over to microbial decolorization and degradation. There is an increasing need of exploring biodiversity for natural dyes and developing eco-friendly methodology for synthetic dyes. All these require more funding in the R&D of respective fields and greater interaction and coordination between industry, academia, and government. Government can do a lot of good work for the cause of Green chemistry by increasing public awareness and by bringing and enforcing strict environmental legislations. One of the recent and controversial examples of government initiative is the conversion of diesel vehicles to compressed natural gas (CNG) in order to reduce pollution in Delhi, the capital city of India.. Relocation of industries into industrial areas away from residential parks is another bold step taken by the Delhi government. Further, the government is also concentrating on new projects such as fuel pellets from municipal waste, aspirated H-cylinder engines for Light Commercial Vehicles (LCVs), meeting India 2000 emission norms, battery-powered cars for pollution-free driving, hydrogen energy and energy towers for new environment-friendly fuel, development of traditional herbal drugs as adaptogens and immunomodulators. The government should also increase funding to encourage research in Green chemistry. By introducing Green chemistry education at all levels, the government can build a solid foundation toward Green chemistry in India.

INDIAN CHAPTER

The recently constituted Green Chemistry Chapter of India has already started working to popularize Green chemistry in India. As a part of environmental movement, a National Workshop on Green Chemistry was organized by the Department of Chemistry, University of Delhi in March 2009 to bring together all who are practicing Green chemistry in India for the first time. For Green chemistry education, a refresher course was organized for college teachers by the Centre for Professional Development in Higher Education in University of Delhi. Inspired by the overwhelming response of participants in these events, recently an IUPAC International Symposium on Green Chemistry was organized by the Department of Chemistry, University of Delhi, which proved to be an excellent event for researchers' world over to interact on the one common platform. Green Chemistry Chapter of India was constituted recently to expand its domain. Some future activities under the banner of the Green Chemistry Chapter of India have been planned. Top on the priority list is to spread the awareness of Green chemistry among researchers and young students by means of workshops, conferences scholarships, and awards. Simultaneously, there is a need to encourage industries to collaborate with academia and government for effective practice of Green chemistry. Another aim of the Green Chemistry Chapter of India is to encourage global partnership for effective environmental management.

CONCLUSION

Finally, on foregoing facts, the practicing of Green chemistry in India is a necessity rather than an option, as this is now a high time to protect our caring environment from further damage. The future of Green India is in the hands of young researchers and students, as the practice of Green chemistry is a moral responsibility for them. Government agencies should enforce the laws strictly to practice Green chemistry. Industries should also understand their moral responsibility toward the fragile environment.

REFERENCES

1. Anastas,P. T and Warner,J. C., Green Chemistry, Theory and Practice, Oxford University Press, New York, 1998.

2. Hall , J. A., Vuocolo, L. D., Suckling, I. D., Horwtz, C. P., Allison, R. W., Wright, L. J. and Collins. T. A New catalysed hydrogen peroxide bleaching process. Proceedings of the 53rd APPITA Annual Conference April 19-22, 1999, Rotorua, New Zealand.

3. Kidwai, M. 2007. New Trends in Green Chemistry, Anamaya Publishers, New Delhi.

4. Bhojak, N., Gudasaria, D. D., Khiwani, N. and Jain, R. 2007. "Microwave Assisted Synthesis, Spectral and Antibacterial Investigation on Complexes of Mn (II) with Amide Containing Ligands", *E-Journal of Chemistry*,4, (2), pp. 232-237.

5. Bhojak, N., Khiwani, N. and Barwar, U. 2010. "Greenchemical Biotechnology: An InnovativeConcept, Theory and Practice", *Advances in Applied Biotechnology and Microbiology,* Chapter No. 16, pp. 95-99.
6. Collins, T. 2001. "Towards Sustainable Chemistry", *Science,* 291, pp. 5501-5548.
7. Garg, B. S. Bhojak, N. and Nadan, D. 2005. "Spectrofluorometric Determination of Lead in Natural Wastewater and Egg samples with N-(2' pyridyl)-2-hydroxybenzamide in Micellar Medium", *Ind. J. Chem.,* 44A, pp.. 504.
8. Grieco, P. A. 1988. Organic Synthesis in Water, Ed Blackie, London

DAM

V.K AHUJA ,

HOD (CIVIL ENGG.)Engg. Email: vkahuja@gmail.com ph. 9255589865

A **dam** is a barrier that impounds water or underground streams. The reservoirs created by dams not only suppress floods but provide water for various needs to include irrigation, human consumption, industrial use,aquaculture and navigability. Hydropower is often used in conjunction with dams to generate electricity. A dam can also be used to collect water or for storage of water which can be evenly distributed between locations. Dams generally serve the primary purpose of retaining water, while other structures such as floodgates orlevees (also known as dikes) are used to manage or prevent water flow into specific land regions.

Types of dams

- Gravity dam
- Arch-gravity dams
- Barrages
- Embankment dams
- Rock-fill dams
- Concrete-face rock-fill dams
- Earth-fill dams
- Saddle dam
- Weir
- Check dam
- Dry dam
- Diversionary dam
- Underground dam
- Tailings dam
- Steel dams
- Timber dams
- Cofferdams
- Natural dams
- Beaver dams
- **Common purpose**

Function	Example
Power generation	Hydroelectric power is a major source of electricity in the world. Many countries that have rivers with adequate water flow, that can be dammed for power generation purposes. For example, the Itaipu Dam on the Paraná River in South America generates 14 GW and supplied 93% of the energy consumed by Paraguay and 20% of that consumed by Brazil as of 2005.
Water supply	Many urban areas of the world are supplied with water abstracted from rivers pent up behind low dams or weirs. Examples include London – with water from the River Thames and Chester with water taken from the River Dee. Other major sources include deep upland reservoirs contained by high dams across

	deep valleys such as the Claerwen series of dams and reservoirs.
Stabilize water flow / irrigation	Dams are often used to control and stabilize water *flow*, often for agricultural purposes and irrigation.[52] Others such as the Berg Strait dam can help to stabilize or restore the water *levels* of inland lakes and seas, in this case the Aral Sea.[53]
Flood prevention	Dams such as the Blackwater Dam of Webster, New Hampshire and the Delta Works are created with flood control in mind.[54]
Land reclamation	Dams (often called dykes or levees in this context) are used to prevent ingress of water to an area that would otherwise be submerged, allowing itsreclamation for human use.
Water diversion	A typically small dam used to divert water for irrigation, power generation, or other uses, with usually no other function. Occasionally, they are used to divert water to another drainage or reservoir to increase flow there and improve water use in that particular area. See: diversion dam.
Navigation	Dams create deep reservoirs and can also vary the flow of water downstream. This can in return affect upstream and downstream navigation by altering the river's depth. Deeper water increases or creates freedom of movement for water vessels. Large dams can serve this purpose but most often weirs and locks are used.
Recreation and aquatic beauty	Dams built for any of the above purposes may find themselves displaced by time of their original uses. Nevertheless the local community may have come to enjoy the reservoir for recreational and aesthetic reasons. Often the reservoir will be placid and surrounded by greenery, and convey to visitors a natural sense of rest and relaxation.

Location

One of the best places for building a dam is a narrow part of a deep river valley; the valley sides can then act as natural walls. The primary function of the dam's structure is to fill the gap in the natural reservoir line left by the stream channel. The sites are usually those where the gap becomes a minimum for the required storage capacity. The most economical arrangement is often a composite structure such as a masonry dam flanked by earth embankments. The current use of the land to be flooded should be dispensable.

Dam failure

Dam failures are generally catastrophic if the structure is breached or significantly damaged. Routine deformation monitoringand monitoring of seepage from drains in and around larger dams is useful to anticipate any problems and permit remedial action to be taken before structural failure occurs. Most dams incorporate mechanisms to permit the reservoir to be lowered or even drained in the event of such problems. Another solution can be rock grouting – pressure pumping portland cementslurry into weak fractured rock.

The main causes of dam failure include inadequate spillway capacity, piping through the embankment, foundation or abutments, spillway design error, geological instability caused by changes to water levels during filling or poor), poor maintenance, especially of outlet pipes, extreme rainfall , earthquakes and human, computer or design

MODELING & SIMULATION OF A GRID INTERACTIVE PV POWER PLANT

Subham Gandhi[1], Satish Kumar[2]

[1]Associate Professor, SBMN Engg College, Rohtak, subhamsavailable@gmail.com
[2]Principal Jat polytechnic College, Kaithal, satishkumarsinghal68@gmail.com

ABSTRACT

This paper presents a simulation model of the electrical part of a grid interactive photovoltaic system. The model contains a detailed representation of the main components of the system that are the solar array, 3 phase inverter and LC filter inverter control circuit and load. A proper control of the DC/AC inverter is developed in order to synchronize the system to the grid. The grid interface inverter transfers the energy drawn from the PV array into the grid by keeping common DC coupling voltage constant. Besides that, the control techniques for inverter as well can be delved into. This research paper also focuses on the different PV technologies (amorphous silicon, polycrystalline) and their effect to the system in terms of energy output. Besides that, this research work would be designed nearly to the projects that have been implemented in India in order to verify the energy output results from the modeling and simulation activities. This model is a simplified approach of the system's individual modules.

Index Terms— PV array, inverter, grid interactive, multicrystalline , PV energy

I. INTRODUCTION

India is a developing and most densly populated country in the world. The electrical power is essentially required for the economic development of a country. The energy crisis is the biggest hurdle in the any country development. The constant cost increasing of fossil fuelas well as it has a negative impact on environment. The ever increasing demand of fossil fuel is a motivating force for researcher towards innovation of renewable energysources like wind energy, solar energy, tidal energy, geo themal energy. In photo voltaic system , the solar energy is converted in to electrical energy. The basic device of the PV system is the PV cell. Cells may be grouped to form a module and modules may be used to form a PV array. More precise applications require electronic converter to process the electricity from the PV device. These converters may be used for operation of PV array at maximum power point and supply the sully the AC sinusodsial current.. The mathematical model of the PV device is used here for analyzing a complete PV system and its component using the Matlab/Simulink simulator. Villalla et.al[4] proposed a mathematical model for photovoltaic module in which they found the parameters of non linear I-V equation by adjusting the curve at three point: open circuit, maximum power, and short circuit. They found the best I–V equation solution by an iterative process for the single-diode photovoltaic (PV) model including the effect of the series and parallel resistances, and warranties that the maximum power of the model matches with the maximum power of the real PV module. In this paper the above fallacies has been removed and a simulink model has been proposed for a real time full scale solar PV power plant and the energy output results has been tested with the data of an operating 14 kW working power plant. Simulation of the mathematical model has been done in Matlab/Simulink software.

Nowadays, renewable energy has been more and more attractive due to the severe environmental protection regulations and the shortage of conventional energy sources. Photovoltaic generation is the technique which uses photovoltaic cell to convert solar energy to electric energy. Photovoltaic energy is assuming increasingly importance as a renewable energy source because of its distinctive advantages, such as simple configuration, easy allocation, free of pollution, low maintenance cost, etc. However, the disadvantage is that photovoltaic generation

is intermittent, depending upon weather conditions. The current contribution of solar energy to the total India's energy needs is insignificant, in the medium and long run, it is expected that solar energy, especially solar PV will form a vital component of the country's energy mix. So the Indian government launched world's largest solar program Jawaharlal Lal Nehru National Solar mission (JNNSM), which includes the establishment of no. of such grid connected PV plants. This research work is towards analyzing of those power plants.The block diagram in fig.1 presents the proposed design for the grid interactive solar PV system. It consist of:

--P-V array that simulates the PV voltage and current output, depending on the temperature and solar irradiance,

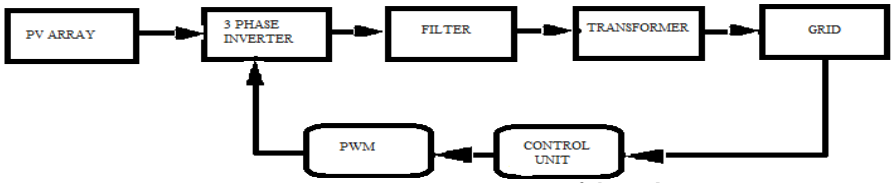


Fig. 1 Components of the grid interactive PV system

-- A 3-phase inverter which converts the generated DC power into a 3-phase AC,

-- A filter which is an LC low pass filter reducing the harmonic distortion by cutting off the high frequency harmonics.

-- A control unit in which a PLL synchronizes the output phase of the inverter with the phase of the grid and the PWM synchronizes the IGBTs.

II. BUILDING MODELING OF PV COMPONENTS

A. PV array

A photovoltaic cell is basically a semiconductor diode whose p–*n* junction is exposed to light [1]. Photovoltaic cells are made of several types of semiconductors using different manufacturing processes. The monocrystalline and polycrystalline silicon cells are the only found at commercial scale at the present time. The incidence of light on the cell generates charge carriers that originate an electric current if the cell is short circuited [1].

There are several models available for modeling of a practical photovoltaic cell. The general model consists of a current source, a parallel diode, a parallel resistor expressing leakage current, and series resistor describing an internal resistance to the current flow. In an ideal photovoltaic cell, there is no series loss and there is no leakage to the ground. That is, the series resistor has a value of zero while the parallel resistor has a value of infinity.

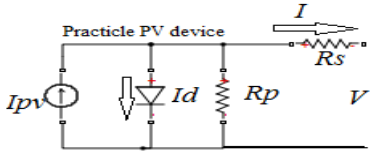


Fig.2 Single Diode Model

Figure 2 shows a single diode model of PV cell. Some authors have proposed more sophisticated models that present better accuracy and serve for different purposes. For example, in [6] & [7] an extra diode is used to represent the effect of the recombination of carriers. A three-diode model is proposed in [10] to include the influence of effects that are not considered by the previous models. For simplicity, the single diode model of Fig.2 is studied in this paper. This model offers a good compromise between simplicity and accuracy [8], and has been used by several authors in previous works, sometimes with simplifications but always with the basic structure composed

of a current source and a parallel diode [10], [13] and [14]. The current obtained from a photovoltaic module consisting of a number of cells (Ns) connected in series is represented by eq.(1).

$$I = I_{PV} - Io\left[exp\left(\frac{q(V+IRs)}{akTNs}\right) - 1\right] - \frac{V+RsI}{Rp} \qquad (1)$$

I_d

Where I_{PV} is the current generated by the incident light (directly proportional to the Sun irradiation); I_d is the Shockley diode equation, I_o is the reverse saturation or leakage current of the diode, q = 1.60217646 × 10−19 C (the electron charge), k =1.3806503 × 10−23 J/K (the Boltzmann constant), T is the temperature of the p–n junction (in Kelvin), a is the diode ideality constant, R_S is series resistance, R_P is shunt resistance. In the case of a number of modules connected in parallel (N_P), the current obtained from eq. (1) is multiplied by N_P. All PV array datasheets bring basically the following information: the nominal open-circuit voltage ($V_{oc,n}$), the nominal short-circuit current ($I_{sc,n}$), the voltage at the maximum power point (V_{mp}), the current at the maximum power point (I_{mp}), the open-circuit voltage temperature coefficient (K_V), the short circuit current temperature coefficient (K_I), and the maximum experimental peak output power ($P_{max,e}$). This information is always provided with reference to the nominal condition or standard test conditions (STCs) of temperature (25 °C) and solar irradiation (1000 W/m2). Some manufacturers provide I–V curves for several irradiation and temperature conditions.

The photovoltaic module used for the calculations in this paper is the BP 380 produced by BP Solar [18]. For simplicity, the single diode model will be studied in this chapter. This model offers a good compromise between simplicity and accuracy [8]. The light-generated current of the PV cell depends linearly on the solar irradiation and is also influenced by the temperature according to the following equation [8], [12]:

$$I_{PV} = \left(I_{PV,n} + K_I\,(T - T_n)\right)\frac{G}{G_n} \qquad (2)$$

$$I_{PV,n} = \left(I_{sc,n} + K_I\,(T - T_n)\right)\frac{G}{G_n} \qquad (3)$$

$$I_{SC} = \left(I_{sc,n} + K_I(T - T_n)\right)\frac{G}{G_n} \qquad (4)$$

Where G is irradiance on device surface; G_n = 1000 W/m2 (nominal solar radiation), T_n = 298.15 K (nominal temperature); K_I = + (0.065±0.015) % ºC, $I_{sc,n}$ = 4.8 A [18]. King et.al [11] found that there is typically less than a 5% change in the voltage coefficients over a tenfold change in irradiance from 100 W/m2 to 1000 W/m2. The temperature of the PV module can be obtained from the following equation [5]:

$$T - T_a = (219 + 832\bar{K})\,\frac{NOCT - 20}{800} \qquad (5)$$

Where T_a is the ambient temperature which can be obtained from the Metrological Department; NOCT °C (nominal cell operating temperature, obtained from the BP380 module datasheet); $\bar{K}$ is the monthly clearness index. The diode saturation current I_o and its dependence on the temperature may be expressed by the following equation [13]:

$$I_o = I_{o,n}\left[\frac{T_n}{T}\right]^3 \exp\left[\frac{qE_g}{ak}\left(\frac{1}{T_n} - \frac{1}{T}\right)\right] \qquad (6)$$

Where E_g is the band gap energy of the semiconductor = 1.12 eV.The nominal saturation current $I_{o,n}$ is obtained by evaluating eq. (1) at the nominal open-circuit condition, with V = $V_{oc,n}$ I= 0, and $I_{PV} \approx I_{sc,n.}$

$$I_{o,n} = \frac{I_{sc,n}}{exp\left[\frac{V_{oc,n}}{aV_{t,n}}\right] - 1} \qquad (7)$$

Where $V_{t,n} = \frac{kT_n}{a}$

The value of a is stated by Huan-Liang Tsai et.al [12] for different types of PV depending on the PV technology. For the calculations in this paper, the value of a will be taken equal to 1.2. The method used here to get R_S and R_P is very simple. The first iterative value of R_S is 0. Then the value of R_P will be calculated from eq. (8) using the values at the nominal conditions which is obtained from the module data sheet. The value of the maximum power is then obtained from the graph ($P_{max.m}$, using a computer code) and compared with the experimental maximum power ($P_{max,e}$). This process is then repeated while increasing the value of R_S by small increment (e.g. 0.01) until the value of the calculated maximum power and the experimental one are equal (or close to each other within a certain tolerance, e.g. 0.001) [4]. It is worth noting that the values of both R_S and R_P obtained are for the nominal conditions. However, the changes in their values due to the temperature changes are small and can be neglected.

$R_p =$

$$\frac{V_{mp}\left(V_{mp} + I_{mp}R_S\right)}{\left[V_{mp}I_{PV-}V_{mp}I_o exp\left[\frac{V_{mp} + I_{mp}R_S}{N_s a}\frac{a}{kT}\right] + V_{mp}I_o - P_{max,e}\right]} \qquad (8)$$

B. **Inverter**

Grid-connected inverters are necessary for dc–ac conversion. To avoid the distortions to the power grid, the generated currents from these inverters are required to have low harmonics and high power factor. When the output currents are in phase with the grid voltages, the maximum real output power is achieved. For conversion of dc-ac in grid connected PV system, there are several types of topologies and inverter designs are used in existing installations [2]. There are still some subjects as yet unproven. Reliability, life span, and maintenance needs should be certified through long-term operation of a PV system. Further reductions of cost, size, and weight are required for the PV systems. The main circuit is the part where the DC electric power is converted to AC. This is virtually implemented with the one that is shown at the Fig.4. In this circuit we use a 3 leg inverter for 3-phase conversion which is composed of 6 IGBT. Control unit generates control pulses to drive the IGBTs.. The frequency of the IGBTs we use is 15 KHz. For the time interval the IGBTs are open, we get a pulse at power circuit, which has the same amplitude of source, as it can be seen at the fig.[3]. A three phase inverter has the basic advantage that generates power in 3-phase and is working without a hitch.

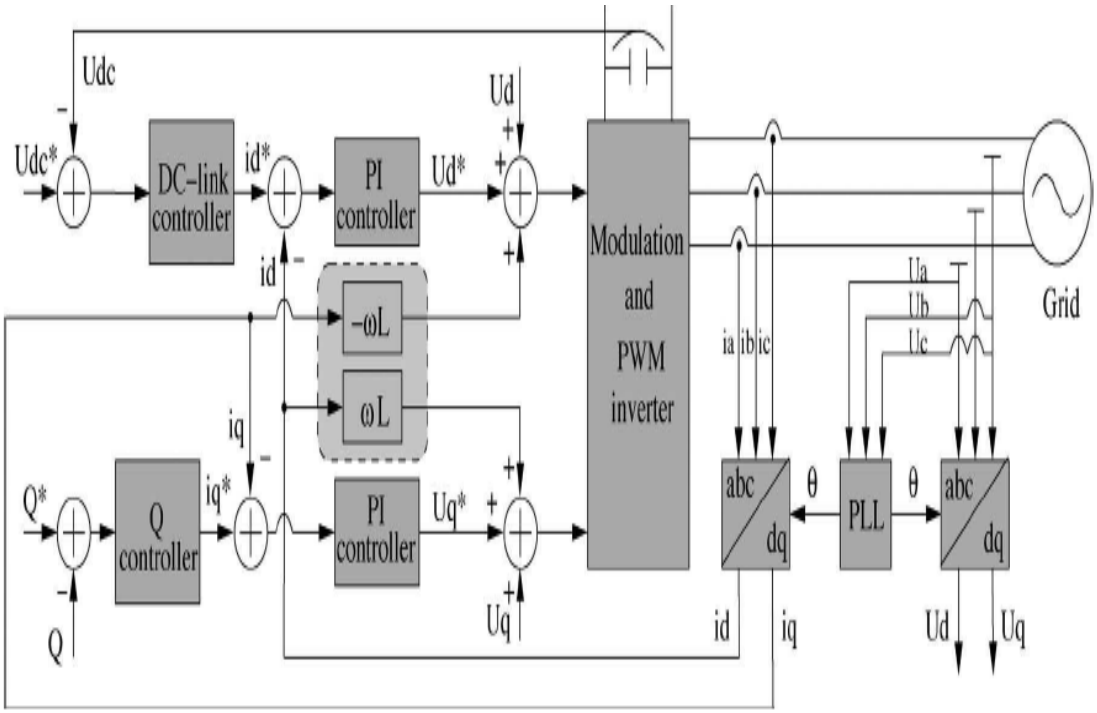


Figure 3 General structure for inverter control unit

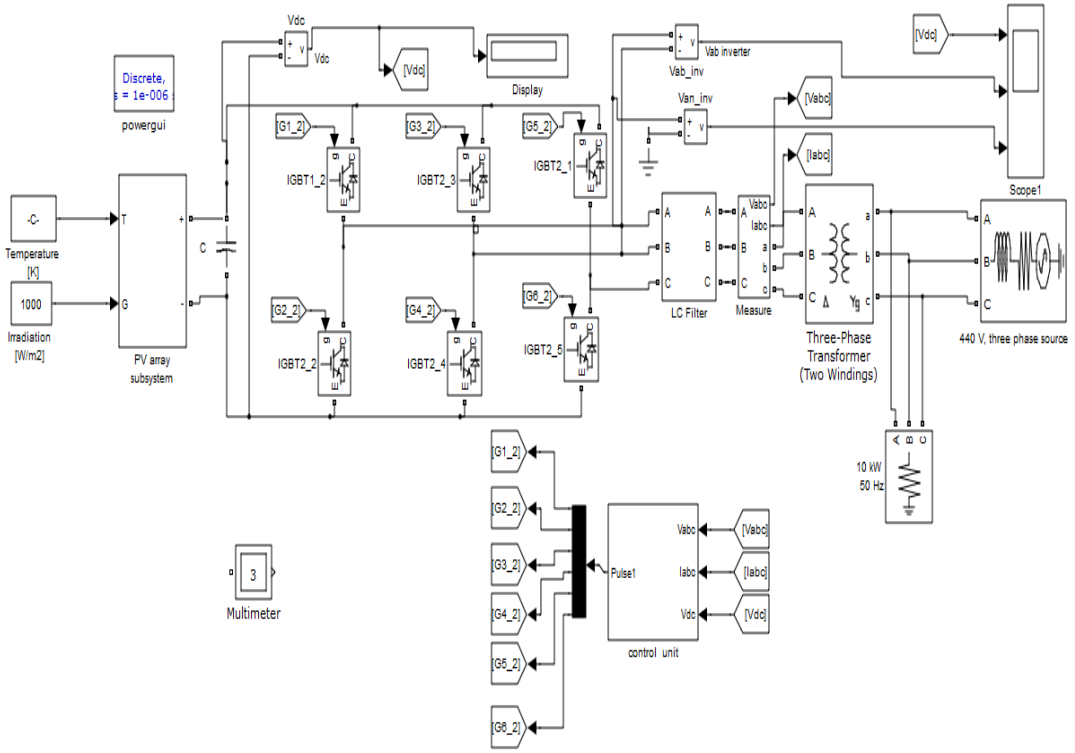


Figure 4 Matlab/Simulink of grid connected PV system

C. Control Unit, Filter, Load, and Grid utility

The control unit combines a number of functions. At first, it provides the appropriate pulses to the IGBTs for the PWM technique, Secondly, it synchronizes the inverter generated phases to the grid. The control circuit has been developed in Matlab/Simulink.Some of the control strategies are given in [5].We have to enhance the functionality of the control circuit with the use of dynamic control. The structure for control and synchronization of inverter is shown in figure3.In order, this simulation give us realistic and accurate output data for each day, of the simulated days from the PV array, we have to simulate the grid utility and the load. We use a high resistance for a preliminary house load estimation of about 10KW. The load we use at the simulation can be seen at fig. [4]. Grid utility is simulated using a three phase source from simpower system library. Instead of this Grid utility can also be simulated using impedance estimation. Such values have been experimentally measured in [17]. The Grid utility is not presented in this contribution. The idea to have a simulation that will work for each day and give us data for such a system has been studied and will be fulfilled at the future.. The control unit uses two regulators: an inner current loop controlling the current and an outer DC voltage regulator controlling the DC voltage. LC filters are used to reduce harmonic (harmonic frequencies around multiples of 15 kHz). The LC filter we use can be seen at fig. [4], with L=5mH and C=20μF.

III. ENERGY MODELING

The array is characterized by its average efficiency, η_p, which is a function of average module temperature T_c and is given by

$$\eta_p = [1 - \beta_p(T_c - T_r)] \qquad (9)$$

where η_r is the PV module efficiency at reference temperature Tr (= 25°C), and β_p is the temperature coefficient for module efficiency. Tc is related to the mean monthly ambient temperature Ta through Evans' formula (Evans, 1981) as given in equation (5). η_r, NOCT and β_p depend on the type of PV module considered. They can be entered by the user or, for standard technologies, are assumed to take the values given in [3]. The energy delivered by the PV array, E_P , is simply:

$$E_p = \eta_p(A * \bar{H}_t) \qquad (10)$$

Whereby given:

η_p = array average efficiency

A = area of the array

$\bar{H}_t$= solar radiation

E_p = energy delivered by the PV array

It has to be reduced by "miscellaneous PV array losses" λ_P and "other power conditioning losses" λ_C

$$E_A = E_p(1-\lambda_P)(1-\lambda_C) \qquad (11)$$

Whereby given:

λ_P= miscellaneous PV array losses

λ_C = other power conditioning losses

E_p = energy delivered by the PV array

E_A = energy available to the load

The energy available to the grid is what is produced by the array, reduced by inverter losses:

$$E_{GRID} = E_A * \eta_{INV} \qquad (12)$$

Whereby given:

E_A = energy available to the load

η_{INV}= inverter efficiency

Depending on the grid configuration not all this energy may be absorbed by the grid. The energy actually delivered is:

$$E_{DLVD} = E_{GRID} * \eta_{ABS} \quad (13)$$

Whereby given:

E_{GRID} = energy delivered to load

η_{ABS} = PV energy absorption rate

E_{DLVD} = energy available to the load

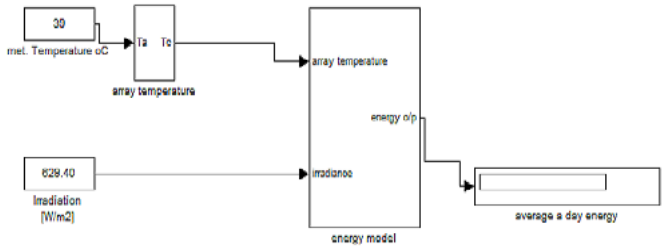


Figure 3 Subsystem Implementation of Energy model

Assumptions

In the energy modeling of the whole of the system it is assumed that the power conditioning losses (λ_C) are 5%, miscellaneous PV array losses are (λ_P) are also considered 5%, and grid absorption rate is 95%. The efficiency of the inverter is taken 95%.

IV. SIMULATION RESULTS AND DISCUSSION

In this chapter, the simulation of single junction PV module will be viewed and verified through the manufacturer's datasheet. The next stage will be simulation of the grid connected PV system and analysis through energy output and also verification through actual case studies. Case study is conducted on 14 kW multicrystalline PV power plant. While energy output yielded from implementing the Case study using MATLAB / SIMULINK were analyzed and later will be verified using the actual monitored data from the PV power plant. The simulation results validate the manufacturer's data sheet. The irradiance data is measured at every 15 minutes from 10:00 am to 5:00 pm. The simulation is conducted on a week data. Figure 3 & 4 gives I-V and P-V curve at Standard Test Condition. Figure 5 & 6 shows I-V and P-V curve at different irradiance level. The results are validating the literature theory.

Figure 7 shows the difference between the actual and simulated energy output. The difference is not so large so the model can be used for any capacity of power plant by changing the parameters. Beyond this the effect of temperature on efficiency can be observed.

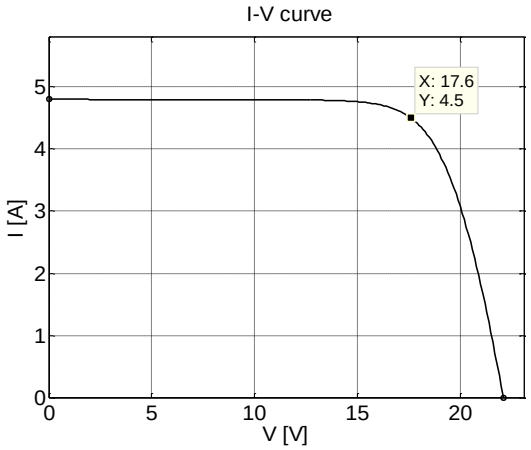


Figure 3: I-V curve at Standard Test condition

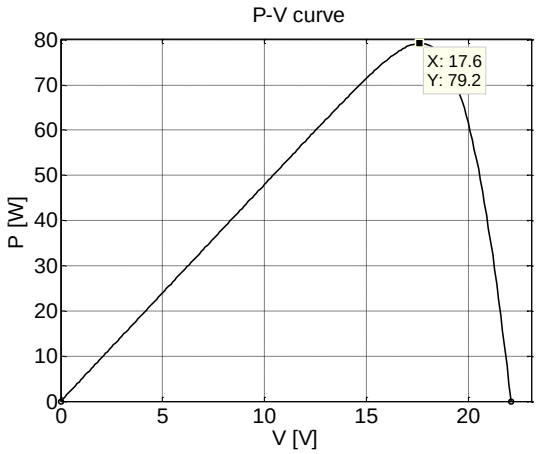


Figure 4: P-V curve Standard Test Condition

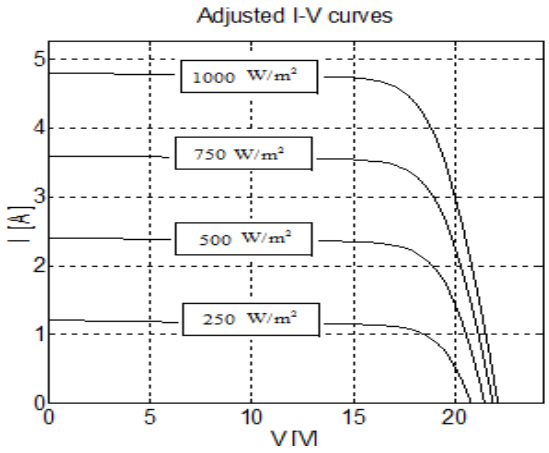


Figure 5: I-V curve at constant temperature (25°C) and different irradiance

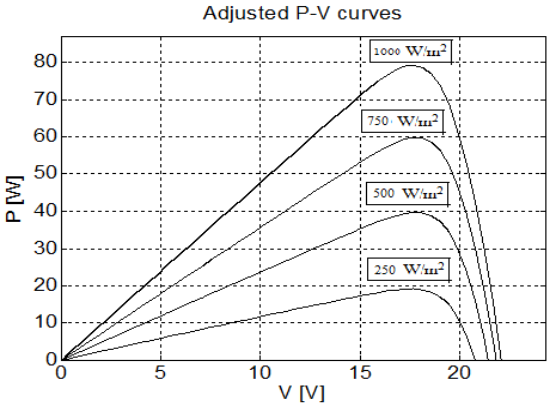


Figure 6: P-V curve at constant temperature (25°C) and different irradiance.

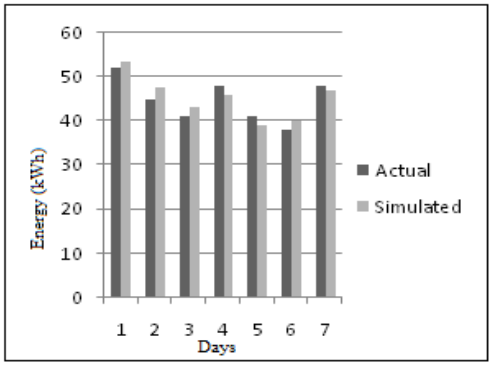


Figure: 7 Comparison between actual energy Vs Simulated energy outputs

Figure 8 shows a comparative idea of energy delivered by both the PV technologies. This energy simulink subsystem is used for energy comparison of different PV technology. Which clearly shows that at same environmental conditions the efficiency of multicrystalline PV module is more as compare to the a-Si technology.

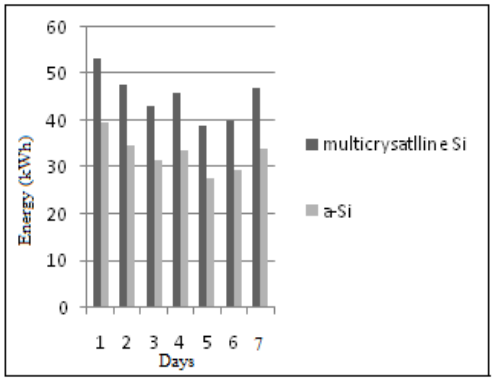


Figure 8 Comparison of energy delivered by multicrystalline Si and a-Si technology

Overall findings indicate that this MATLAB modeling can be further used for investigation and make improvement in order to identify which best technologies to be implemented. Another conclusion that can be drawn, the multicrystalline PV System yields higher energy output compared to the amorphous silicon technology.

From simulink model of figure 4, we can analyze the waveforms at any point in the circuit. The output across the dc coupling capacitor is shown in figure 9. The generated dc voltage is 220 volt. The dc coupling capacitor smoothens the generated voltage. Figure 10 is the PWM output line voltage of the inverter, which has large harmonic components. Filter reduces harmonics up to a desirable level which is sufficient for interconnection of grid as per the CERC regulation. Figure 10 is the phase voltage. The simulation results are at Standard Test Condition. These may vary as the environmental condition varies. The current injected in the grid is shown in fig.13. the total harmonic distortion is 4.78% which is obtained from fft analysis.

Figure 9 DC coupling voltage

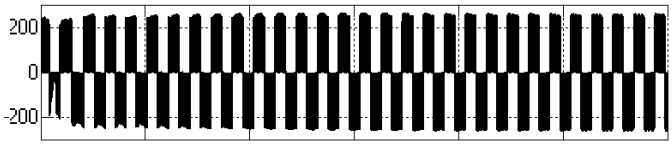


Figure10: Phase to phase output of inverter

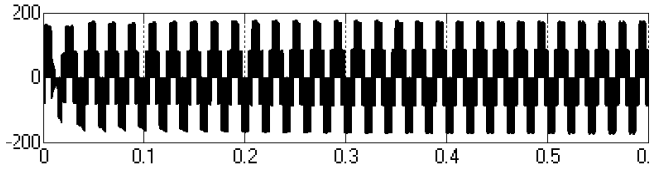


Figure 11: Phase to ground voltage

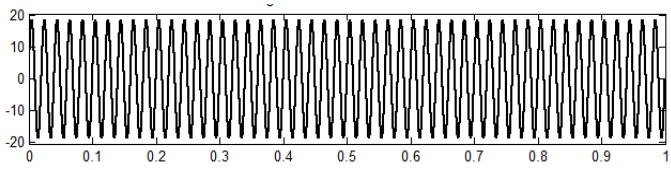


Figure 13: Current injected in the grid

IX. CONCLUSION

This paper presents the model and simulation of a grid interactive PV plant and analyzes its electrical performance for different operating conditions. Overall system is built by integrating individual models of a PV array, interfaces, and load. Simulation results show that the overall simulink model of the grid interactive PV plant performs well to reflect the system behavior for different operating scenarios. This model does not require storage of the PV module characteristics as data storage but calculates the power generated and energy delivered by the PV module at different solar irradiance. The developed PV model gives properly the I-V and P-V characteristics of a photovoltaic module. This system is able to take into account the variations in the environmental conditions such as temperature and irradiance.

The model analyzes the impact of different technologies on the PV system. By this multicrystalline technology is proven higher than amorphous silicon technology. Thus in selection of the best PV technology actually it depends on the usage and also niche area on where the PV system would be applied. Finally through verification using the real time system, the grid connected model can be used for energy output prediction and it is a user friendly system.

REFRENCES

[1] Darren M. Bagnall, Matt Boreland, " *Photovoltaic Technologies",Energy policy, Vol 36, pp. 4390-4396, 2008*.

[2] Power electronics by M.H.Rashid

[3] Trishan Esram Patrik L, Chapman, " comparison of photovoltaic Array Maximum Power Point Tracking Technique", IEEE Transanction on Energy Conversion, Vol 22, No.2 pp 439-499,2007.

[4] J. Bratt, "Grid connected PV inverter, Modeling and Simulation"Master of Scince in Electrical Engineering , Sag Diego State university, 2011.

[5] Amal A. Hassan, Faten H. fahmy,Abd Ei-Shafy A Nafeh, Mohamed A El Sayed, " Modeling and simulation of a single phase grid connected photo voltaic system"WSEAS TRANSACTION on SYSTEM and CONTROL issue 1,vol 5 2010.

[6] J.A. Gow and C. D. Manning, "Development of a photovoltaic array model for use in power-electronics simulation studies," *IEE Proc. Elect. Power Appl.*, vol. 146, no. 2, pp. 193–200, 1999.

[7] J.Hyvarinen and J. Karila, "New analysis method for crystalline silicon cells," in *Proc. 3rd World Conf. Photovoltaic Energy Convers.*, 2003,vol. 2,pp.1521-1524.

[8] C. Carrero, J.Amador, and S. Arnaltes, "A single procedure for helping PV designers to select silicon PV module and evaluate the loss resistances,"*Renewable Energy*, vol. 32, no. 15, pp. 2579–2589, Dec. 2007.

[9] D. Sera, R. Teodorescu, and P. Rodriguez, "PV panel model based on datasheet values," in *Proc. IEEE Int. Symp. Ind. Electron. (ISIE)*, 2007,pp. 2392–2396.

[10] W. Xiao, W. G. Dunford, and A. Capel, "A novel modeling method for photovoltaic cells," in *Proc. IEEE 35th Annu. Power Electron. Spec. Conf. (PESC)*, 2004, vol. 3, pp. 1950–1956.

[11] King, D. L., Kratochvil, J. A., and Boyson, W. E., 1997, "Temperature Coefficients for PV Modules and Arrays: Measurement Methods, Difficulties, and Results," Proc. 26th IEEE Photovoltaic Specialists Conference, Anaheim, CA, September 29–October 3, pp. 1183–1186

[12] Huan-Liang Tsai, Ci-Siang Tu, and Yi-Jie Su, Member, IAENG, Development of Generalized Photovoltaic Model Using MATLAB/SIMULINK, Proceedings of the World Congress on Engineering and Computer Science 2008

[13] K.Nishioka,N. Sakitani,Y. Uraoka, and T. Fuyuki, "Analysis of multicrystalline silicon solar cells by modified 3-diode equivalent circuit model taking leakage current through periphery into consideration," *Solar Energy Mater. Solar Cells*, vol. 91, no. 13, pp. 1222–1227, 2007.

[14] W. De Soto, S.A.Klein, and W. A. Beckman, "Improvement and validation of a model for photovoltaic array performance," *Solar Energy*, vol. 80, no. 1, pp. 78–88, Jan. 2006.

[15] E. Koutroulis, K. Kalaitzakis, and V. Tzitzilonis. (2008). Development of a FPGA-based system for real-time simulation of photovoltaic modules, *Microelectron. J.* [Online].

[16] W.Yi-Bo,W. Chun-Sheng, L. Hua, and X.Hong-Hua, "Steady-state model and power flow analysis of grid-connected photovoltaic power system," in *Proc. IEEE Int. Conf. Ind. Technol. (ICIT'08)*, pp. 1–6.

[17] L. Asiminoaei, R. Teodorescu, F. Blaabjerg, U. Borup, *A digital controled pv inverter with grid impedance estimation for ens detection*, IEEE Transactions on powe
r electronics, VOL 20, NOV 2005.

SUMMER WEATHER EFFECT ON CEMENT MORTAR

Ajay Sharma

*Civil Engineering Department, Shri Baba Mastnath Engineering College, Asthal Bohar, Rohtak, Haryana, India
aajay.055@ymail.com

When water is added to cement, paste is formed which gradually stiffens and then hardens. The stiffening of cement paste is called setting. Basically, setting is a process of transformation from an initial state, a scattered concentrated suspension, to a final state, a connected and strengthened system of particles. This transformation in the practice of cement and concrete is obtained by chemical reactions between cement particles and water (i.e., cement hydration). Normal setting of cement is associated with the hydration of Alite (impure C3S) and formation of the calcium silicate hydrate (CSH) phase. Cement paste / concrete sets gradually under the standard laboratory conditions (temperature ~230C and relative humidity not less than 90%) but outside the laboratory concreting has to be done under the prevailing climatic conditions. In some countries including the author's country concrete is subjected to summer weather, which is defined as "any combination of high air temperature, low relative humidity, wind velocity and intensities of solar radiation tending to adversely affect the quality of fresh and hardened concrete Retarding admixture is an admixture that retards the setting of concrete. A retarding admixture causes cement set retardation by one or more of following mechanisms:

(1) Adsorption of the retarding compound on the surface of cement particles, forming a protective skin which slows down hydration.
(2) Adsorption of the retarding compound on to nuclei of calcium hydroxide, poisoning their growth, which is essential for continued hydration of cement after the end of induction period.
(3) Formation of complexes with calcium ions in solution.
(4) Precipitation around cement particles of insoluble derivatives of the retarding compounds formed by reaction with the highly alkaline aqueous solution, forming a protective skin.

MATERIAL WORK

CEMENTS

Three different types of cements were used for setting time tests. These are denoted as type-A, type-B, and type-C. Their oxide and compound composition and some other properties provided by the manufacturers are given -A cement is obtained by adding 6-20% calcined clay to the normal Portland cement clinker during manufacturing while in type-B cement the calcined clay ranges from 21 to 35%. Their compound composition cannot be calculated by using Bogue's or other such formula.

MIXING WATER AND RETARDING ADMIXTURE

Normal tap water was used as mixing water. The retarding admixture used was ASTM C 494 type D admixture. Its density was about 1.02 mg/ml and its chloride content was claimed nil. The amount of the admixture incorporated into the paste was expressed in m//100g of cement indicated as percentage.

MIXES

Cement pastes were prepared for determination of consistency and setting times tests. The cement content and w/c ratios were kept constant for all tests for a given cement type. The amounts of cement and water used per test.

EQUIPMENT

A Vicat apparatus was used for determination of both the standard consistency and setting times of pastes. The apparatus was similar to that recommended by the ASTM C 187-77

and C 191-77 except the minor difference in the needle and ring (mold) dimensions. The needle of the apparatus was 1.13mm in diameter and 46mm long. The ring had an inside diameter of 90mm at the base and 80mm at its top.

DETERMINATION OF STANDARD CONSISTENCY AND SETTING TIMES

For standard consistency determination, the procedure of the ASTM C 187-77 was followed and for setting time determination, the Turkish Standard 19 (TS-19) was followed. The TS-19 nearly follows the ASTM C 191-52 procedure with minor amendments as described below: The initial set is said to have taken place when the needle (1.13mm dia.) of theVicat apparatus ceases to pass 3-5 mm above the bottom of cement paste taken in the Vicat mould. Final set is said to have occurred when the needle penetrates the cement paste to a maximum depth of 1mm. In both cases, the setting time is reckoned from the moment when mixing water is added to the cement.

CURING CONDITIONS

In order to simulate the approximate normal and adverse outdoors climatic conditions, the following three categories of curing conditions were provided to the test specimens:

(1) First curing condition (CC-I): Temperature = 220C, Relative Humidity = 55-65%

(2) Second curing condition (CC-II): Temperature = 350C, Relative Humidity = 35-45%

(3) Third curing condition (CC-I): Temp = 500C, Relative Humidity = 25-35%

For maintaining the desired curing conditions, a temperature controllable cabinet was used. The required relative humidity at various temperatures was obtained by placing saturated salt solutions (sodium nitrate at 220C, potassium carbonate at 350C and potassium chloride at 500.

TEST RESULTS AND DISCUSSION

Setting time tests with varying admixture contents were performed under the specified curing conditions. An average of three test readings was taken as the final reading. To compare the changes occurred in setting times by incorporation of the admixture, the setting time of cement paste with out admixture content under CC-1 was used as reference. The setting times were recorded in minutes. These results are presented in the following tables and figures. Setting time tests with varying admixture contents were performed under the specified curing conditions. An average of three test readings was taken as the final reading. To compare the changes occurred in setting times by incorporation of the admixture, the setting time of cement paste with out admixture content under CC-1 was used as reference. The setting times were recorded in minutes. These results are presented in the following tables and figures. The temperature effects on setting times in the range of 22 – 350C are greater than in the range 35 - 500C. For example, for the type-A cement paste without admixture, the initial setting time were reduced by about 40% when comparing 35 to 220 C and 21% when comparing 50 to 350C.

CONCLUSIONS

(1) High temperature and low humidity accelerated the setting of cement pastes for all mixes with and without the retarding admixture.

(2) The retarding admixture successfully retarded cement setting under each curing condition.

(3) The retarder showed lower retarding tendency at higher temperatures and lower humidity.

(4) The loss in setting times (with respect to the reference setting times) at 35oC was recovered by adding 0.125% of the admixture to the mix while at 500C, it was recovered by adding 0.25% of the admixture.

(5) With the type-B cement, the admixture showed accelerating effects on initial set. So, caution is needed when using retarders with pozzolanic type cements.

AN ARTIFICIAL NEURAL NETWORK BASED APPROACH FOR DOS ATTACKS DETECTION IN MANET

Anil Dudy

Department of Electronics and Communication Engineering, Baba Mast Nath College of Engineering and Technology, Rohtak, Haryana, India.

ABSTRACT:

Denials of Service (DoS) attacks are a serious threat for the prominent e-commerce internet sites such as Amazon, CNN, E*Trade, Yahoo and eBay. DoS attacks can consume memory, CPU, and network resources and damage or shutdown the operation of the resource under attack (victim). The quality of service enabled networks (QoS), which offer different levels of service, is vulnerable to such DoS attacks. Denial of service (Dos) is a type of attack in which a hacker issues a huge amount of packets to congeal specific servers' services, consequently.

Keywords: Artificial Neural Network(ANN), Delay of Service (DOS), Feed Forward Neural Network, Intrusion Detection Sytsem (IDS), Network Security

INTRODUCTION

Mobile ad hoc networks (MANETs) are dynamic mobile networks that can be formed in the absence of any pre-existing communication infrastructure. In addition to node mobility, a MANET is characterized by limited resources such as bandwidth, battery power, and storage space. The underlying assumption in MANETs is that the intermediate nodes cooperate in forwarding packets. However, this assumption does not hold in commercial and emerging civilian applications. MANETs are vulnerable to Denial of Service (DoS) due to their salient charac-teristics. There is a need to provide an incentive mechanism that can provide cooperation among nodes in the network and improve overall network performance by reducing DoS attacks. DoS attacks committed by selfish and malicious nodes were investigated. Our scheme motivates nodes to cooperate and excludes them from the network only if they fail to do so. We evaluated the perfor-mance of our scheme using the packet delivery ratio, the routing and communication overhead, and misbehaving node detection in a discrete event-simulation environment. The results indi-cate that a reputation-based incentive mecha-nism can significantly reduce the effect of DoS attacks and improve performance in MANETs.In MANETs, nodes act as both routers and ordinary nodes. Due to dynamic network topology and lack of centralized infrastructure, network security has brought a new challenge to networking communities. Unlike traditional net-works, MANETs are more vulnerable to DoS attacks due to limited resources that force nodes to be greedy in resource utilization. When there is no cooperation, activities of even a small number of nodes may significantly decrease the performance of the network. For example, a mis-behaving node that discards any packets passing through it can result in repeated retransmissions, which in turn cause network congestions. Also, a wireless link does not provide the same protection for data transmissions as does its wired link counter part. Hence, any user or receiver within the transmissions range can eavesdrop or interfere with data packets or routing information. Battery power is another critical resource for mobile nodes.If the battery power has been used up due to malicious attacks such as the sleep deprivation attack, the victim will not be able to provide network services. Since all nodes can be mobile, changes in network connectivity and resource availability also expose a network to various attacks. This calls for detection and prevention of attacks in the network (Mieso K Denko, 2010).

Although MANETs are based on the fundamental assumption that the nodes will cooperate in providing services or sharing available resources, non-cooperation is a critical problem when deploying these networks for civilian applications. Lack of cooperation in MANETs can be a result of misbehaving nodes or lack of sufficient resources. Misbehaving nodes can either be

malicious or selfish. Selfish nodes are nodes that participate in the network to maximize their own benefit by using network resources while saving their own resources. Malicious nodes directly attack a network by disrupting its normal operation. The absence of a trusted third party in ad hoc networks necessitates the development of protocols for collecting, storing, and distributing reputations. Enhancing cooperation among nodes in the network can help in detecting and mitigating DoS attacks caused by the misbehaving nodes (Mieso K Denko, 2010).

DOS ATTACK SCENARIOS

The DoS attacks that target resources can be grouped into three broad scenarios. The first attack scenario targets Storage and Processing Resources. This is an attack that mainly targets the memory, storage space, or CPU of the service provider. Consider the case where a node continuously sends an executable flooding packet to its neighbourhoods and to overload the storage space and deplete the memory of that node. This prevents the node from sending or receiving packets from other legitimate nodes. Neighbour-hood watch and monitoring can prevent the occurrence of such events by gradually excluding such malicious nodes. The second attack scenario targets energy resources, specifically the battery power of the service provider. Since mobile devices operate by battery power, energy is an important resource in MANETs. A malicious node may continuously send a bogus packet to a node with the intention of consuming the victim's battery energy and preventing other nodes from communicating with the node. The use of localized monitoring can help in detecting such nodes and preventing their consequences. The third attack scenario targets bandwidth.Consider the case where an attacker located between multiple communicating nodes wants to waste the network bandwidth and disrupt connectivity. The malicious node can continuously send packets with bogus source IP addresses of other nodes, thereby overloading the network. This consumes the resources of all neighbours that communicate, overloads the network, and results in performance degradations. Such attacks can be prevented based on the reputation information exchanged among the involved nodes or the cluster head. We attempt to prevent both selfish and malicious nodes from degrading network performance by providing incentives to encourage cooperation and punishing nodes that do not cooperate (Mieso K Denko, 2010).

back propagation network with one layer of z-hidden units. The Y output unit has W_{ok} bias and Z hidden unit has V_{ok} as bias. It is found that both the output units and the hidden units have bias. The bias acts like weights on connection from units whose output is always 1. This network has one input layer, one hidden layer and one output layer. There can be any number of hidden layers. The input layer is connected to the hidden layer and the hidden layer is connected to the output layer by means of interconnection weights. The bias is provided for both the hidden and the output layer, to act upon the net input to be calculated (Amit Garg and Ravindra Pratap Singh, 2013).

ARTIFICIAL NEURAL NETWORK

Artificial neural networks born after McCulloc and Pitts introduced a set of simplified neurons in 1943. These neurons were represented as models of biological networks into conceptual components for circuits that could perform computational tasks. The basic model of the artificial neuron is founded upon the functionality of the biological neuron (Afrah Nazir, 2013). By definition, "Neurons are basic signaling units of the nervous system of a living being in which each neuron is a discrete cell whose several processes are from its cell body. One can differentiate between two basic types of networks, networks with feedback and those without it. In networks with feedback, the output values can be traced back to the input values. However there are networks wherein for every input vector laid on the network, an output vector is calculated and this can be read from the output neurons. There is no feedback. Hence only, a forward flow of information is present. Network having this structure are called as feed forward networks. There are various nets that come under the feed forward type of nets. A multilayer feed forward

TRAINING PROGRAM IN MATLAB

Now from these different values of Start Time, Duration and Service a neural network is trained through a x: which is a input training vector as. Input training vector in this case is Start Time, Duration and Service. Output target vector t is Attack or Normal Traffic in this case. If it is a Normal traffic output t is 0 if it would be attack it would be 1. Weight W and v are initialized to small random values. W_0 and V_0 are bias. Initially Neural Network is trained using 18 different values of Start Time, Duration and Service. Iterations done in this program are 990000.

```
clc;
clear;
W=[-3.2184; -12.5463; 0.6328];
Wo=[9.5708];
v=[0.8153 -0.0614 0.9105; 6.6872 19.4263 0.2000; 0.1294 0.0324 0.2064];
vo=[0.4346 0.8483 -0.4499];
x=[.020 .020 .040 .040 .040 .040 .008 .008 .015
.015 .030 .030 .035 .035 .002 .002 .004 .004;
     .004 .004 .010 .010 .015 .015 .025 .025 .002
.002 .002 .002 .002 .002 .010 .010 .020 .020;
    .005 .010 .005 .010 .005 .010 .005 .010 .005
.010 .005 .010 .005 .010 .005 .010 .005 .010]; t= [1 1 0 1 1 0 1 1 0 1 0 1 1 1 0 0 1 0 ]; epoch=1;
alpha=.3;
while(epoch<990000)  for I=1:18
for i=1:3 zin1=0; for j=1:3
zin1=zin1+x(j,I)*v(j,i); end
zin(i)=zin1  + vo(i)*1; z(i)=1/(1+exp(-zin(i))); end
yin1=0; for i=1:3
yin1=yin1+z(i)*W(i,1); end
yin1=yin1+Wo; y(I)=1/(1+exp(-yin1)); delta1=(t(I)-y(I))*(y(I))*(1-y(I)); for
i=1:3 delin(i)=delta1*W(i,1); del(i)=delin(i)*z(i)*(1-z(i)); end
for i=1:3 for j=1:3
delv(i,j)=alpha*x(i,I)*del(j);
v(i,j)=v(i,j)+delv(i,j); end delvo(i)=alpha*1*del(i);
vo(i)=vo(i)+delvo(i); end
for i=1:3 delW(i)=alpha*z(i)*delta1; W(i,1)=W(i,1)+delW(i); end
delwo=alpha*1*delta1; Wo=Wo+delwo;
end epo(epoch)=epoch; error(epoch)=t(5)-y(5); n(epoch)=y(5);
epoch=epoch+1;
end figure
plot(epo,error,'r'); xlabel('Epoch Nubmer'); ylabel('Error');
title('Plot between Epoch and Error'); figure
plot(epo,n,'b');
title('Plot between Epoch and Output Value'); y
epoch W Wo
v vo
```

Output of the Training Program are Weights (W and V) and Bias (W_0 and V_0). Now using these values of Weights and Bias in Neural Network at any values of Start Time, Duration and Service and the condition of Attack or Normal Traffic can be easily forecasted through a program in MATLAB

FORECASTING PROGRAM IN MATLAB

```
clc;
clear;
```

```
W= [69.7724; -92.1417; 75.1616]; Wo=[5.7396];
v=[-101.2418 -48.9699 238.5258; 180.6523 215.8780 -0.1458; -245.0327 -
305.1403 9.9378]; vo=[ 0.9382 4.7101 -1.8095];
x=[.130; .570; .2100]; I=1;
for i=1:3 zin1=0; for j=1:3
zin1=zin1+x(j,I)*v(j,i); end
zin(i)=zin1 + vo(i)*1; z(i)=1/(1+exp(-zin(i))); end
yin1=0; for i=1:3
yin1=yin1+z(i)*W(i,1); end
yin1=yin1+Wo; y(I)=1/(1+exp(-yin1)); y(I)
```

Now using this MATLAB Program at values of Start Time, Duration and Service Condition of ATTACK or NORMAL TRAFFIC is forecasted. A backpropagation neural network was used in this research. The parameter selection of parameters and the parameter settings for the training of the neural network is of great importance. The neural network was tested with different parameter settings.

DARPA INTRUSION DETECTION EVALUATION

The Information System Technology Group at Massachusetts Institute of Technology - Lincoln Laboratory, sponsored by Defence Advanced Research Project Agency (DARPA) and Air Force Research Laboratory, has collected and evaluated the first standard corpora for evaluation of computer network Intrusion Detection Systems. This is called the DARPA Intrusion Detection Evaluation.

The goal for DARPA is to develop Intrusion Detection Systems, or aggregate of systems, that can detect more than 99% of the attacks with a false alarm rate less than 1% (Bhavin Shah and Bhushan H Trivedi, 2012). Over the last years, a large quantity of data has been gathered by the Lincoln Laboratory for the purpose of testing and comparing Intrusion Detection Systems. The data sets from Massachusetts Institute of Technology's Lincoln Laboratory are the most well-known and used data sets for Intrusion Detection System research.

These evaluations measure probability of detection and probability of false alarm for each system under test. These evaluations are contri-buting significantly to the intrusion detection re-search field by providing direction for research efforts and an objective calibration of the current technical state-of the-art. They are of interest to all researchers working on the general problem of workstation and network intrusion detection. The evaluation is designed to be simple, to focus on core technology issues, and to encourage the widest possible participation by eliminating se-curity and privacy concerns, and by providing data types that are used commonly by the majority of intrusion detection systems. The DARPA Intrusion Detection Evaluation is designed to find the stren-gth and weaknesses of existing approaches and lead to large performance improvements and valid assessment of Intrusion Detection Systems. The concept was to generate a set of realistic attacks, embed them into normal data, evaluate then false alarms and detection rates of systems with these data, and the improve systems to corr-ect weaknesses found. The data sets from DARPA are used by many researchers around the world to test new Intrusion Detection Systems, either they are anomaly based or misuse based systems.

To train and testing the ANN models, very rich training data set is required. All the authors had used either KDD CUP 1999 dataset (The Third International Knowledge Discovery and Data Mining Tools Competition) or DARPA dataset. These both the datasets are standardized and freely available on the internet. KDD Dataset has 42 columns. Last column of this dataset is related with given row is either attack or normal. KDD Dataset contains approximately 500000+ rows for training and 40000000+ rows for the testing.

THE PROPOSED METHOD

BPNN is supervised learning method in which set of the input and expected output must be provided. BPNN can have one or multiple hidden layers. Optimal number of the hidden units

for given number of inputs and outputs, can be decided by the trial and error method. For the problems like detecting and categorizing the attack, one hidden layer is more than sufficient. Even two or three hidden layer can be implemented but it will increase the complexity of the system and hence will reduce the convergence rate. Number of the units in each layer can also increase or decrease the complexity of the system. Too much hidden units can reduce the performance of the system, while too low hidden units can reduce the detection rate.

We implemented the BPNN by using the KDD CUP 1999 dataset Mieso (K Denko, 2010) of MIT Lincoln Laboratory. We took 10% (= 500000+ input rows) of the dataset as training and 90 %(= 40000000+ input rows) dataset for the testing. The dataset contains 24 types of training attacks, with an additional 14 types in the test data. All the attacks fall into four main categories: DOS, R2L, U2R and Probe attacks. We implemented two experiments. For both the experiments, we took 41 inputs, one output, learning rate as constant (0.9), and initial weights were set as random value. In our first experiment, we took two hidden layer with 41 hidden units in each layer. While in the second experiment, we took one hidden layer with 42 hidden units. Our experiments shows that model with one hidden layer takes less time for the training as compare to the two hidden layer. Even convergence rate is also high for the model with one hidden layer. During our both the experiments, we observed that BPNN is suffering from the local minima, and slow coverage. Performance is good in detection of the known and unknown attack. But, to train the BPNN, number of the epochs required was very high which lead to very high training time. If network is over trained then it can decrease the performance, and to overcome, one has to define the early stopping condition. As BPNN can support multiple output unit, it is possible to classify the given data record in to one of the attack category of KDD CUP. Usage of neural networks for intrusion detection with the input data from the DARPA project was presented in many papers. Unfortunately, since 1999 when KDD 99 was published many new attacks have been launched. Therefore for more credible tests of IDS it is important to update the KDD 99 data set. Selection of the input data is a very important issue. Representation of all types of attacks and normal activity should be included in the learning data set. Therefore besides new attack connections, representation of normal traffic not presented before in KDD99 was added.

RESULTS AND DISCUSSION

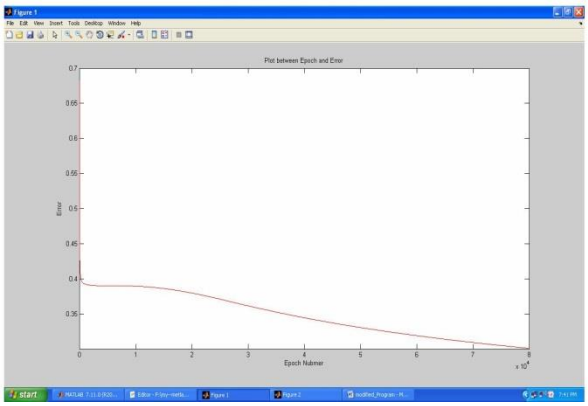
After executing the Program of Training of Neural Network specified in section IV values of y, W. W0, V, V0 are as under

```
Y = 0.7593   0.9990   0.0949   0.9694  0.9234
0.1133       1.0000   1.0000   0.1660  0.8952
0.8883       1.0000   0.6756   1.0000  0.0000
0.0000       0.9569   0.0504
EPOCH = 990000
W =
[ 69.7724
-92.1417
75.1616 ]
WO =
[ 5.7396]
V =
[ -101.2418 -48.9699 238.5258
180.6523 215.8780  -0.1458
-245.0327 -305.1403  9.9378]
VO = [ 0.9382  4.7101  -1.8095
```

The Graphical representation of Epoch and Error is as shown in Figure 7.1 which is plot bet-

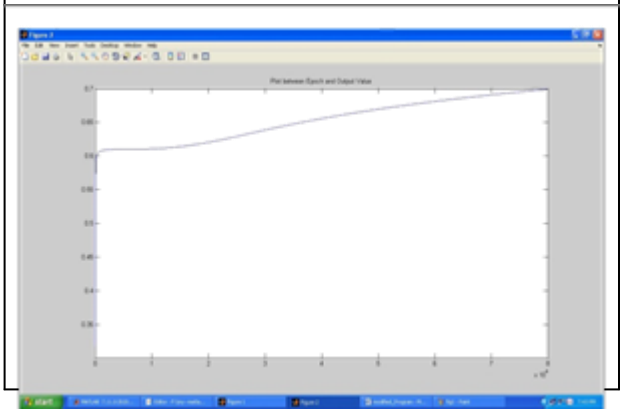
ween Epoch Number and Error. As we know, when iterations error between target value and output value decreases which is clear from Figure 1.

Figure 1: Plot between Epoch Number and Error



Now, as we know as the iterations increases output value reaches to target value it is clearly shown in Figure 7.2 which is a plot between epoch and output value.

Figure 2: Plot between Epoch Number and Output Value



LIMITATIONS

As for many studies; there are some different challenges viewed in the intrusion detection systems. In this study, some limitations were faced. They can be summarized as follows:

[12] Intrusion detection systems need a periodic update to the training set and profiles.

[13] Using a static training data might become outdated and deficient for prediction.

[14] The accuracy of classification for the data do not 100%.

CONCLUSION

There are various techniques of Artificial Neural Network, which can be applied to Intrusion Detection System. Each technique is suitable for some specific situation. BPNN is easy to implement, supervised learning artificial neural network. Number of the epochs required to train the network is high as compare to the other ANN techniques. But, detection rate is very high. BPNcan be used when one wants to not only detect the attack but also to classify the attack in to specific category so that preventive action can be taken. By combining the different ANN techniques, one can reduce the number of the epochs required and hence can reduce the training time. As the DOS attack will be resolved in this work, the throughput of the network will be improved and network delay will be reduced. Change in the number of nodes in hidden layer

resulted in the change in classification rate and also change in the false positive rate and false negative rate for the neural network based intrusion detection systems. The work does not require any additional hardware and is software based. In the future this system could be extended to an online system by little effort.

REFERENCES

1. Afrah Nazir (2013), "A Comparative Study of different Artficial Neural Networks Based Intrusion Detection Systems" *International Journal of Scientific and Research Publications*, Vol. 3, No. 7.
2. Amit Garg and Ravindra Pratap Singh (2014), "Voltage Profile Analysis in Power Transmission System Based on STATCOM Using Artificial Neural Network in MATLAB/ SIMULINK", *International Journal of Applied Information Systems (IJAIS), Foundation of Computer Science, New York, USA*, Vol. 6, No. 1.
3. Bhavin Shah and Bhushan H Trivedi (2012), "Artificial Neural Network Based Intrusion Detection System", *International Journal of Computer Applications*, Vol. 39, No. 6.
4. M Dondo and J Treurniet (2004), "Investigation of a Neural Network Implementation of a TCP Packet AnomalyDetection System", *Defence Research an d Development Canada*.
5. Manoranjan Pradhan, Sateesh Kumar Pradhan and Sudhir Kumar Sahu (2012), "Anomaly Detection Using Artificial Neural Network" *International Journal of Engineering Sciences & Emerging Technologies*.
6. Mehdi Moradi and Mohammad Zulkernine (2009), "A Neural Network Based System for Intrusion Detection and Classification of Attacks".
7. Mieso K Denko (2010), "Detection and Prevention of Denial of Service Attacks in Mobile Adhoc Networks Using Reputation-Based Incentive Scheme", *Systemics, Cybernetics and Informatics*, Vol. 3, No. 4.
8. Przemyslaw Kukielka and Zbigniew Kotulski (2012), "Adaptation of the neural network-based IDS to new attacks detection".
9. Przemyslaw Kukielka and Zbigniew Kotulski (2010), "Analysis of Neural Networks Usage for Detection of a New Attack in IDS", *Annales UMCS Informatica AI X*, Vol. 1, pp. 51-59.
10. S Devaraju and S Ramakrishnan (2012), "Detection of Accuracy for Intrusion Detection System Using Neural Network Classifier", *International Journal of Emerging Technology and Advanced Engineering (IJETAE)*.
11. Samaneh Rastegari, M Iqbal Saripan and Mohd Fadlee A Rasid (2009), "Detection of Denial of Service Attacks Against Domain Name System Using Neural Networks", *IJCSI International Journal of Computer Science Issues*, Vol. 6, No. 1.

PARKING:- PROBLEM & SOLUTION

Sonu Panchal[1], Vikram[2]

Department of Civil Engineering, SBMNEC, Asthal Bohar, Rohtak, Haryana, India Email: sonu.007panchal@yahoo.com

ABSTRACT

Parking management refers to various policies and programs that result in more efficient use of parking resources. This guide describes and evaluates more than two-dozen such strategies. It investigates problems with current parking planning practices, discusses the costs of parking facilities and the savings that can result from improved management, describes specific parking management strategies and how they can be implemented, discusses parking management planning and evaluation, and describes how to develop the optimal parking management program in a particular situation. Cost-effective parking management programs can usually reduce parking requirements by 20-40% compared with conventional planning requirements, providing many economic, social and environmental benefits.

INTRODUCTION

Parking is the act of stopping and disengaging a vehicle and leaving it unoccupied. Parking on one or both sides of a road is often permitted, though sometimes with restrictions. Parking facilities are constructed in combination with some buildings to facilitate the coming and going of the buildings' users .Every country has it own rule to design parking space

A typical automobile is parked 23 hours each day, and uses several parking spaces each week. Parking facilities are an essential component of a transportation system. They are also costly. Parking conflicts are among the most common problems facing designers, operators, planners and other officials. Such problems can be often defined either in terms of supply (too few spaces are available, somebody must build more) or in terms of management (available facilities are used inefficiently and should be better managed). Management solutions tend to be better than expanding supply because they support more strategic planning objectives:

[15] Reduced development costs and increased affordability.

[16] More compact, multi-modal community planning (smart growth).

[17] Encourage use of alternative modes and reduce motor vehicle use (thereby reducing traffic congestion, accidents and pollution).

[18] Improved user options and quality of service, particularly for non-drivers.

[19] Improved design flexibility, creating more functional and attractive communities.

[20] Ability to accommodate new uses and respond to new demands.

[21] Reduced impervious surface and related environmental and aesthetic benefits.

TYPES OF PARKING

Table 1 describes various types of parking facilities and the role they play in an efficient parking system. These categories overlap: surface parking lot can be unpriced, priced but serve just one destination, or commercial. Parking facilities that are regulated and priced to favor higher value trips (such as deliveries and customers over commuters and residents), and serve multiple destinations tend to be used most efficiently. Parking facilities that are priced and serve multiple destinations tend to be most efficiently used. Figure 1 shows the various types of parking facilities in the cities

Type	Images	Costs and Density	Role
On-Street (or Curb) Designated parking spaces located within a road right-of-way, usually in the curb lane.		Moderate construction costs and high density (relatively little land used per space) because they require no driveway.	Convenient to use, and can serve multiple destinations. On-street parking should be managed for maximum efficiency.
Surface Parking A parking lot directly on the ground (either paved or unpaved).		Low to moderate construction costs. Low density (they require lots of land per space, including driveways and circulation lanes.	Inefficient if they serve a single destination. Should be minimized and managed for efficiency.
Structured or Underground Any multi-story parking structure (often called a *parking garage*, *parkade* or *ramp*), including parking facilities within or under a building.		High construction costs but relatively low land costs and high densities.	Supports compact development but must be efficiently managed to justify their high construction costs.
Priced (or Metered) Any parking facility where motorists are charged directly for use, including on-street metered parking, and off-street lots where motorists pay by the hour, day, week, month or year.		Varies. Can be applied to any type of parking structure.	Pricing, particularly congestion pricing (fees are higher at times and places with high demand) tends to encourage efficient use of parking facilities.
Commercial Parking A for-profit parking lot available to any motorist and serves multiple destinations.		Varies. Can be applied to any type of parking structure.	Tends to be efficient because it is priced and usually serves multiple destinations.

Figure 1: Types of Parking Facilities

Table 2 Old and New Parking Paradigms Compared

Old Parking Paradigm	New Parking Paradigm
Parking problem means inadequate parking supply.	There can be many types of parking problems, including inadequate or excessive supply, too low or high prices, inadequate user information, and inefficient management.
Transportation means driving.	Travelers may use various modes. Not everybody drives.
Abundant parking supply is always desirable.	Too much supply is as harmful as too little.
All parking demand should be satisfied on-site. Motorists should not be forced to walk to their cars.	Parking can often be provided off-site, allowing sharing of parking facilities among various destinations.
Parking should generally be provided free, funded indirectly, through rents and taxes.	As much as possible, users should pay directly for parking facilities.
Parking should be available on a first-come basis.	Parking should be regulated to favor higher priority uses and encourage efficiency.
Parking requirements should be applied rigidly, without exception or variation.	Parking requirements should reflect each particular situation, and should be applied flexibly.
Innovation faces a high burden of proof and should only be applied if proven and widely accepted.	Innovations should be encouraged, since even unsuccessful experiments often provide useful information.
Parking management is a last resort, to be applied only if increasing supply is infeasible.	Parking management programs should be widely applied to prevent parking problems.
Land use dispersion (sprawl) is acceptable or even desirable.	Dispersed, automobile-dependent development can be harmful.

1. OFF-STREET PARKING: Off street parking projects have been implemented on Housing complexes, office buildings, shopping malls, cinemas, airports, railway stations, auditoriums and fair grounds. Off-street parking provides the following benefits:

[20] Quick and effortless parking for pick-up and drop-off
[21] Pay only for the parking duration
[22] Safety and security for vehicles and passengers
[23] More parking bays for genuine parkers

Some of the off street parking are:

a) Multi Level Parking
b) **Basements Parking**

2. ON-STREET PARKING: On-street parking is a key factor in promoting businesses in cities, particularly within central business districts. The effective management of on-street parking space with benefits for both parking operators and parking space users. Generally, the layout of a parking lot seeks to strike a balance among maximizing capacity, maneuverability, and circulation. . . .

The general advantages of 90° parking, as compared with lesser angles, are:

1. Most common and understandable;
2. Can sometimes be better fitted into buildings;
3. Generally most efficient if site is sufficiently large;
4. Uses two-way movement (can allow short, dead-end aisles);
5. Allows unparking in either direction. Thus it can minimize travel distances and internal conflict;
6. Does not require any aisle directional signs or markings;
7. Wide aisles often provide room to pass vehicles stopped and waiting for an unparking vehicle;
8. Wide aisles increase separation for pedestrians walking in the aisle and between moving vehicles;

9. Wide aisles increase clearance from other traffic in the aisle, during unparking maneuvers;

10. Fewer total aisles (hence easier to locate parked vehicle).

Several advantages and disadvantages of angle parking (usually 45° to 75°), are:

1. Easiest in which to park
2. Can be adapted to almost any width of site by varying the angle;
3. Requires slightly deeper stalls but much narrower aisles and modules;
4. Drivers must unpark and proceed in original direction; hence producing greater out-of-way travel and conflict;
5. Unused triangles at end of parking aisles reduce overall efficiency;
6. To avoid long travel, additional cross aisles for one-way travel are required, which adds to gross area used per car parked;
7. Difficult to sign one-way aisles.

On-street parking is a highly efficient form of parking because the typical on-street parking space requires only about half the land that the typical space in a surface parking lot requires. On-street parking provides the following additional benefits to redeveloping areas:

- Improved street-level retail activity – provides a cheap and convenient supply of parking for street-oriented businesses.
- Allows pricing of parking supply – can easily be metered and meters can be adjusted over time for greater efficiency.


Disadvantages of on-street parking include:

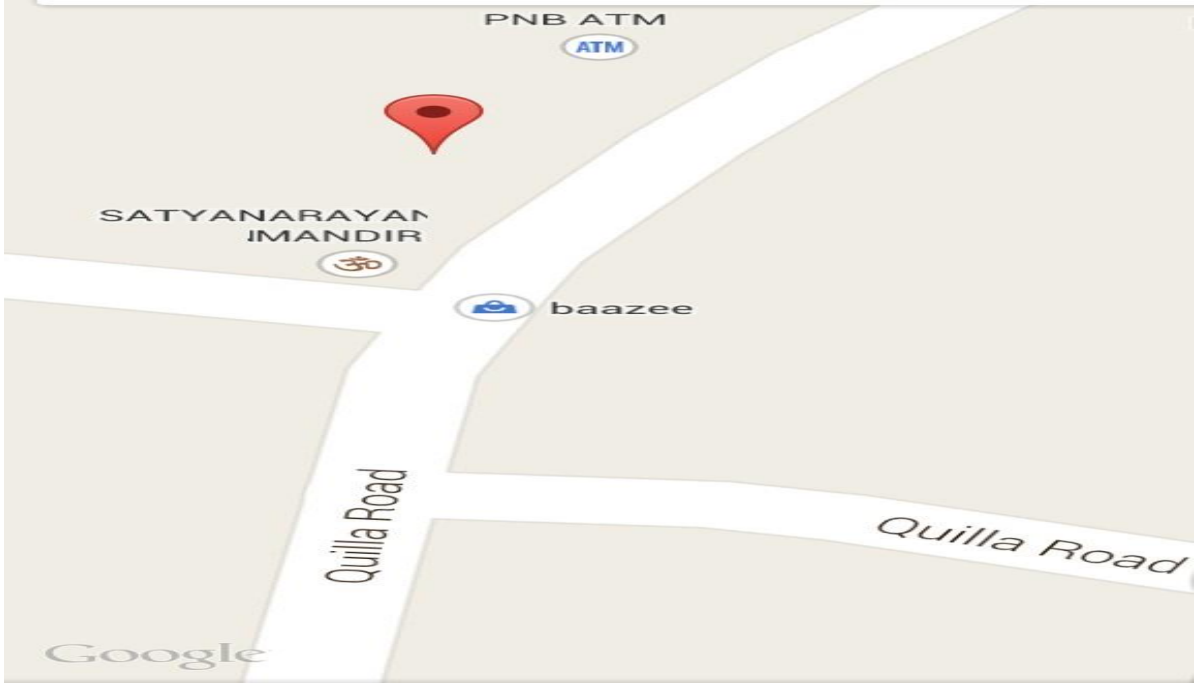
- Takes place in street rights-of-way.
- Requires public enforcement resources to deter crime and unauthorized use.
- Needs to be coordinated with the locations of transit bus stops and bicycle lanes.
- Vehicular/ Bicycle/ Pedestrian conflict with passing traffic as pedestrians exit their cars.

Parking Problems in Quilla Road of Rohtak

Due lack of parking space on the roads the people face congestion around the Quilla road market area of Rohtak. The parking situation is as shown below:-

PROPOSED SOLUTION TO PARKING PROBLEM

It is proposed to come up with a Multi Level Parking lot around the study area to resolve the parking problem.



REFERENCES

[1] Er. Sanjay Poswal, Er. Nikhil Jain, Er. Neeraj Saxena (2014), "Solutions to Parking Problem in Urban Cities of India", National conference on advanced research in science engineering and technology, jan.-feb 2014.
[2] Highway Engineering by khanna & justo
[3] http://en.wikipedia.org/wiki/Parking
[4] https://www.google.co.in/maps/place/Quilla+Rd

SUSPENSION SYSTEM

Yati Parjesh[2]

Associate Prof., S.B.M.N.E.C., Asthal Bohar, Rohtak
Asst. Prof., S.B.M.N.E.C., Asthal Bohar, Rohtak Email: Yati3k@gmail.com

INTRODUCTION:

Suspension is the system of tires, tire air, springs, shock absorbers and linkages that connects a vehicle to its wheels and allows relative motion between the two. Suspension systems serve a dual purpose — contributing to the vehicle's road holding/handling and braking for good active safety and driving pleasure, and keeping vehicle occupants comfortable and a ride quality reasonably well isolated from road noise, bumps, and vibrations, etc.

These goals are generally at odds, so the tuning of suspensions involves finding the right compromise. It is important for the suspension to keep the road wheel in contact with the road surface as much as possible, because all the road or ground forces acting on the vehicle do so through the contact patches of the tires. The suspension also protects the vehicle itself and any cargo or luggage from damage and wear. The design of front and rear suspension of a car may be different.

SPRING RATE

The spring rate (or suspension rate) is a component in setting the vehicle's ride height or its location in the suspension stroke. When a spring is compressed or stretched, the force it exerts is proportional to its change in length. The *spring rate* or *spring constant* of a spring is the change in the force it exerts, divided by the change in deflection of the spring. Vehicles which carry heavy loads will often have heavier springs to compensate for the additional weight that would otherwise collapse a vehicle to the bottom of its travel (stroke). Heavier springs are also used in performance applications where the loading conditions experienced are more extreme.

Springs that are too hard or too soft cause the suspension to become ineffective because they fail to properly isolate the vehicle from the road. Vehicles that commonly experience suspension loads heavier than normal have heavy or hard

MATHEMATICS OF THE SPRING RATE

Spring rate is a ratio used to measure how resistant a spring is to being compressed or expanded during the spring's deflection. The magnitude of the spring force increases as deflection increases according to Hooke's Law. Briefly, this can be stated as

$$F = -kx$$

where

F is the force the spring exerts

k is the spring rate of the spring.

x is the deflection of the spring from its equilibrium position (i.e., when no force is applied on the spring)

Spring rate is confined to a narrow interval by the weight of the vehicle, load the vehicle will carry, and to a lesser extent by suspension geometry and performance desires.

Spring rates typically have units of N/mm (or lbf/in). An example of a linear spring rate is 500 lbf/in. For every inch the spring is compressed, it exerts 500 lbf. A non-linear spring rate is one for which the relation between the spring's compression and the force exerted cannot be fitted adequately to a linear model. For example, the first inch exerts 500 lbf force, the second inch exerts an additional 550 lbf (for a total of 1050 lbf), the third inch exerts another 600 lbf (for a total of 1650 lbf). In contrast a 500 lbf/in linear spring compressed to 3 inches will only exert 1500 lbf.

The spring rate of a coil spring may be calculated by a simple algebraic equation or it may be measured in a spring testing machine. The spring constant *k* can be calculated as follows:

$$k = \frac{d^4 G}{8 N D^3}$$

where *d* is the wire diameter, G is the spring's shear modulus (e.g., about 12,000,000 lbf/in² or 80 GPa for steel), *N* is the number of wraps and *D* is the diameter of the coil.

WHEEL RATE

Wheel rate is the effective spring rate when measured at the wheel. This is as opposed to simply measuring the spring rate alone.

Wheel rate is usually equal to or considerably less than the spring rate. Commonly, springs are mounted on control arms, swing arms or some other pivoting suspension member. Consider the example above where the spring rate was calculated to be 500 lbs/inch (87.5 N/mm), if you were to move the wheel 1 in (2.5 cm) (without moving the car), the spring more than likely compresses a smaller amount. Let's assume the spring moved 0.75 in (19 mm), the lever arm ratio would be 0.75:1. The wheel rate is calculated by taking the square of the ratio (0.5625) times the spring rate, thus obtaining 281.25 lbs/inch (49.25 N/mm). Squaring the ratio is because the ratio has two effects on the wheel rate. The ratio applies to both the force and distance traveled.

ROLL RATE

Roll rate is analogous to a vehicle's ride rate, but for actions that include lateral accelerations, causing a vehicle's sprung mass to roll about its roll axis. It is expressed as torque per degree of roll of the vehicle sprung mass. It is influenced by factors including but not limited to vehicle sprung mass, track width, CG height, spring and damper rates, roll center heights of front and rear, anti-roll bar stiffness and tire pressure/construction. The roll rate of a vehicle can, and usually does, differ front to rear, which allows for the tuning ability of a vehicle for transient and steady state handling. The roll rate of a vehicle does not change the total amount of weight transfer on the vehicle, but shifts the speed at which and percentage of weight transferred on a particular axle to another axle through the vehicle chassis. Generally, the higher the roll rate on an axle of a vehicle, the faster and higher percentage the weight transfer on that axle.

DAMPING

Damping is the control of motion or oscillation, as seen with the use of hydraulic gates and valves in a vehicle's shock absorber. This may also vary, intentionally or unintentionally. Like spring rate, the optimal damping for comfort may be less than for control.

Damping controls the travel speed and resistance of the vehicle's suspension. An undamped car will oscillate up and down. With proper damping levels, the car will settle back to a normal state in a minimal amount of time. Most damping in modern vehicles can be controlled by increasing or decreasing the resistance to fluid flow in the shock absorber.

SEMI-ACTIVE AND ACTIVE SUSPENSIONS

If the suspension is externally controlled then it is a semi-active or active suspension — the suspension is reacting to what are in effect "brain" signals. As electronics have become more sophisticated, the opportunities in this area have expanded.

For example, a hydropneumatic Citroën will "know" how far off the ground the car is supposed to be and constantly reset to achieve that level, regardless of load. It will *not* instantly compensate for body roll due to cornering however. Citroën's system adds about 1% to the cost of the car versus passive steel springs.

Semi-active suspensions include devices such as air springs and switchable shock absorbers, various self-levelling solutions, as well as systems like hydropneumatic, hydrolastic, and hydragas suspensions. Mitsubishi developed the world’s first production semi-active electronically controlled suspension system in passenger cars; the system was first incorporated in the 1987 Galant model. Delphi currently sells shock absorbers filled with a magneto-rheological fluid, whose viscosity can be changed electromagnetically, thereby giving variable control without switching valves, which is faster

and thus more effective.

Fully active suspension systems use electronic monitoring of vehicle conditions, coupled with the means to impact vehicle suspension and behavior in real time to directly control the motion of the car. Lotus Cars developed several prototypes, from 1982 onwards, and introduced them to F1, where they have been fairly effective, but have now been banned. Nissan introduced a low bandwidth active suspension in circa 1990 as an option that added an extra 20% to the price of luxury models. Citroën has also developed several active suspension models (see hydractive). A recently publicised fully active system from Bose Corporation uses linear electric motors (i.e., solenoids) in place of hydraulic or pneumatic actuators that have generally been used up until recently. Mercedes introduced an active suspension system called Active Body Control in its top-of-the-line Mercedes-Benz CL-Class in 1999.

4 G WHAT'S NEXT ?

Arvind Batra

Asstt. Prof., Baba Mastnath Engg. College, Asthal Bohar, Rohtak-124001 E-Mail : Batra_Arvind@rediffmail.com

The convergence of economic and technological factors is driving the proliferation of a wide range of wireless devices from sophisticated smartphones to broadband-enabled laptops, which is whetting peoples' appetites for broadband services outside of the typical home or office environment. Increasingly, people want to be connected wherever they are, at any given time. Additionally, with recent, dramatic reductions in the cost of devices for basic connectivity – see the success of USB dongles – the world is poised for massive adoption of mobile broadband. Just think about the 6.5 billion applications that have been downloaded on just one operating system from Apple's app store and consider that the Android operating system (launched recently by Google) is witnessing similar growth. At the same time, wireless networks have come a long way since their introduction to consumers two decades ago. Two key factors essentially contributed to this evolution. Firstly, the move from analog to digital in major parts of a radio network (such as radio access and core network) that has dramatically improved cellular network performance and capabilities, and secondly, the continuous optimization of the usage of precious radio resources, enabling more and more data to be transmitted over the air. Consumers are excited by faster speeds and the ability to remain broadband connected when on-the-go. Findings from recent European research indicate Live Messaging, Next Generation Music and Enhanced Mobile Video are the top three applications consumers would love to get permanent access to. Businesses also see the value of wireless multimedia for increased efficiency in business processes. Both market segments have different drivers for wireless broadband adoption. However, they both expect to increase their use of mobile online data activities as they see the benefit of new, high-value applications such as Mobile Collaboration or in-car connected services. An extensive study conducted in several developed countries was reviewed by the Bell Labs Business Modeling team. It confirmed that consumers and businesses are ready for 4G LTE applications and the interest in new LTE-optimized applications could be translated into a significant market opportunity for service providers. Participants ranged in age from 14 to 65 and used both mobile and broadband services. Enterprise respondents included decision makers in small, medium, and large companies responsible for the purchase, implementation, or deployment of employee mobile devices and services. In fact, consumers and enterprises report a notable willingness to switch operators in order to obtain such new services. In the European countries surveyed 51 percent of consumers and almost 80 percent of medium to large enterprises said they would sign up for 4G LTE service to enjoy the benefits of this technology innovation. Wireless network goes IP One technological change inherited from the wireline/Ethernet world is at the heart of current cellular networks evolution: Internet Protocol (IP). Today's 2G and 3G networks – to a large extent – are primed to transition and implement IP technology. But LTE (Long Term Evolution), as the 4th Generation of the evolution of cellular networks, is natively built on IP architecture and extends network transformation and capabilities substantially further than the previous generation of networks. 4G LTE is sure to bring dramatic improvements to the subscribers' mobile broadband experience thanks to higher data speeds and lower latency. 4G LTE is three times more spectrally efficient than 3G. 4G LTE boasts technological innovations such as Orthogonal Frequency Division Multiplexing (OFDM) and Multiple Input, Multiple Output (MIMO). When coupled with an all-IP end-to-end connection between devices and the network application servers (or between the devices), with each and every equipment using IP, these innovations significantly boost network efficiency. Besides providing the user with increased speeds more than 10 times faster than 3G, one of the key attributes redefining the LTE user experience is the lower latency. With three to five times better responsiveness when compared to 3G, bringing the wireless experience closer to the wired broadband desktop experience enables real-time applications like voice, video, and gaming to be delivered to the mass market over a pure wireless IP network. As a result, LTE promises to support the same kind of rich, broadband multimedia services (such as high-definition video, music, multiplayer video gaming) that people have come to expect at home, delivered wherever they are, any time of the day. In addition, LTE's

massive bandwidth and real-time capabilities offer the prospect of delivering a wide array of new services that haven't yet been defined. Think about for example, the introduction of fixed broadband and how it spurred the growth of the web and the myriad services that have become available over the past decade. New business model opportunities for service providers Changes in users' behavior and technology innovations open tremendous opportunities and perspectives for service providers. Opportunities to better serve their customers and increase their competitiveness, while at the same time exploring new business models and getting new revenue streams from untapped user demand. Examples of new business models enabled by 4G LTE broadband wireless are multiple. Early LTE adopters will be younger, more affluent and more open to advertising-subsidy models. This is a new and immediate source of revenue for operators, at almost no cost. Lower subscription prices would attract more users and in turn generate more data traffic, while advertising companies would find a new media for their targeted campaigns. Another example illustrating the huge potential of 4G is the ngConnect Program's "LTE Connected Car", a solution concept that drives an innovative ecosystem of devices, applications and content. The concept has a strong appeal among the under-35 age group, with Internet connectivity perceived as highly desirable. Along with always-on connection to the Internet, it offers an array of entertainment, traffic, navigation, car maintenance and safety features. Machine-to-Machine (M2M) is another key segment that is growing rapidly. Although it does not require a lot of bandwidth, it can strain the network because of the scale of the connections which run into millions in a small coverage area. LTE will support this demand more efficiently. Recent research from Technology Business Research Inc. showed that the volume of M2M connections will grow to more than 55 billion devices in this decade. Seizing such business opportunities also means addressing challenges for operators. To play the game, service providers require several assets. First, they need a high bandwidth, scalable and cost effective intelligent network with a high quality of experience (QoE). LTE's advanced capabilities are set to deliver on that even if implementations differ from one vendor to another. Second, they need 'application enablers' to monetize the broadband data explosion. Only by combining the trusted capabilities of their networks with the speed and innovation of the web, will operators give end-users and enterprises what they demand: a richer and more trusted web experience. In this area, relying on an industry partner with experience in both wireless and IP multi-media solutions is vital. Finally, and this is a recurrent concern for operators, they need to mitigate risk of business and network transformations. Few mobile service providers are familiar with wireline networks and the IP transformation they experienced in the past decade. They need the help of a trusted partner. Business transformation for enterprises For many businesses over much of the world including verticals, broadband services are becoming a necessity – an indispensable feature of our daily lives that extends well beyond the simple connectivity that today's networks have provided. In light of this ever-increasing demand for data, the need for a mobile solution to bridge the connectivity gap between home and office has become an important market driver. A recent study by Alcatel-Lucent in North America shows that messaging comes as the first wireless activity that will increase dramatically (87 percent of respondents), followed by download/view of documents (67 percent) and access to Internet/Intranet (60 percent) and video activities (59 percent). Among eight typical applications spanning Group Video Calling, Location-based Project Updates, Mobile Broadband File Sharing, Video Calling, Navigation Helper, Presence Contact List, Mobile Collaboration Tools and Priority of Critical Applications, respondents said they would use them daily with mobile broadband file sharing coming first on the list. Enterprises with over US$1 million in revenue and more than ten full time employees are the most likely to endorse such 4G applications. As a proof-point of 4G attractiveness, the majority of enterprises (74 percent) responded they would definitely or probably sign up for 4G service. 4G LTE: A huge potential and a revolution to our lifestyle In a fast-changing world with ever-increasing competitiveness, 4G LTE is set to unlock new business opportunities for both service providers and enterprises alike. It's up to them to get ready to leverage the huge potential this technology offers in terms of business efficiency and revenue streams. 4G LTE promises to redefine the user experience and the way we communicate. It will revolutionize our lifestyle.