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Design and Development of N type Coaxial to Waveguide (WR284) to Transition

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Abstract

This paper deals with design and development of coaxial to rectangular waveguide (WR284) transition at frequency 2998MHz. The CST microwave studio is used for electromagnetic simulation of transition. The S parameter analysis is carried out using transient solver of CST microwave studio. For broadband applications, a tuning probe is incorporated in the wider side wall of the waveguide. The simulated values of VSWR in 182 MHz and 500MHz bandwidth ranges are found to be 1.105 and 1.25 respectively. Based on design, the transition is fabricated and experiment was carried out to measure return loss. The simulated and measured return loss values at 2998 MHz are found to be 51dB and 33 db respectively.

Keywords: Coaxial to waveguide transition, electromagnetic simulation, S parameter, broad band

Introduction

Coaxial to waveguide transition is an important passive device required for low power measurement of various microwave sources and components [1]. The vector network analyzer (VNA) provides its output through coaxial line, so a coaxial to rectangular waveguide transition is required to test a waveguide type active and passive device. It is necessary that the transition should have minimum VSWR value. The various types of transitions are available in literature [1-4] but to the best of author's knowledge, this type of wide band transition is hardly available in the literature.

This paper deals with coaxial to rectangular waveguide transition. This transition is designed for low power testing of 3.1 MW magnetron and RF match load. Figs.1 and 2 show the schematic diagram of low power transition and magnetron cold test setup, respectively.

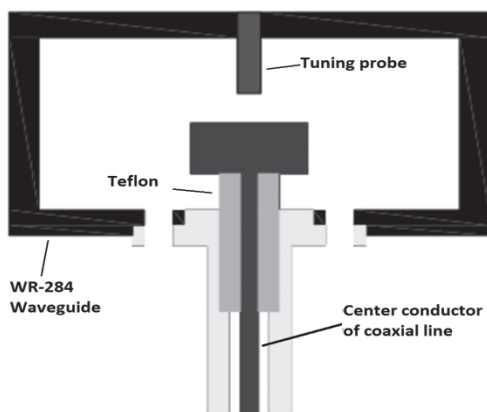


Fig.1: Schematic diagrams of coaxial to waveguide transition
(b)

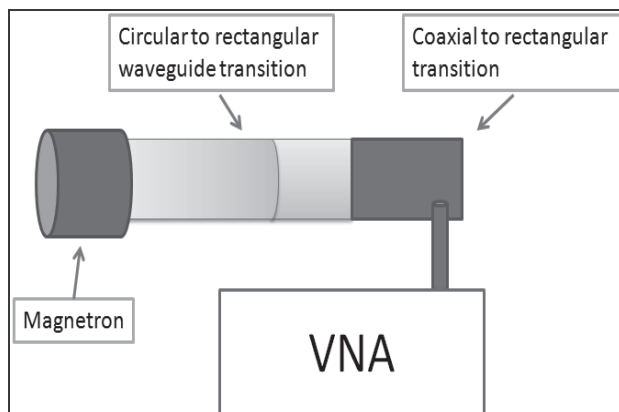


Fig.2: Schematic diagrams of magnetron cold testing setup

This paper is organized into four sections including the introduction section I. Section II briefly describes the electromagnetic simulation of coaxial to waveguide transition. Section III presents the fabrication and testing of coaxial to waveguide transition and finally section IV concludes the study.

Electromagnetic Simulation

In the coaxial to rectangular waveguide transition, the TEM modes convert into TE_{01} mode. The impedance of coaxial line and waveguide are different so at the junction point a matching network is required. Fig. 3 shows the simulation model of coaxial to waveguide transition.

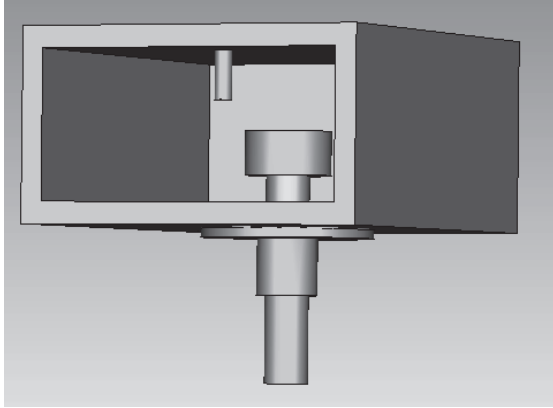


Fig.3: 3D simulation model of coaxial to waveguide transition.

Both coaxial and wave guide transmission lines have different values of impedances. To achieve impedance matching, the center conductor of the coaxial line extends into the waveguide up to half height of the waveguide and end of the center conductor is terminated with higher radius cylinder as shown in Fig. 3. The one end of WR 284 waveguide is shorted by end plate at a proper distance from center conductor of coaxial line. A tuning probe is inserted in the waveguide to achieve the broadband. The tuning probe is also used for frequency tuning. The dimensional parameters of transition are optimized by using transient solver of CST microwave studio. The optimized dimensional parameters of transition are given in table-I.

Table I: Dimensional parameters of transition

Parameter	Symbol	Unit	Value
Waveguide width and height	a, b	mm	72.136, 34.04
Coaxial line outer radius	D	mm	3.5
Coaxial line inner radius	d	mm	1.5
Height of center conductor	h	mm	17.2
End plate distance from center conductor	L	mm	34

The simulation results of S_{11} and S_{21} are shown Fig.4. The minimum value of S_{11} is found to be -51.24 dB at 3.0 GHz resonance frequency. The S_{21} value on frequency 3.0 GHz is found to be -0.00964 dB.

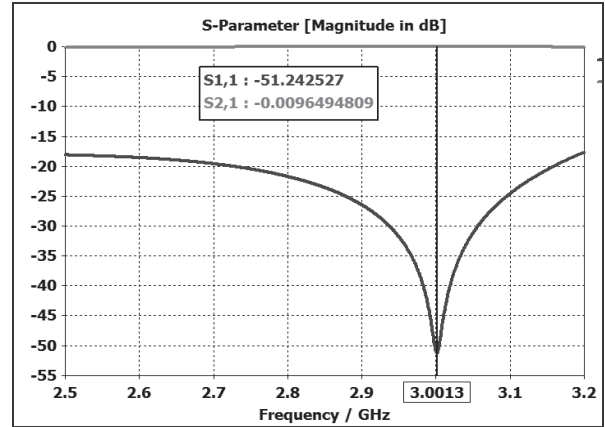


Fig.4: Frequency versus S_{11} and S_{21} variation for coaxial to waveguide transition.

The variation VSWR with respect to frequency is shown in Fig.5. It is obvious from this figure that the minimum value of VSWR is 1.005. The bandwidth at 1.1 and 1.25 are found to be 182 MHz and 500 MHz respectively. The effect of center conductor height (h) and end plate distance (l) on S_{11} are shown in Fig. 6 and Fig. 7. It is found from Fig. 6 that the endplates affects the resonance frequency. It is found from Fig. 7 that due to change in height of central conductor, the value of S_{11} is affected. These studies are used in optimization of transition.

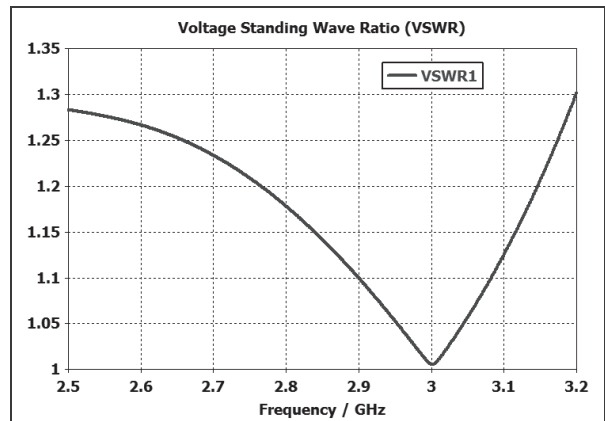


Fig.5: Voltage standing wave ratio (VSWR) versus frequency plots

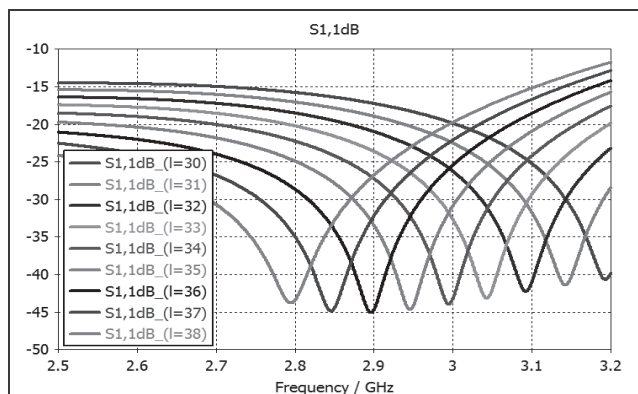


Fig.6: Frequency versus S_{11} for different endplate distances at fixed height of center conductor

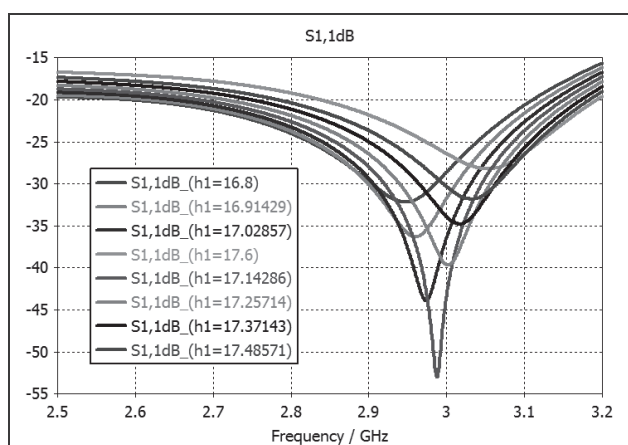


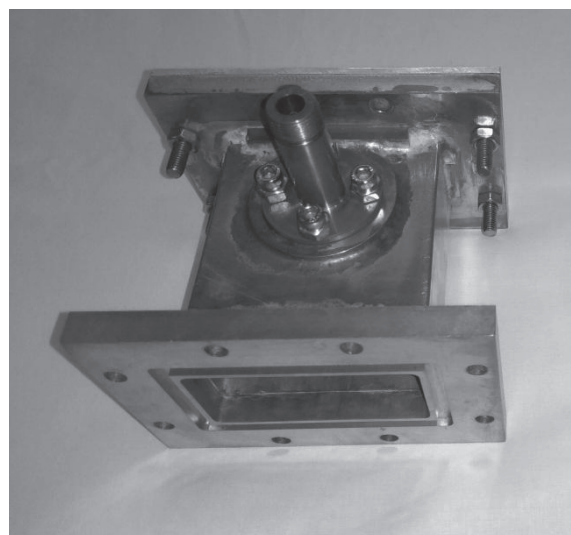
Fig.7: Frequency versus S_{11} for different height of center conductor at fixed endplate distances.

Fabrication and Testing of Coaxial to Wr- 284 Transition

On the basis of design, the coaxial to rectangular waveguide transition is fabricated. For fabrication, a WR-284 waveguide section of length 96 mm is taken. Rectangular flanges with 8 holes are brazed on both sides of the waveguide. The one side of the brazed section is shorted. The other side is open which is used for testing purpose. N type coaxial connector, central conductor, and teflon cylinder are fabricated which are shown in Fig. 8. This assembly of these parts is fitted in the wider side of the waveguide.



(a)



(b)

Fig. 8: (a) Fabricated parts of transition, (b) final fabricated coaxial to WR284 transition

For measurement of return loss, the fabricated transition is attached with a RF load which is shown in Fig. 9. The return loss is measured a VNA. The measured return loss is shown in Fig. 10. It is obvious from this that the measured value of return loss is 33.78dB at frequency 2.998 GHz.

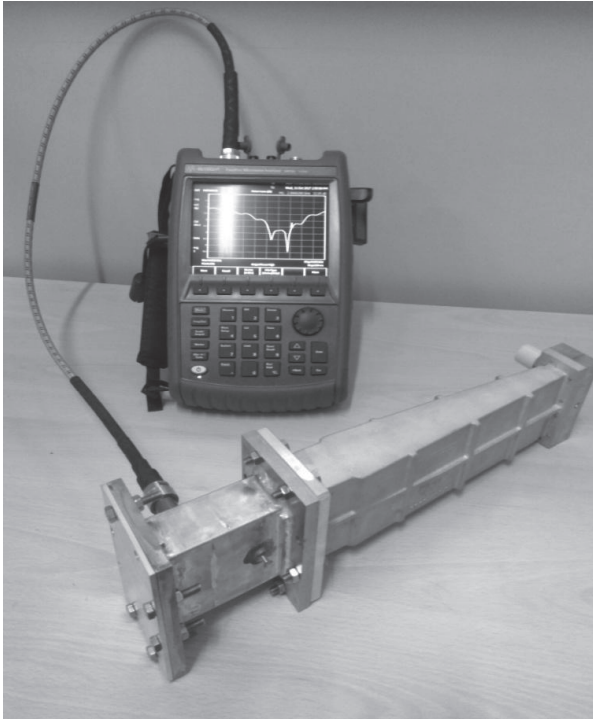


Fig.9: Experimental set-up of coaxial to WR 284 transition

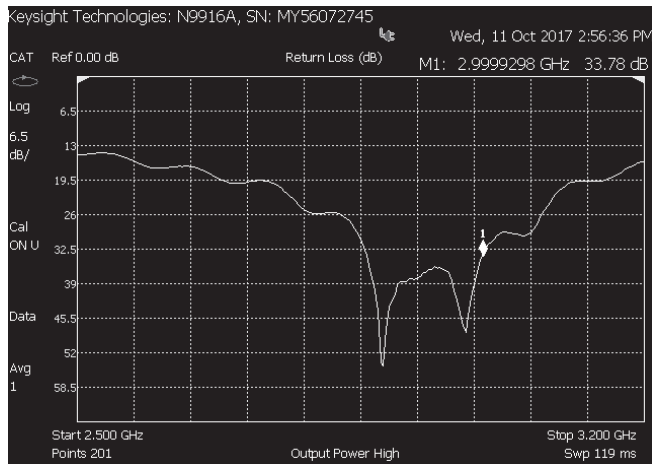


Fig.10: Experimental result

Conclusion

The coaxial to rectangular waveguide (WR 284) transition at frequency 2998MHz is designed and developed. The CST microwave studio is used for electromagnetic simulation of transition. On the basis of design, parts of transition are fabricated and finally the transition is fabricated. The simulated and measured values of return loss are found to be 51dB and 33 dB respectively at spot frequency 2998 MHz

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Monitoring of Toxicity in Cosmetic Powders using Laser Induced Breakdown Spectroscopy

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Abstract

Laser Induced Breakdown Spectroscopy (LIBS) is a well known atomic spectroscopic technique for instant and simultaneous elemental analysis of variety of materials in any phase (solid, liquid, gas or aerosol) with no or minimal sample preparation. The study deals with LIBS analysis of cosmetic powders of different brands (A, B and C). The LIBS spectra of these samples have been recorded by focusing the high power pulsed laser on the surface of sample (pellets of three brands of these sample) in the spectral range of 200-500 nm. LIBS spectra show the presence of spectral signatures of Ca, Mg, Si, Fe, Zn and Al. Out of them Al is the toxic element which is harmful for human skin. Thus, our result emphasizes the capability of LIBS for monitoring of toxicity in cosmetic powders.

Keywords: LIBS, Cosmetic Powders, toxicity

Introduction

Cosmetic products (like powders) are used by peoples due to many different cultures for various reasons such as traditional, artistic, or aesthetic purposes. Some compounds that have been used for cosmetic purposes, vary from malachite, green copper ore, galena, khol, mercuric sulfide and white lead in ancient times to phthalates, iron oxides, parabens, alkylphenols, synthetic musks, and many others in today's modern society. Most cosmetics are poorly tested, especially for chronic application causing low-grade toxicity. Most have been scarcely tested at all, and only a few out of thousands have had expensive toxicity testing. Consumers who suffer from reactions of these products rarely complain and they just stop using the product. Make-up material may contain aluminum, a well known toxin in humans to make them long-lasting. In 1982 Daniel Cramer, MD, reported in the journal Cancer that women in Boston who used talcum powder on their genitals and sanitary napkins had a 328 times greater risk of ovarian cancer.

Laser induced breakdown spectroscopy (LIBS) has been one of the most popular techniques since last decades in different branches of science due to its unique features [1]. It is an atomic emission spectroscopic technique which is growing steadily and showing a strong potential for performing direct

spectro-chemical elemental analysis of versatile materials, solids including non-conducting solids, liquids, gases, aerosol and molten state with minimal sample pretreatment procedures [2]. High energetic pulse laser beam is focused on the sample surface which is absorbed through various energy transfer mechanism including thermal and non-thermal heating, melting and finally ablation of the sample and form plasma plume on its surface. Here, focused laser beam initiates an avalanche ionization of the sample elements, giving rise to the breakdown effect and the subsequent plasma formation. On cooling of plasma generated by the intense laser pulse-sample interaction, characteristic photons are generated which is collected and fed to spectrometer for recording LIBS spectra [3-5]. The spectral analysis of the atomic and ionic emission lines of the sample species are performed with the help of NIST data base [6].

Experiments

The LIBS experimental set-up shown in Figure 1, consists of a laser source, a sample stage, and a spectrometer [7-8]. The frequency-doubled Nd:YAG laser (532 nm) source capable of delivering a maximum energy of 425 mJ, pulse duration of 4 ns (FWHM) and maximum pulse repetition rate of 10 Hz was used to record the LIBS spectra [7-9].

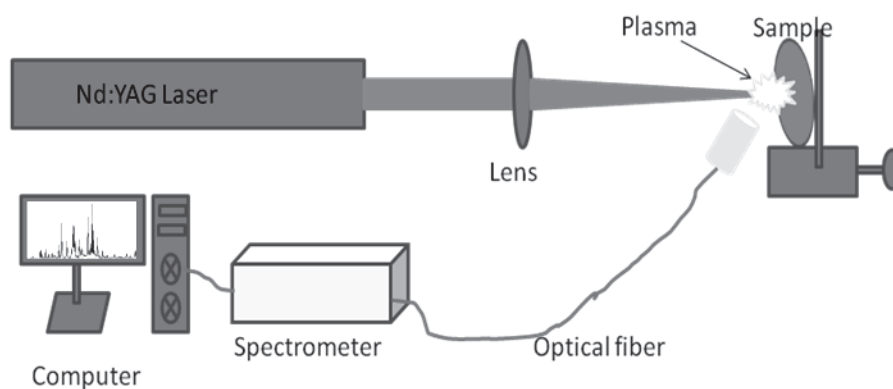


Figure 1: Typical experimental setup for LIBS analysis

15 cm focal length lens was used to focus the laser beam on the surface of the cosmetic powder pellet. The characteristic photons emitted from the plasma were collected using optical

fiber bundle at an angle of 45° to the incident laser beam and fed into the spectrometer (Ocean Optics, LIBS 2000+) equipped with a charge-coupled device (CCD) [7-9].

Result and Discussion

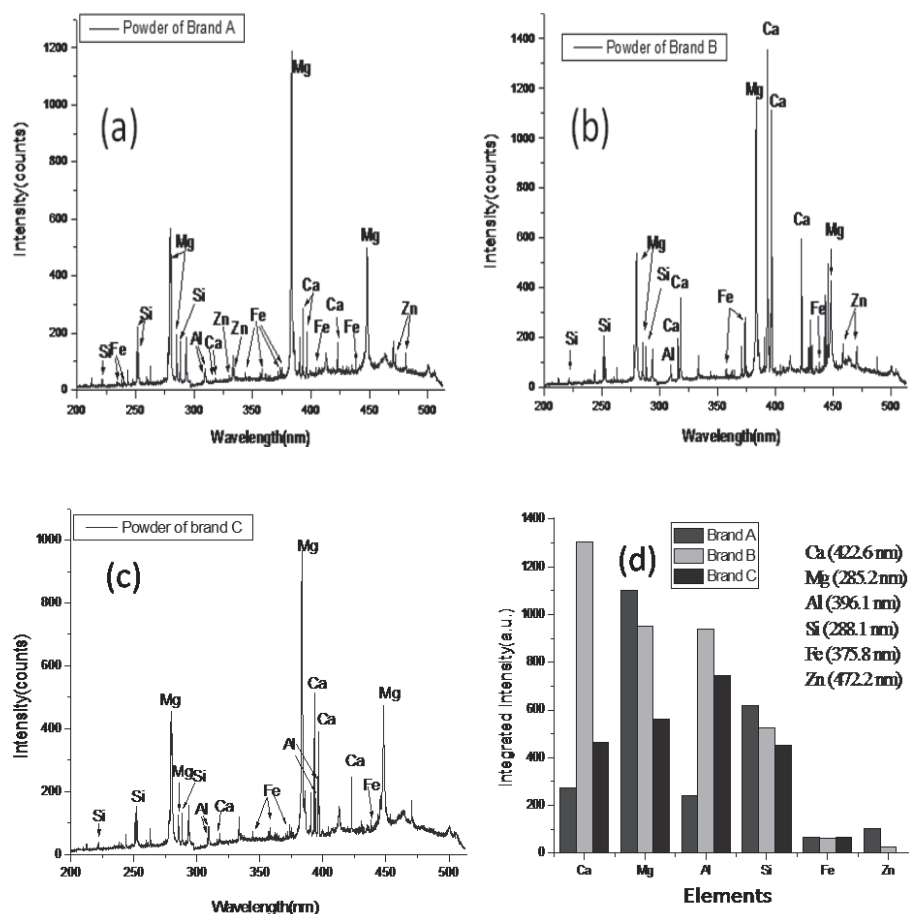


Figure 2: LIBS spectra of cosmetic powder of (a) brand A, (b) brand B, (c) brand C and (d) relative spectral intensity of elements in three brand samples

The LIBS spectra of the cosmetic powders of different brands have been collected from the local market of Allahabad. The pellets were prepared using KBr hydraulic press, and then the pellets are used as sample. The LIBS spectra of these samples have been recorded by focusing the laser light on the surface of pellet and are shown in Figure 2. It is clear from the Figure 2(a-b) that spectral signatures of Ca, Mg, Si, Fe, Zn and Al are seen in the LIBS spectra of cosmetic powders of Brand A and B, while spectral signatures of Ca, Mg, Si, Fe and Al (Figure 2c) are seen in the LIBS spectra of cosmetic powder of Brand C. Figure 2(d) show the relative spectral intensity of elements present in different brand samples. It is clear from Figure 2(d) that Ca and Al are highest in brand B and Mg, Si, Fe and Zn are higher in Brand A. Al is present in all the three brand samples which is one of the toxic elements hence all the samples are toxic for human skin. But, Al is highest in Brand B powder sample and hence it is more toxic in all samples.

Conclusion

The LIBS study of cosmetic powder of three brands have been studied and found the presence of spectral signature of Ca, Mg, Si, Fe, Zn and Al. Out of them Al is the toxic element for human skin and Fe and Zn are the heavy metals which may also be toxic for human skin. Thus our result clearly demonstrates the capability of LIBS to detect presence of the toxic elements in cosmetic powder samples and hence LIBS technique may be used as a monitoring tool for increasing the quality of such sample sold in the market.

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Water Quality Criteria, Some Common Pollutants their Toxic Effects and Remediation Technique: A Critical Review

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Abstract

Freshwater systems are very important to our society today. Aquatic ecosystems not only provide food and water for human consumption; they maintain and improve water quality by filtering out, storing, converting and diluting contaminants. Water quality criteria was defined by various national and international agency like World Health Organization (WHO), Central Pollution Control Board (CPCB) and bureau of Indian Standards. Among the various undesirable and naturally occurring pollutants in water, the coliform bacteria, fluoride, iron, nitrate and arsenic are very important as these cause serious health problems. Many methods have been used to remove fluoride, iron, nitrate, arsenic and coliform bacteria from water. Most of them are conventional and expensive. Conventional methods for removal of fluoride, arsenic and coliform bacteria involve coagulation followed by separation of the produced insoluble settling or by direct filtration through sand beds. Other methods include reverse osmosis, ion exchange, lime softening, flotation and adsorption on iron oxides or activated alumina. These conventional water treatment methods are not affordable in rural communities of developing countries.

Introduction

Water is a very precious gift of nature to all living things. Fresh water, essential for our life, is available as small portion of the total water available on this earth. About 2/3rd of the earth is covered by water. According to the global hydrological cycle, 97 percent of water always remains in the sea. Of the left-over 3 percent, nearly 1.998 percent is locked in the masses of ice caps and glaciers in the Poles and mountains and in the clouds, 0.85 percent falls directly on the sea as rainfall and 0.152 percent is available to humanity as utilizable water in the form of precipitation on land. However, out of the 0.152 percent, 0.15 percent flows through the rivers back to the sea as surface run-off, and only 0.002 percent of water is available as surface water in lakes / rivers and groundwater reserves, which could be harnessed to meet the demand for various designated best uses. Thus, for every 250 litres of global water, only one teaspoonful of fresh water (5ml) is available to mankind, rendering it a scarce and precious resource. [1]

Water Quality Criteria

Various National and International Agencies involved in water quality assessment and pollution control have defined water quality criteria for different uses of water considering different indicator parameters. Classification schemes for water quality criteria/standards developed by these agencies

differ in addition to terminologies used such as Action level, Guide level etc. in defining the concentration values in these classes. The concentration ranges in these classes are defined in Indian scenario considering Indian Standards [2] and Central Pollution Control Board (CPCB) criteria [3]. World Health Organization (WHO), Geneva, also gives guidelines for Drinking Water Quality [4]. WHO Guidelines are indicated in terms of Action Levels for important organic, inorganic and bacteriological parameters. The term Action Level indicates a level above which the reasons for the presence of a substance should be investigated and appropriate remedial measures is applied.

Water quality is a relative term used for proposed use of water for different purposes i.e. domestic, industrial, agricultural etc. but from user's point of view, water quality is defined as, "*The physical, chemical or biological characteristics of water by which the user evaluates the acceptability of water*", water required for drinking purpose should be pure, wholesome, and potable. For management of water quality, in India, the Central Pollution Control Board (CPCB), an apex body in the field of water quality management, has developed "*designated best use*". The CPCB has identified 5 such "*designated best uses*" as follows (Table 1.1).

Table1: Water quality criteria: Use based classification of surface water in India (According to CPCB) [3]

Designated Best use	Class	Criteria
Drinking Water Source without conventional treatment but after disinfection	A	1. pH between 6.5 – 8.5 2. Dissolved Oxygen 6 mg/L or more 3. BOD (5 days) at 20° C 2 mg/L or less 4. Total Coliforms organism MPN/100 ml shall be 50 or less
Outdoor bathing	B	1. pH between 6.5 – 8.5 2. Dissolved Oxygen 5 mg/L or more 3. BOD (5 days) at 20° C 3 mg/L or less 4. Total Coliforms organism MPN/100 ml shall be 500 or less
Drinking Water Source after conventional treatment and disinfection	C	1. pH between 6-9 2. Dissolved Oxygen 4 mg/L or more 3. BOD (5 days) at 20° C 3 mg/L or less 4. Total Coliforms organism MPN/100 ml shall be 5000 or less
Propagation of Wildlife and Fisheries	D	1. pH between 6-9 2. Dissolved Oxygen 4 mg/L or more 3. BOD (5 days) at 20°C 3 mg/L or less
Irrigation, Industrial cooling, Controlled waste disposal	E	1. pH between 6-8.5 2. Conductivity at 25°C in $\mu\text{S}/\text{cm}$ max. 2250 3. Sodium Absorption Ratio Max. 26 4. Boron Max. 2 mg/L

Table: 1.2 Indian standard specification for drinking water (IS-10500:2012) [2]

Parameters	Acceptable Limit
Colour, Hazen units, max.	10
Odour	Unobjectionable
Taste	Agreeable
Turbidity NTU, max.	1
pH value	6.5 to 8.5
Total Dissolved Solids mg/L, max.	500
Total Hardness (as CaCO_3) mg/L, max.	200
Calcium (as Ca) mg/L, max.	75
Magnesium (as Mg) mg/L, max.	30
Copper (as Cu) mg/L, max.	0.05
Iron (as Fe), mg/L, max.	0.3

Parameters	Acceptable Limit
Manganese (as Mn), mg/L, max.	0.1
Chloride (as Cl^-), mg/L, max.	250
Sulphate (as SO_4^{--}) mg/L, max.	200
Nitrate (as NO_3^-), mg/L, max.	45
Fluoride (as F^-), mg/L, max.	1
Arsenic (as As), mg/L, max.	0.01
Zinc (as Zn), mg/L, Max.	5
Chromium (as Cr^{6+}), mg/L, Max.	0.05
Pesticides, Max.	Absent
Alkalinity (as CaCO_3), mg/L, Max.	200

Table: 1.3 World Health Organization (WHO), Geneva, 2011, Guidelines for Drinking Water Quality [4]

Characteristics	Guideline Value	Unit
Arsenic	0.01	mg/L
Cadmium	0.003	mg/L
Chromium	0.05	mg/L
Fluoride	1.5	mg/L
Lead	0.05	mg/L
Mercury	0.006	mg/L
Nickel	0.07	mg/L
Nitrate	50	mg/L
Nitrite Nitrogen	3.0	mg/L
Selenium	0.04	mg/L
Chloride	250	mg/L
Sulphate	250	mg/L
Hardness as CaCO ₃	500	mg/L
Total Dissolved Solids	1000	mg/L
Aluminium	0.2	mg/L
Copper	2.0	mg/L
Iron	0.3	mg/L
Manganese	0.1	mg/L
Sodium	200	mg/L
Zinc	0.1	mg/L
2,4,6-Tri chlorophenols	200	µg/L
Chloroform	300	µg/L
DDT	1.0	µg/L
Lindane	2.0	µg/L
1,4-Dichlorobenzene	300	µg/L
Gross Alpha Activity	0.1	Bq/L
Gross Beta Activity	1.0	Bq/L
Colour	15	TCU
Turbidity	5	NTU
Taste	Not Objectionable	-
pH	6.5 to 8.5	-
Coliforms	Absent in 100mL	cfu

Some Common Pollutants

Iron is one of the most common elements in nature as it represents about 5% of the earth's crust [5], and it can be found in fresh waters at a concentration of 0.5-50 mg/L. Although iron has been classified as a secondary contaminant as it is not harmful for human beings, it causes aesthetic problems such as taste issues, turbidity, and discolouration [6]. Moreover, presence of iron motivates the bacterial growth (iron bacteria such as *Gallionella*, and *Leptothrix*), which block the plumbing in consequence [7-8].

The exposure to low or high concentrations of arsenic (As), either due to the direct consumption of As contaminated drinking water, or indirectly through daily intake of As contaminated food may be fatal to the human health [9-10].

About 80% of communicable diseases in the world are waterborne [11]. To reduce the incidence of water-borne diseases and make the water suitable for human consumption, the removal of pathogenic organisms, fecal matter, suspended solids, algae, organic matter and harmful chemicals is absolutely necessary [12-13]. Among the various undesirable and naturally occurring pollutants in water, the coliform bacteria, fluoride, iron, nitrate and arsenic are very important as these cause serious health problems [13].

Fluoride

Excessive fluoride concentrations have been reported in ground waters of more than 20 developed and developing countries including India where 17 states are facing acute fluorosis problems [14]. But on the contrary, the lower concentration of fluoride in water like 1 mg/L, is even found to be beneficial in preventing dental caries. However, excessive fluoride concentration in drinking water causes dental fluorosis, skeletal fluorosis and crippling skeletal fluorosis [4, 15, 16].

Defluoridation of drinking water is the only practicable option to overcome the problem of excessive fluoride in drinking water, where alternate source is not available. During last several years following the extensive research has been done on various methods for removal of fluoride from water and wastewater due to discovery of fluoride as the cause of fluorosis,. These methods are based on the principle of adsorption [17], ion-exchange [18], precipitation– coagulation [19, 20], membrane separation process [21, 22], electrolytic defluoridation [23], electrodialysis [24–26], etc.

Iron

Iron is one of the most abundant elements of the earth's crust. Groundwater easily gets contaminated with iron usually in its (+II) valence state. Although iron is an essential mineral for human, its presence in groundwater above a certain level makes the water unusable mainly for aesthetic considerations such as discoloration, metallic taste, odor, turbidity, staining of laundry and plumbing fixtures. Moreover, iron oxides, which are formed in reservoirs upon aerial oxidation of dissolved iron promotes growth of micro-organism in water. Thus, World Health Organization has set a guideline value of 0.3 ppm of iron in drinking water [4].

Aeration and separation is the most common method for removal of iron from groundwater in public water supply systems, which is, however, not so popular at domestic level. The iron content in surface water is very low due to natural aerial oxidation of iron to its oxides, which subsequently separate out. There are various other methods for iron removal from groundwater, viz., ion-exchange method [27], oxidation with oxidizing agents such as chlorine and potassium permanganate [28], activated carbon and other filtering materials [29], supercritical fluid extraction [30], bioremediation [31] and treatment with limestone [32].

Nitrate

Nitrate contamination of water resources has become an increasing problem globally due to the extensive use of nitrogen fertilizers and improper treatment of waste water from industrial sites [33]. Within the human body, nitrates may be reduced to nitrites that combine with haemoglobin to form methaemoglobin, which can be fatal to neonates [34]. The limit for nitrates in drinking water recommended by WHO is 45 ppm, and the higher levels are considered unsafe [4].

Because of health risks, several methods to reduce nitrate in drinking water have been presented [35-37]. Without addition of a carbon source, biological denitrification is difficult [38, 39]. The problem of contamination by dead bacteria has to be solved to make such processes safe enough to utilize in drinking water treatment [40]. Ion exchange and reverse osmosis processes [41] cannot transfer nitrate into harmless compounds but only concentrate nitrate from water to brine, requiring further treatment.

Arsenic

The arsenic problems are of more concerns for small communities in rural areas around the world, where groundwater comprises the main drinking source of water

[42]. In high concentrations, arsenic poisoning can also lead to an acute condition called arsenicosis [43-45]. The World Health Organization (WHO) revised the guideline for arsenic from 0.05 to 0.01 mg/L in 1993 [4].

There are number of methodologies were used to remove arsenic from aqueous system. The treatment methodologies chosen will depends upon on the scale of the system, arsenic concentration level in the water and other physiochemical quality of water. Arsenic cannot be destroyed; it only can be transformed from one form to another form. Inorganic form of arsenic i.e. arsenite and arsenate is generally present in real water whereas organic form occurs in saline water and biological system. It is inorganic species of arsenical compound that are generally treated. An extensive research has been done for removing excess arsenic from water which includes coagulation followed by lime softening, precipitation, membrane separation, ion exchange, adsorption etc [9].

Coliform Bacteria

Coliform bacteria are the indicator of contaminated water with human or animal wastes and if these are absent then only water can be considered as safe for drinking purposes. These bacteria are generally not harmful but other microbes along with these bacteria can cause short-term effects like diarrhea, cramps, nausea, headaches, and other symptoms [46].

Many methods have been used to remove fluoride, iron, nitrate, arsenic and coliform bacteria from water. Most of them are conventional and expensive. Conventional methods for removal of fluoride, arsenic and coliform bacteria involve coagulation followed by separation of the produced insoluble settling or by direct filtration through sand beds [47]. Other methods include reverse osmosis, ion exchange, lime softening, flotation and adsorption on iron oxides or activated alumina [48, 49]. These conventional water treatment methods are not affordable in rural communities of developing countries [50]. So other small-scale economical methods are needed for the decentralized communities. Various low-cost materials like kaolinite, bentonite, charfines, lignite and pumice were investigated to assess their capacity for removal of fluoride, arsenic and coliform bacteria from water by batch adsorption studied [51-53]. Slow sand, diatomaceous earth (DE) filtration and sandfilled glass columns having grown plants were also used for removal of *Giardia muris*, *cryptosporidium* and coliform bacteria from water [46, 54-57].

Conclusions

Various technologies had been used for the removal of fluoride, iron, nitrate, arsenic and coliform bacteria from drinking water but this problem has not still been rooted out. Several small-scale water treatment techniques are practiced to alleviate these problems especially for rural communities of developing countries. Homemade slow sand filter is one of the options which had been practiced for a long time and this is also economical to construct, operate and maintain, but the filtrate of this homemade slow sand filter does not fulfill the recommended guideline value suggested by World Health Organization (WHO), Central Pollution Control Board (CPCB) and bureau of Indian Standards in removing these pollutants and pathogens. The data have demonstrated that there are serious consequences to degraded water quality. The time has come to take a more intensive look into the drinking water quality to safeguard people's health. Therefore, for better cleanup of the environment, new technologies with the options of new hybrid technologies are needed to challenge the menace of contaminated water. Among various techniques, membrane technologies, electroremediation and phytoremediation are seemed to be more suitable for removal of As, F, Fe, nitrate and Coliform bacteria from water. However, further research is needed to develop more efficient techniques without drawbacks for providing pure drinking water to the common man.

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Greenhouse Gases and their Role in Climate Change

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Abstract

Global climate change has been a growing concern for International community. Climate data show that the terrestrial climate is remarkably sensitive to global forcing. It has been reported that the global climate is influenced by both natural forcing as well as anthropogenic forcing. In the past decades considerable speculation has been ongoing regarding the probable effects on the global climate of the increase in the atmospheric Carbon dioxide and other greenhouse gases. International panel on climate change (IPCC) in its report has said that human activity since industrial revolution has increased the amount of greenhouse gases in the atmosphere leading to global warming. In this paper, an attempt has been made to present comprehensive picture of the relative role of greenhouse gases in modulating the climate change.

Keywords: Global forcing, climate change, greenhouse gases, global warming.

Introduction

The United Nations Intergovernmental Panel on Climate Change has been studying global warming for years. Their most recent report, issued in February 2007 says that the globally averaged net effect of human activities has been one of warming climate. It has been reported that anthropogenic greenhouse gases are primarily responsible for the most of the observed temperature increase since the middle of the twentieth century. It is pertinent to note here that the global climate change has been affecting human health, agricultural productivity and global environment. It is in this backdrop that the international panel on climate change (IPCC) has issued warning to the scientific community to the alarming situation caused by climate change. The release of chemicals into the atmosphere has grown greater over the last century. Increased reliance on synthetic chemicals, deforestation, biomass burning and fossil fuel combustion have all contributed to the observed perturbations of trace chemicals in the atmosphere. This chemical change has effected an alteration of the earth's climate. Several research groups [Rama Nathan et al 1982, Hansen et al (1982, 1997, 2005b, 2007); Spahni, R. et al. (2005)] have undertaken detailed study for understanding combined effects of the trace atmospheric gases on the global climate. It is to be noted that trace chemicals modify the radiation energy balance of the earth- atmospheric system. Since carbon dioxide and methane absorb infra-red radiation, an expected effect of adding CO₂ and CH₄ to the lower atmosphere is to raise the average surface temperature of the earth. The direct lower

atmosphere warming is primarily responsible for global climate change. Therefore, scientific study of the problem is of fundamental importance for understanding the complex mechanisms involved in global climate change.

Greenhouse Effect

The "greenhouse effect" is a term used to describe the retention of heat in the earth lower atmosphere (troposphere) due to concentration of certain trace gases and water vapor in the atmosphere. These gases are known as greenhouse gases.

Greenhouse effect is responsible for the average global temperature and is favorable for sustaining life on the earth. However, as the concentration of green house gases in the atmosphere increases, the more heat energy remains trapped in the blanket of the atmosphere. The lag of heat transfer raises the total amount of heat present in the troposphere. Consequently, surface – air temperature goes on increasing resulting in global warming. Global warming in its turn causes global climate change.

Greenhouse gases and their abundance in the atmosphere These greenhouse gases are

1. Carbon Dioxide:

The CO₂ concentrations in the atmosphere have been measured to be increasing at a rate of approximately 1.5ppm/y [MRC (1983)]. Fossil fuels combustion is regarded as a major contributor to the increase in the concentration of atmospheric carbon dioxide. Wuebbles,

D. J; (1981) described the CO₂ concentration empirically as,

$$[\text{CO}_2] = 330.0e^{0.0056(t-1975)} \quad \text{---- (1)}$$

1975 < t < 2100

The CO₂ concentration of year 2030 is computed to be 450 ppm.

2. Methane

The most abundant atmospheric hydrocarbon, methane is believed to be increasing globally. From trapped air in dated ice cores, Craig and Chou (1982) have deduced that CH₄ concentrations have approximately doubled in the last 350 years with a greater rate of increase in the last century. Principal sources of atmosphere methane appear to be centric fermentation in sediments below shallow water bodies as rice paddies and quite possibly, production by termites and biomass-burning.

3. Nitrous Oxide

The most abundant nitrogen containing chemical from the climate viewpoint is N₂O. It is now accepted that N₂O has a very long atmospheric life-time (>100 yrs) with stratospheric photolysis the only known removal process. Microbial production in solids and oceans has been found to be source as well as a sink of N₂O. From the Weiss mathematical model (1981), Rama Nathan et al (1985) estimated the year 2030 concentration to be 375 ppb and a likely range of 350-450 ppb.

4. Ozone

The climatic effects of ozone change depend very strongly on whether troposphere or stratosphere ozone is being altered [Rama Nathan and Dickinson, 1979; Wang, 1982]. Data suggest that ozone concentrations are increasing with time. National Aeronautics and space Administration for the 2Km to 8 Km region reported increase in ozone concentrations by (8+4) %.

5. Chlorofluorocarbons (CFC's)

These chemicals came into major use in the 1960's and initially exhibited a rapid growth. The most important CFC's to date have been CFCL₂ (F12) and CFC's (F12 or F11). THE global emissions of the major CFC's actually declined due to ban on some nonessential usages. CFC emissions increased sharply in 1983, [Wuebbles (1983b)].

A brief description of climate model

Several numerical models have been used in early studies to estimate the impact on the atmosphere of CO₂ increase [Rama Nathan et al. 1979, Hansen et al. 1981]. Although there is uncertainty, modeling results consistently predict an increase in troposphere temperature under increasing CO₂. Most of the early models show the largest temperature increase occurring in the polar or sub-polar regions. Because of the overlap between proxy data and the instrumental data, researchers have to use sophisticated computer analysis and statistical methods to reconstruct yearly temperatures with a certain range of uncertainty. A key finding from the records is that the last few decades of the twentieth century were unusually warm in comparison with preceding centuries. Hansen (2004) with the help of climate model simulations has suggested that CO₂ emissions can be stabilized and other greenhouse gas concentrations can be substantially decreased in the future.

Global Forcing

A forcing is an impressed change of the planet's energy balance with space. The most common technical measure of forcing [Hansen et al (1997), IPCC (2001)], is the adjusted forcing (Fa). The adjusted forcing is the imbalance, caused by the forcing agent, between solar energy absorbed by the planet and thermal emission to space, measured after stratospheric temperature adjusts to presence of the agent. The effective forcing due to all three long- lived Greenhouse gases is given by

$$F_e = 1.15 [F_a(\text{CO}_2) + 1.4 F_a(\text{CH}_4)] \quad \dots (2)$$

Some forcing agents have greater efficacy than others for changing global temperature [Hansen et al. (2005b)]. This is the reason for different numerical coefficients in the equation (2). On the basis of the equation (2), Hansen et al (2007) have inferred that CO₂ is the largest human- made climate forcing, but other trace gases are also important. The most important of the non- CO₂ forcing is methane, as it causes the second largest human- made GHG climate forcing. It is also the principal cause of increase in troposphere Ozone.

Discussion

From the perusal of the estimates of the abundance of the atmospheric trace gases mentioned in the table (1), we have gained some ideas regarding effects of green house gases on global climate. The earth's atmosphere currently contains trace gases with atmospheric life- times that vary from much less than 1 hour to several hundred years. The abundance of

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trace gases is therefore dictated as much by their removal rates as by the growth in sources. From a viewpoint of global climate effects, species with extremely short life times are unlikely to play an important direct role. Since N₂O climate forcing is a fraction of the total greenhouse gases forcing and because N₂O time-variations are similar to those all three long-lived greenhouse gases. In this direction, a lot of research work has been done.

Our analysis suggests that present knowledge does not permit accurate specification of the dangerous level of human-made greenhouse gases. With Greenhouse gases

continuing to increase, the planetary energy imbalance provides ample energy to melt ice for raising the sea-level. With this danger in mind, it is appropriate to monitor ice-sheet conditions. For averting ice-sheet disintegration, efforts should be to reduce both CO₂ emissions and non-CO₂ climate forcing. Therefore, it is necessary to develop some sophisticated model which could throw some light on the mechanisms involved in the CO₂ and CH₄; it is possible to reconstruct accurately the climate forcing caused by the sum of modulation of climate change.

Table 1: Estimate of the abundance of the trace gases (chemicals) in the global atmosphere of 1980 and 2030.

Chemical group	Chemical formula	Dominant source	Dominant sink	Year 1980 Global average mixing ratio ppb *	Year 2030, probable global average concentration (ppb) best estimate/ possible range	Life-time (years)
Carbon dioxide	CO ₂	Natural and Anthropogenic	Oceanic	339×10 ³	450×10 ³	2
Nitrogen compounds	N ₂ O	Natural and Anthropogenic	Stratosphere (UV)	300	375, /350-450	120
Hydrocarbons	CH ₄	Natural	Troposphere(OH)	1650	2340, /1850-3300	5-10
Ozone	O ₃	Natural	Troposphere(UV;SL,O)	25-70	12.5%	0.1-0.3
Chlorofluorocarbons	CCLF ₃ (F ₁₃)	Anthropogenic	Stratosphere(UV)	0.007	0.06, / 0.04-0.1	400
	CCLF ₂ (F ₁₂)	do	do	0.28	1.8, / 0.9-3.5	110
	CCLF(F ₁₁)	do	do	0.18	1.1, / 0.5-2.0	65

Source: Data have been collected from scientific literature, WMO (1982b) and Ramnathan et al.(1985), Journal of Geophysical Research; Vol. 90, no: D3 p-5547 to 5566

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Study on Islanded Wind Power Transmission Processes

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Abstract

Wind energy around the world produced 273 TWh of electricity in 2009 from an estimated installed capacity of 159 GW. Wind power development is booming globally and large scale onshore wind power is under development in China, the Americas and India. Compared with AC transmission, HVDC technology can due to lower losses be attractive if the distance from generation to consumption is long. VSC (Voltage Source Converter) HVDC is rapidly maturing as a technology, and offers some significant advantages compared to traditional LCC (Line Commutated Converter) HVDC, such as independent control of active and reactive power. This independent control enables VSC HVDC to operate in very weak / islanded AC systems; however LCC HVDC still has a considerably higher DC current capability than VSC HVDC. Onshore wind conditions require both very high DC current capability, and control flexibility due to very weak/islanded AC systems.

Keywords: Wind energy, AC transmission, HVDC technology, Voltage source converter (VSC), STATCOM and line commutated converter (LCC)

Introduction

Wind power development is booming globally and large scale onshore wind power is under development in China, the Americas and India. Compared with AC transmission, HVDC technology can due to lower losses be attractive if the distance from generation to consumption is long. VSC (Voltage Source Converter) HVDC is rapidly maturing as a technology, and offers some significant advantages compared to traditional LCC (Line Commutated Converter) HVDC, such as independent control of active and reactive power. This independent control enables VSC HVDC to operate in very weak/islanded AC systems; however LCC HVDC still has a considerably higher DC current capability than VSC HVDC. Onshore wind conditions require both very high DC current capability, and control flexibility due to very weak / islanded AC systems. Therefore, there is good reason to investigate new solutions to transmit large scale onshore wind power via LCC HVDC, which complements existing VSC and LCC HVDC technology. In this paper, a novel control method that enables LCC HVDC combined with a small STATCOM to transmit the rated HVDC power capacity from an islanded large scale PMSG (Permanent Magnet Synchronous Generator) wind farm AC system is presented. As known from micro grids [2], an islanded AC system with full converter power sources changes the

fundamental behaviour of the AC system; the AC voltage amplitude depends on the active power balance instead of the reactive power balance; the AC voltage frequency depends on the reactive power balance instead of the active power balance. Considering these non-conventional relations, an LCC HVDC rectifier is used to control the AC voltage amplitude, and a small STATCOM is used to control the AC voltage frequency. Wind power development is booming globally and large scale onshore wind power is under development in for instance China, America and India. It is also perceivable that the abundant onshore wind resources in for instance Mongolia and Kazakhstan could be utilized in the future by large capacity, long distance HVDC transmission. Feasibility studies regarding arctic wind power has also been performed, where for instance northern Russia is shown to be very rich in arctic wind resources. Onshore wind conditions therefore require both very high DC current capability, and control flexibility due to weak/islanded AC systems with fluctuant power. Therefore, there is good reason to investigate new solutions to complement existing VSC and conventional HVDC technology. In this paper, a novel control method that enables LCC HVDC combined with a relatively small STATCOM (~ 0.06 p.u of the HVDC) to transmit the rated HVDC power capacity from an islanded

large scale PMSG (Permanent Magnet Synchronous Generator) wind farm AC system, is presented.

Mostly due to the German energy transition, there have in recent years been large amounts of offshore wind power integrated by VSC HVDC. The offshore station is of special interest since it needs to operate in a completely islanded AC network. By firstly forming an offshore AC voltage that is fixed in amplitude, frequency and phase, the wind generators can relatively easily be started up. Then all power generated by the offshore wind farm is transmitted by the HVDC link. A DC Chopper is used, where the onshore DC Chopper can temporarily burn the active power produced by the offshore wind farm. This way the offshore AC network does not have to deal with a sudden and large active power surplus. The onshore placement of the Chopper is purely an economical choice, there is practically nothing preventing placing a Chopper offshore. This is because offshore systems are cable based, hence there is no need to consider temporary DC faults. DC cable faults are not temporary, and in addition they are very costly to repair. Based on this, the whole power transmission must trip immediately should there be a DC fault.

A typical PMSG wind turbine has a generator side converter, and a grid side converter. The grid side converter is a voltage source converter, which interfaces the AC grid via an inductance. Based on the voltage difference between the AC grid voltage and the controlled converter AC voltage, there will be a controlled current injection from the generator. Therefore with a typical control scheme, a PMSG wind turbine will appear as a current source generator in the AC power system. Two fundamental AC circuits are firstly compared with each other.

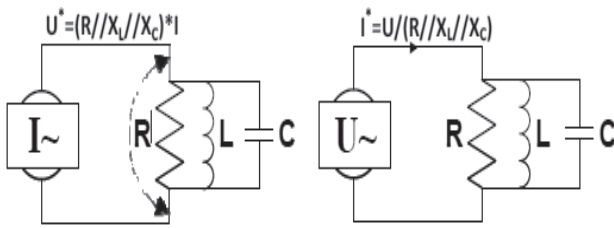


Figure 1

Both circuits have a resistance R . There is also one capacitor C with the resulting impedance X_C , and an inductance L with the resulting impedance X_L connected in parallel. For the fundamental AC frequency, the impedances X_C and X_L are equal but with different signs. Thus they cancel out, leaving the AC circuit purely resistive. For the current source

generator circuit on the left hand, the resistance R decides the voltage amplitude across the circuit, due to the current which the current source drives and the higher the resistance, the higher the voltage U . For the voltage source generator circuit on the right hand, the resistance R decides the amplitude of the current flowing in the circuit, which the voltage source drives and the higher the resistance, the lower the current I .

In Figure 2, it is assumed that there are three main components in the onshore AC network, namely the PMSG based wind farm, the LCC HVDC rectifier, and a supporting AC network consisting of conventional voltage source generators.

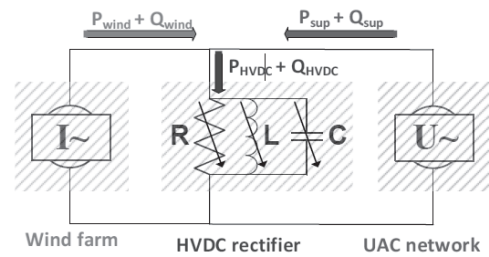


Figure 2

The active and reactive power that is consumed by the HVDC rectifier, is hence the sum of the two sources

$$P_{HVDC} = P_{Wind} + P_{Sum}$$

The LCC HVDC rectifier is modelled as a variable apparent resistance R and variable apparent inductance L , with a variable apparent capacitance C in parallel. At fundamental AC frequency, the reactive power consumption of the inductance L is cancelled by the reactive power generation of capacitance C . An islanding of this system means that the supporting AC network is disconnected, as shown in Figure 3.

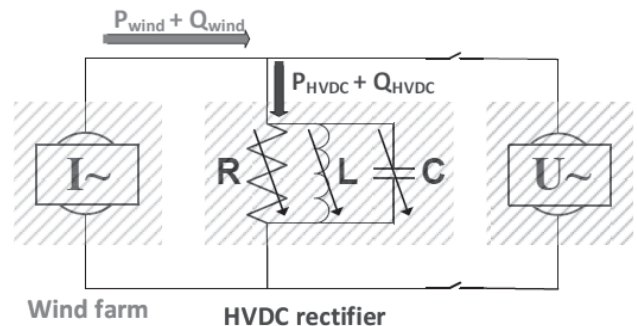


Figure 3

Hence in an AC network with only current sources, the amplitude of the AC voltage can be controlled, by controlling the active power order of the LCC HVDC

system. This means that the frequency can then in turn be controlled, by controlling the reactive power drawn. Hence the AC voltage amplitude can be controlled, by controlling the LCC HVDC active power order. The difference between measured and ordered AC voltage amplitude is firstly sent to a dead band, and then to a normal PI regulator. Finally the output delta power order is added to the existing power order. Since reactive power consumption is a function of active power for LCC HVDC, it is not possible to control the reactive power independent of active power. Therefore a STATCOM is used to balance out the relatively small difference between the reactive power consumed by the LCC HVDC, and the reactive power generated by the related AC filters and shunt capacitors.

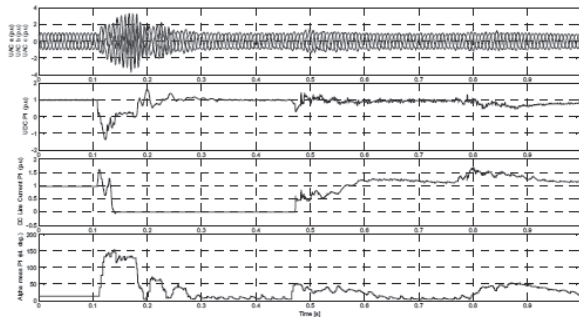


Figure 4

This way the frequency can be controlled in the islanded system. Since the HVDC system is unable to transmit any active power on either pole, the AC voltage amplitude

quickly becomes extremely high, peaking at ~ 3.4 p.u. As soon as both DC Choppers are connected, the AC voltage amplitude can be controlled down to its nominal value. As shown in figure 4.

Conclusion

In this paper, a novel control method for islanded wind power transmission was presented. By identifying and using the special characteristics of an islanded AC network consisting of only PMSG wind generators, both the AC voltage amplitude and frequency can be controlled by a LCC HVDC rectifier combined with a relatively small STATCOM.

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Study of a New Fractal Antenna which is Applicable in 1 GHz -- 30 GHz Frequency Range

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Abstract

With advancement in communication technology over the past decade, there is an increasing demand for miniaturization, cost effective, multiband and wideband antennas. Wireless applications, particularly with multiple resonances, put new demands on antennas pertaining to size, gain, efficiency, bandwidth, and more. One promising approach in this regard is to use fractal geometries to find the best distribution of currents within a volume to meet a particular design goal. In this paper, we have investigated a new fractal antenna with multi-band and broad - band properties. The proposed design is a loaded the 2nd iteration of a new fractal geometry to a square loop antenna. The simulation is performed with HFSS simulator software. The simulation results show that the proposed antenna is applicable in 1GHz to 30 GHz frequency range. Radiation patterns are also studied.

Keywords: Multiband antenna, fractal geometry, radiation pattern, bandwidth and HFSS software.

Introduction

In the present scenario of wireless communication, people demand high data rate and reliable communication. Federal Communication Commission (FCC) has allocated the unlicensed frequency band of 3.1GHz to 10.6GHz with an EIRP of – 41.3 dB m / MHz to avoid interference with present narrow band communication systems. This has drawn the attention of scientists and researchers to the usage of UWB towards short-range high - data rate wireless communication applications as well as long-range but low data rate wireless connectivity applications. After the announcement of unlicensed 7.5 GHz of spectrum, there has been wide variety of applications developed for indoor / outdoor communications, high accuracy radars and imaging systems etc.

In UWB Communication systems, antenna is one of the most critical components to be realized in order to have a good system performance. One of the good solutions is using fractal geometry to design of UWB antennas. Applying fractals to the antenna elements allows for smaller size, multi-band and broad-band properties.

Fractals have self-similar shapes and can be subdivided in parts such that each part is a reduced size copy of the whole. Self-similarity of fractals causes multi-band and broad-band properties and their complicated shapes provide design of antennas with smaller size. Fractals have convoluted and jagged shapes with many corners that these discontinuities increase bandwidth and the effective radiation of antennas. Fractals can be placed long electrical length in to a small area using their ability of space - filling. Fractals describe by one

algorithm or repetitive function, iterating several times that integral dimension does not define for them.

Several wire antenna configurations based on fractal geometries have been investigated including Koch, Minkowski, Hilbert, and fractal tree antennas in recent years. These antennas have been simulated using the moment method and discussed. .

In this paper, new fractal geometry is presented. By applying this fractal generator to square loop antenna elements, we have achieved a super wideband antenna.

The method of moment method and HFSS software has been used for the design and simulation of the proposed antenna. According to the results this new fractal antenna is applicable in 1 GHz --30 GHz. Also, the radiation patterns are studied in multi frequencies.

Antenna Geometry

The geometric construction of this new fractal curve starts with a straight line, called the initiator, which is shown in Figure 1 ($n = 0$). This is partitioned in to four equal parts and the two centric segments are replaced with three others of the same length which are formed like a trapezoid with the indentation angle $\theta = 60^\circ$. This is the first iterated version of the new fractal geometry and is called the generator, which is shown in figure 1 ($n = 1$). The process is repeated in the generation of the 2nd iteration, which is shown in figure 1 ($n = 2$).

The proposed design is based on a loaded the 2nd iteration of the new generator to a square loop antenna. If the length of each side at square is assumed to be X with applying the 2nd iteration of this generator, the length will be $X = (5/4)^2$

In this paper, we have supposed the length of each side at square loop is 8 cm and 2 mm in diameter. The feed location is placed at the corner. The structure of this fractal antenna is shown in figure 2.

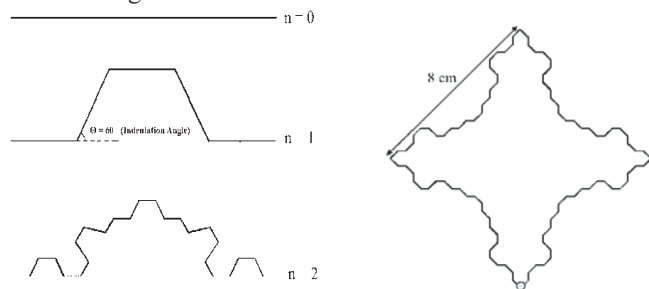


Figure 1 & 2: Iterations of the proposed fractal geometry and antenna structure

Simulation Results and Discussion

The MOM (method of moments) is a very powerful technique which can be applied to the analysis of three dimensional structures. MOM is used for simulation of this antenna based on the HFSS simulator software. The wire conductivity of all conductors is assumed to be 5.7×10^7 . The voltage source is 1 volt and the frequency range is from 1 GHz - 35 GHz. Figure 3 shows the simulated VSWR in entire bandwidth.

According to the simulated VSWR, this super wideband antenna is applied between frequencies 1 GHz - 30 GHz and approximately can be matched with a 50Ω coaxial cable. Also, Figures 4 (a) and 4 (b) present real and imaginary parts of the input impedance.

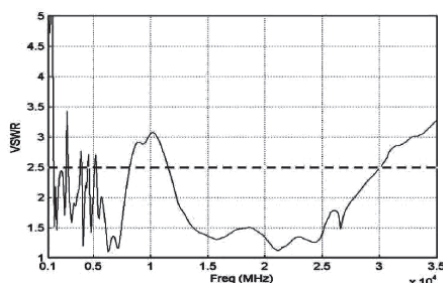


Figure 3: Proposed antenna VSWR

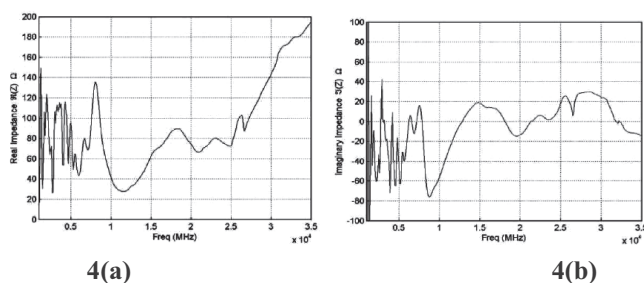


Figure 4 (a) & 4 (b): Real part of the input impedance and imaginary part of the input impedance respectively

Conclusion

The concepts of fractals can be applied to the design of ultra wideband antennas. Applying fractals to antennas allows for miniaturization of antennas with multi-band and broad-band properties. The proposed design is a loaded 2nd iteration of new fractal geometry to a square loop antenna. The proposed structure has dimensions of $8 \times 8 \text{ cm}^2$.

The simulated results conducted by the HFSS simulator. Due to the simulation results, the proposed design is a super wideband antenna which is applicable for frequencies between 1 GHz - 30 GHz. This super wideband antenna is simple to design and easy to fabricate. The proposed antenna is of small size, low cost and can be integrated with MMIC/MICs. This antenna can be a good candidate for UWB wireless communication applications which are going to proliferate in wide variety of ways in all wireless scenarios.

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Study of Design and Characteristics of a 2.42 GHz Inset Fed Microstrip Patch Antenna for RFID Applications

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Abstract

Microstrip antennas in recent years have been one of the most innovative topics in antenna theory and design. The basic idea of microstrip antenna came from utilizing printed circuit technology not only for the circuit component and transmission lines but also for the radiating elements of an electronic system. Microstrip patch antenna offers a potential solution and therefore is a popular printed resonant antenna for narrow - band applications. Radio Frequency Identification (RFID) systems, due to recent technological advances, have been used for various advantages in industries like production facilities, supply chain management etc. Since used in wide range of applications, RFID systems demand low cost and low profile antennas. In this paper, a 2.42 GHz inset fed patch antenna has been designed for RFID applications with a gain of 4.6772dB and a return loss of -21.0579 dB.

Keywords: Microstrip patch antenna, Radio Frequency Identification (RFID), narrow bandwidth, printed resonant antenna and return loss and HFSS software.

Introduction

Over the last few years, RFID (Radio Frequency Identification) has drawn a great deal of attention and it is now widely believed that RFID can bring revolutionary changes. RFID is based on radio communication for tagging and identifying an object. It consists of two blocks namely, RFID transceivers (readers) and RFID transponders (tags). A RFID tag consists of a small integrated circuit for storing information and an antenna for communication. A basic RFID system is based on wireless communication between a reader and a tag. RFID readers can read information stored in no line-of sight RFID tags in their vicinity and communicate it to central database system through wired or wireless interface. Applications of RFID systems include supply chain automation, security and access control, cold chain management (temperature logging) and identification of products at check-out points, to name a few. Some of the major retailers have already invested significantly in RFID and mandated their manufacturers to place tags on cases and pallets, which resulted in mass production of inexpensive RFID tags. In recent years, more efforts have been made to implement RFID applications in inventory control and logistics management. RFID based system leads to significant reduction on processing time and labour as inventory in warehouses can be tracked more accurately in a simple, timely and more efficient manner. More importantly, RFID based system provides complete visibility of accurate inventory data, from manufacturer's warehouse, shop floors

and brings opportunities for improvement and transformation in various process of the entire supply chain.

Integration of RFID systems with wireless sensors has broadened the scope of RFID applications. RFID tags can be interfaced with external sensors such as shock, temperature, and light sensors. Similar to wireless sensor networks, RFID systems can be deployed on-line instead of pre-installed statically.

RFID readers are transceivers which transmit and receive at the same time at the same frequency. To maximize read range, readers typically transmit with maximum allowable power using directional antennas with high gain. For proper decoding and detection of weak tag signals, good isolation between transmit and receive channels is very important. Figure 1 shows a typical RFID system.

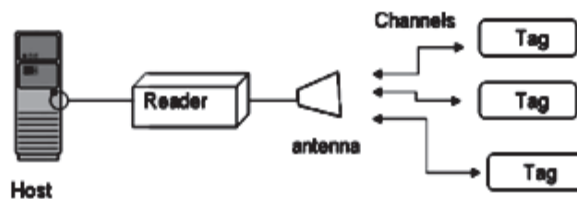


Figure 1: Schematic diagram of RFID system

In a RFID system, the reader is at heart, a radio transceiver that communicates with tags. Many different varieties of reader antennas are available and the proper choice is often guided by the regional regulations for maximum allowable radiated power and antenna beamwidth. As per applications

demand, RFID systems can use advanced antennas such as switched beam antennas, smart antenna arrays, etc. To reduce the size of the antenna and to achieve wideband, different techniques such as meandered ground plane, chip loading, feed modification, stacked shorted patch, slot-loading and teardrop dipole in an open sleeve structure have been reported. Several frequency bands have been assigned to the RFID applications, such as 125 kHz, 13.56, 869, 902 - 928 MHz, 2.45 and 5.8 GHz. According to ISO -18000 (International Standard Organization), the operated frequency range of the 2.45 GHz RFID band is 2400 - 2483.5 MHz, and the bandwidth is about 83 MHz, i.e. a narrow bandwidth compared to UHF RFID band of 860-960 MHz.

The frequency of 2.4 GHz is higher than that of UHF RFID so the size of the tag antenna is smaller. Thus, the application of 2.45 GHz RFID system can be applied to the smaller objects and systems.

This paper presents the design and simulation of an inset fed microstrip patch antenna. The antenna is designed at resonant frequency of 2.425 GHz and was intended for RFID applications.

Proposed Antenna Design and Geometry

A. Microstrip Patch Antenna

Microstrip antennas, in recent years, have been one of the most innovative topics in antenna theory and design. The basic idea of microstrip antenna came from utilizing printed circuit technology not only for the circuit component and transmission lines but also for the radiating elements of an electronic system. They are used in wide range of modern microwave applications because of their simplicity and compatibility with printed-circuit technology. It is easy to manufacture and either as stand-alone elements or as elements of arrays. A microstrip antenna in its simplest form consists of a rectangular (or other shapes such as circular, triangular ... etc) on top of a substrate which is backed by a ground plane. Fig. 2 shows a typical microstrip patch.

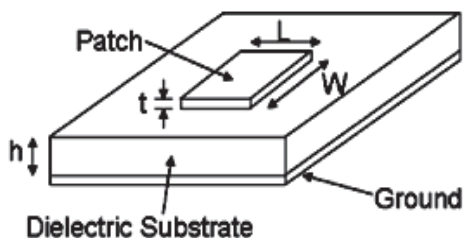


Figure 2: Microstrip patch antenna

There are various types of feeding methods available for patch antennas. In general, the microstrip antenna is fed

either by strip line or coaxial line. In either case, electromagnetic energy is first guided along the feed line to the region under patch, which acts like a resonant cavity with open circuits on the sides. Some of the energy leaks out of the cavity and radiates into space, resulting in antenna. There is another form of non-contracting feed available, called aperture coupled. The aperture coupled feed eliminates feed-line radiation and also allows thick substrate as probe reactance is not an issue.

Microstrip antennas radiate a relatively broad beam broad-side to the plane of the substrate. Thus the microstrip antenna has very low profile, and can be fabricated using printed circuit techniques. This implies that the antenna can be made conformable and potentially at low cost.

In microstrip antennas, linear and circular polarizations are possible with simple feed. They can easily be used to fabricate dual-frequency and dual-polarization antennas.

Some other advantages include easy fabrication into linear or planar arrays, and easy integration with microwave integrated circuits.

Some of the disadvantages of microstrip antenna configurations include narrow bandwidth, spurious feed radiation and poor polarization purity, excitation of surface waves, limited power capacity, and tolerance problem. In microstrip antennas, the dielectric and conductor losses can be large for thin patches resulting in poor antenna efficiency. Sensitivity to environmental factors such as temperature and humidity are also an issue with microstrip antennas.

However, narrow bandwidth disadvantage of patch antenna turns out to be advantage for RFID reader as RFID applications do not need much bandwidth because antenna rejects the signals that are out of the band and accordingly increases the quality factor.

Design and Configuration of Patch Antenna

In modern communications low cost and low profile antennas are in demand. microstrip patch antenna due to its conformal nature and capability to integrate with the rest of the printed circuitry satisfy those requirements. In antenna design feeding mechanism plays an important role. Coaxial probe feeding is sometime advantageous for applications like active antennas, while microstrip line feeding is suitable for developing high gain microstrip array antennas.

In general, patch antennas have the length of half - wave structures at the operation frequency of fundamental resonant mode. Since the fringing field acts to extend the effect length of patch, the length of the half - wave patch is slightly less than a half wavelength in the dielectric substrate material.

Approximate value for the length of a resonant half - wavelength path is given by

$$L = 0.49 \frac{\lambda}{\sqrt{\epsilon_r}} \quad \text{..... (1)}$$

Where λ is the free - space wavelength and ϵ_r is the substrate dielectric constant.

The width is given by

$$W = \frac{c}{2f_0} \sqrt{\frac{2}{\epsilon_r + 1}} \quad \text{..... (2)}$$

Where f_0 is the resonant frequency of the patch antenna, c is the free - space velocity of light. The effective dielectric constant for the case of ($w/h > 1$) is given by

$$\epsilon_{eff} = \frac{\epsilon_r + 1}{2} + \frac{\epsilon_r - 1}{2} \left[1 + 12 \frac{h}{W} \right]^{-1/2} \quad \text{..... (3)}$$

The extension of patch length due to fringing effects can be determined by

$$\frac{\Delta L}{h} = 0.412 \frac{(\epsilon_{eff} + 0.3) \left(\frac{W}{h} + 0.26 \right)}{(\epsilon_{eff} - 0.258) \left(\frac{W}{h} + 0.8 \right)} \quad \text{..... (4)}$$

The effective length of patch after taking into fringing effect can be calculated by

$$L = \frac{c}{2f_0 \sqrt{\epsilon_{eff}}} - 2\Delta L \quad \text{..... (5)}$$

The 2.45 GHz is taken as a resonant frequency of the microstrip patch antenna. Rogers Duroid material has been selected for the substrate with dielectric constant 2.2. The height of substrate is kept at 0.8 mm.

In inset fed microstrip patch antenna the location of feed determines the antenna input impedance. Therefore, to achieve 50Ω impedance, we the inset length (y_0) as shown in Fig. 3 by using the relation.

$$R_{in}(y = y_0) = R_{in}(y = 0) \cos^2\left(\frac{\pi}{L} y_0\right) \quad \text{..... (6)}$$

Where $R_{in}(y = 0)$ is the input impedance at the leading radiating edge of the patch and $R_{in}(y = y_0)$ is the desired input impedance (50Ω).

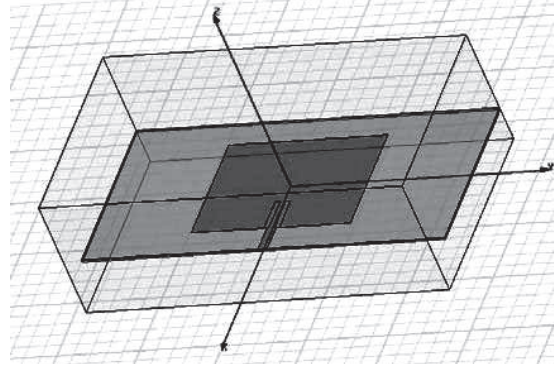
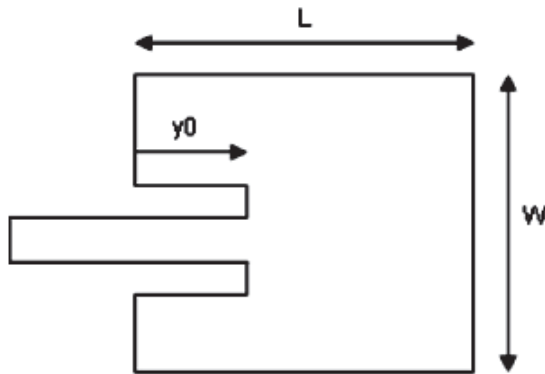


Figure 3: Inset fed antenna and **Figure 4:** Proposed antenna structure

The antenna has three layers or three blocks: first for the radiating plate, second for the microstrip line, and third for the substrate. After optimization, we got improved results at patch length of 41.21 mm and width of 48.4 mm. The inset feed length was optimized to 12.5 mm. The feed line is a 50Ω transmission line, is directly connected to the port. Simulation of the proposed antenna has been done by HFSS software. An optimization of the antenna parameters was performed to obtained satisfactory results. As shown in Fig. 5, the obtained input return loss (S_{11}) is equal to -21.0579 dB which is in good agreement with the obtained results. Fig. 6 shows the antenna gain and Fig. 7 presents the radiation pattern of proposed antenna which is in good agreement with the proposed derived result.

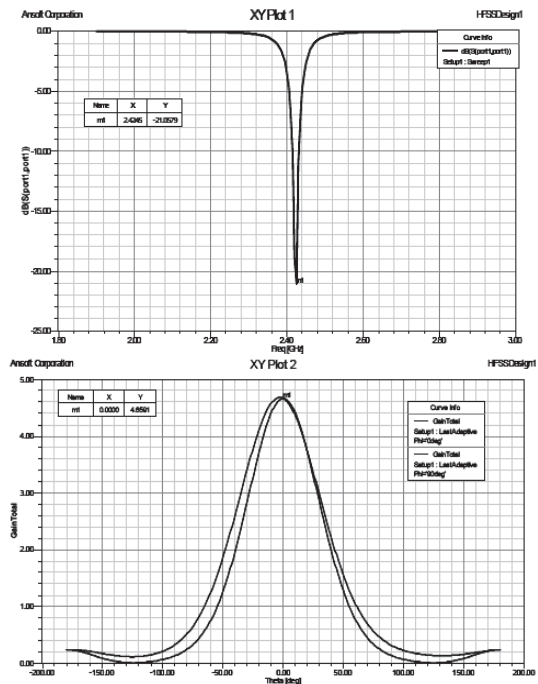


Figure 5: Return loss S_{11} and **Figure 6:** Frequency gain

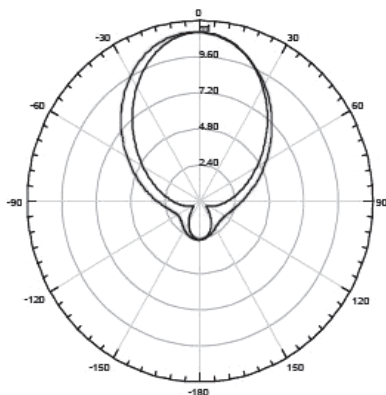


Figure 7: Radiation pattern

Conclusion

In this paper, the basic properties and applications of the RFID antennas was discussed. . RFID antenna engineering represents a relatively new field of research that combines the attributes with antenna theory. Research in this field has recently yielded a rich class of new designs for antenna elements as well as arrays. An inset fed microstrip patch antenna was designed and simulated at resonant frequency of 2.42. An input return loss of - 21.0579 dB and a gain of 4.6772dB were obtained. The proposed design has simple design structure and can easily be constructed at low cost. This antenna design can be a good solution for many RFID applications such as warehouse, supply chain automation etc.

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Noise Pollution, Cause and Effects

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Abstract

Sound is that part of energy, which excites us, the sensation of hearing. This is because some type of energy originating from the source and is entering our ear, sound becomes unwanted when it either interfaces with normal activities such as sleeping, conversion or disrupts or diminishes one's quality of life and such sound is called unwanted, hence we call it sound pollutant or noise. The traditional definition of noise is also “unwanted or disturbing sound”. A sound might be unwanted because it is Loud, unpleasant and intrusive or distracting.

A constant hearing of sound of level above 120 dB can cause headache and permanent damage to the ears of the listeners. The sound of level 30 dB to 10 dB has the soothing sensation, while the level 0 dB of loudness of sound represents the limit of hearing. We should not confuse about the limit of hearing with the limit of audibility. The limit5 audibility refers to the frequency of sound 20 Hz to 20 KHz which a man can hear, while the limit of hearing is the level of loudness (= 0dB) which a man can hear. Noise is also considered a mixture of many different sound frequencies at high decibels level. No one on earth can escape the sounds of noise. Noise is a disturbance to the human environment that is escalating at such a high rate that it will become a major threat to the quality of human lives. Noise in all areas, especially in urban areas, has been increasing rapidly. There are numerous effects on the human environment due to the increase in noise pollution. In this paper, cause and effects of noise pollution will be presented in some details.

Keywords: Loudness, unpleasant, decibels, Audibility, intensity, sensation and annoying.

Introduction

Most of the society is now aware that a noise can damage hearing. However, short of a threat that disaster would overtake the human race if nothing done about noise, it is unlikely that many people today would become strongly motivated to do something about the problem. Yet the evidence about the ill effects of noise does not allow for complacency or neglect. For instance, researchers working with children with hearing disorder are constantly reminded of the crucial importance of hearing to children. In the early years the child can not learn to speak without special training if he has enough hearing loss to interface effectively with the hearing of words in context.

There are two types of hearing loss : Conductive and Sensorineural. In conductive deafness sound pressure waves never reach the cochlea, must often as a consequence of a ruptured eardrum or a defect in the ossicles of the middle ear. Conductive hearing loss can be minimized, even overcome by use of the familiar hearing aids. The most common is worm over the mastoid bond behind the pinna. It picks up sound waves and transmits them through the skull to the cochlea.

Sensorineural hearing loss, the most common form in the United States, occurs as a result of advancing age as well as exposure to loud noises. In both instances there is a disruption of the organ of corti. The organ serves two functions: converting mechanical energy to electrical and dispatching to the brain a coded version of the original sound with information bout frequency, intensity and timbre. The hair cells of the organ of corti send their electrochemical signals into the central nervous system are picked up by thousands of auditory nerve fibres and transmitted to the brain. It is the decoding of all the information that enables a person to distinguish the unique and separate sounds of a violin, trumpet and clarinet even all three are playing the same note.

The organ of corti a gelatinous mass is one of the best protected parts of the body encased as it is within the cochlea which in turn is deeply embedded in the temporal bone, perhaps hardest of the 206 bones. None the less, loud noise can damage the hair cells and the auditory nerve producing at times depending on the types of noise, sudden and often total deafness.

Sustained noise over a period of time can also engender sensorineural deafness in the form gradual losses in hearing. This is the most common loss in teenagers today listening to loud rock music.

Until a few years ago, sensorineural deafness could not be helped by hearing aids. However with advances in electronic wizardry and into the auditory canal are available.

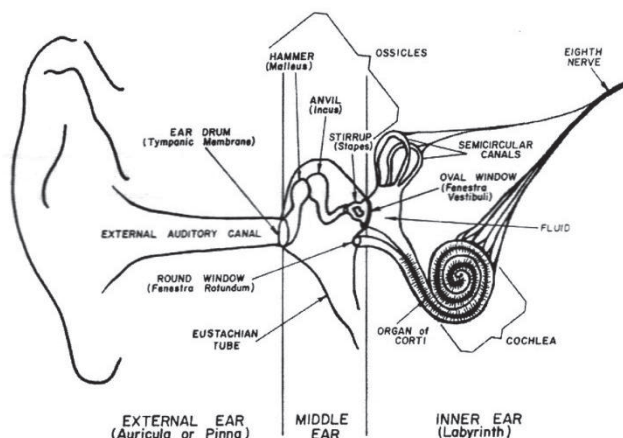


Figure 1: Anatomy of the Ear

The question is now: how much noise communities will tolerate or at what point the citizenry will have reached its threshold has never had greater currency. The number of towns enacting strict and enforceable ordinances to reduce and control noise levels both day and night suggests the point have been reached.

Barking dogs, lawn mowers, leaf blowers, power saws, motorcycles, airplanes, car stereo systems and traffic generally have combined to such a degree that noise induced irritations, annoyance, discomfort and hearing impairment have become a significant public health issue, certainly enough of one to motivate a political response.

Decibels	Traffic	Aircraft	Construction	Rail	Industrial
More than 80	0.1	0.1			
More than 75	1.1	0.3	0.1		
More than 70	5.7	01.3	0.6	0.8	
More than 65	19.3	4.7	2.1	2.5	0.3
More than 60	46.6	11.5	7.7	3.5	1.9
More than 55	96.8	24.3	27.5	6.0	6.9

Table 1: Population exposed to noise level and source (Million people exposed)

Source	Rank	% of Respondents Highly annoyed by source	Rank	% of Respondents Highly annoyed by source	Rank	% of Respondents Highly annoyed by source
Motorcycles	1	12.7	1	13.2	1	9.4
Automobiles	2	9.4	3	7.4	3	4.2
Large trucks	3	7.3	2	10.0	7	2.6
Construction	4	6.5	4	7.2	4	3.7
Sports cars	5	5.9	5	7.0	4	3.7
Constant traffic	6	4.7	6	5.5	10	1.5
Buses	7	4.7	8	3.5	11	1.1
Small trucks	8	4.1	7	4.1	9	1.5
Helicopter	9	3.9	10	3.1	2	5.3
Airplanes	10	3.6	9	3.4	5	3.2
Power garden tools	11	1.2	11	2.1	8	1.8
Total		66.0		62.2		55.9

Table 2: Population reporting “Highly annoying sources” population per square mile more than 20,000, 3,000 – 20,000 less than 3,000

Results and Discussion

Considering that earphone listening has been around for some 30 years, why has the problem only recently surfaced? Apparently the pattern of listening has changed. Currently, earphones are used while walking or running on noisy easy streets rather than in the privacy of the home or other relatively quiet area where the listener did not wish to disturb others. Now the volume must be tuned up to overcome the noise of city traffic.

Radio industry believes it is not their product that are the problem;; rather it is improper use. If they say, the volume is kept down, there would be no problem, which is equivalent

to saying that if we all drove cautiously there would be no accidents.

Since to these people louder are better the makings for an epidemic of hearing loss is at hand.

The occupational safety and Health Agency (OSHA) has set the danger level at 95 dB and above for 4 or more hours per day as likely to induce permanent hearing impairment. Although noise is an integral part of civilization it would appear that unless some definite steps are taken to reduce the present inordinate levels in both industry and community generally, more people will become auditory cripples.

Fortunately, in the past few years, large portions of the population have expressed an increasing determination to revolt against noise levels. A measure of this is seen in the number of communities moving forcefully to reduce noise levels in their home areas.

Studies of both traffic noise and noise communities near major airports have concluded that elevated blood pressure, heart disease and psychological trauma are direct consequences of noise exposure.

One example of psychological trauma is the research of Jenkins and his group at the London Institute of Psychiatry. It was reviewed the findings of two studies conducted in the area of London is Heathrow Airport. These studies had compared rates of admission at Springfield psychiatric hospitals among residents living near Heathrow, findings suggested that area closest to the airport with presumably higher levels of noise, also had the highest rates of hospital admission.

Sudden and unexpected noise has been observed to produce marked changes in the body, such as increased blood pressure, increased heart rate and muscular contractions. Moreover, digestion, stomach contractions and the flow of saliva and gastric juices all stop. Because the changes are so marked repeated exposure to unexpected noise should obviously be kept to a minimum. These changes fortunately wear off as a person becomes accustomed to the noise. However, even when a person is accustomed to an environment where the noise level is high, psychological changes occur.

Noise has psychological effects is undoubted. The question is how these effects can be assessed and whether they lead to damage. No clear case has been made thus far for psychological damage caused by moderately high levels of noise, that would cause hearing damage to only a small fraction of the people exposed. Indeed, fears have been expressed that “..... over emphasis on damage may backfire when people came to realize that the truth of the matter

seems to be simply that people can express violently their dislike about being disturbed by noise.

Future Aspects

Even with the relatively ambitious steps currently being taken or envisioned to control noise in most countries, sound levels and exposure to noise will remain high and possibly increase. At the same time rising living standards will bring about demands for better environmental quality and more vigorous and more organized protests against noise. These protests may even be triggered by lower noise levels than in the past, for it is highly likely that as the public acquires more amenities, it will want to be exposed to “comfortable” rather than merely tolerable levels of sound.

According to a WHO report to the UN conference on environment of all environmental problems, noise is the easiest to control”. But the question of control will arise only after there is awareness among the people of the need for control and for the government to find some solution for it.

1. The first approach has been to reduce noise at source: Design and fabrication of silencing engines, trucks, cars, motorcycles, industrial machines and home appliances would be an effective measure; protection to workers can be provided through wearing devices such as earplugs and ear muffs.
2. Making a change in design and operation of machines, vibration control, sound proof cabins and sound absorbing materials can reduce it.
3. It can get reduced by prescribing noise limits for vehicular traffic, ban on honking of horns in certain areas and planning main traffic arteries, industrial establishment, amusement areas, residential colonies, creation of silent zones near schools and hospitals and resigning of building to make them noise proof. Other measure can involve reduction of traffic density in residential areas giving preferences to mass public transport system.
4. **Control of indoor noise:** Following method can be applied for reducing the effect.
 - (a) Locate in the building as far as possible from noise source. The noise level drops about 6 dB each time the distance is doubled.
 - (b) Trees and shrubs may be planted in form of building to provide some absorption for the sound.
 - (c) Back to back bathrooms or toilets should be avoided unless they are effectively sound isolated. Bathrooms, kitchen should not be adjacent to the floor.

(d) Bathroom walls, floor and ceiling should be sound insulated using construction of high sound insulation glasses.

1. **Road noise:** Vegetation buffer zones must be created in different parts of the city. Efforts should be made for roadside plantations.
2. **An urgent need for legislation to control noise pollution:** We have seen that in India, in advances of a specific legislation for control and prevention of the noise pollution, one has to seek provisions in various branches of law and regulations. There has been no doubt that the available provisions in various branches of law are adequate, unscientific and crude. In most of the developed countries specific legislations have been made and scientific methods for investigations of noise pollution have been invented. The science of audiometer and other branches related to sound have been developed and it becomes comfortable to device various legal provisions to control and prevent noise pollution.

At present, there is no specific and detailed legislation to control the noise pollution. However, there is an urgent need that the central? Government of India should manage to get a legislation passed for the control of noise pollution. Some legislation regarding water and air pollution has been made in India.

Government should pass the noise pollution control act to meet special India condition. Apart from such kind of central legislation, there should be a city noise control code for all major cities in India. Creation of unnecessary noise has to be prohibited and should be punishable under law.

3. **Education:** people can be educated through radio TV, newsreels in cinema halls about noise pollution. In the family, elders can teach children to keep the radio and TV volume low, low voice talking, not to horn unnecessarily on the roads and avoid quarrelling amongst each other's and so on. There should be complete ban of loudspeakers from 8 p.m. to 7 a. m. mornings.

4. **Public Awakening and the control:** it is also important that public awakening is also very essential for the control and preventions of the noise pollution. In India, most persons lack any idea about the ways in which pollution could be controlled. Very few scientists are aware of the problem and its control. Masses are still ignorant of the grave effects of the noise pollution. In this regards television, radio, internet and newspapers should give a campaign for wide publicity.
5. There standard or "ideas" may be easily controlled of all environmental problems, but just as every other problem, do we want to change for the sake of noise pollution? If the question is answered with integrity and honestly, it is sure that not many would want to change over some small portions of noise pollution. With this exception of the throwing the blame at someone else makes it so much easy to control noise, since it is not the life of yourself being at stake to change.

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Cellular Mobile Communication System with Application to Smart Antenna

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Abstract

In view of explosive growth in the number of digital cellular subscribers, service providers are becoming increasingly concerned with the limited capacities of their existing networks. This concern has led to the deployment of smart antenna systems throughout major metropolitan cellular markets. These smart antenna systems have typically employed multibeam technologies, which have been shown, through extensive analysis, simulation, and experimentation, to provide substantial performance improvements in FDMA, TDMA and CDMA networks. Multibeam architectures for FDMA and TDMA systems provide the straight-forward ability of the smart antenna to be implemented as a non-invasive add-on or appliqué to an existing cell site, without major modifications or special interfaces.

Keywords: Magnetic nanoparticles, Biomedical applications, low toxicity.

Introduction

The main impediments to high-performance wireless communications are interference from other users (co-channel interference), the inter-symbol interference (ISI)[1-4] and signal fading caused by multipath. Co-channel interference limits the system capacity, defined as the number of users which can be serviced by the system. However, since the desired signal and co-channel interference typically arrive at the receiver from different directions, smart antennas can exploit these differences to reduce co-channel interference, thereby increasing system capacity. The reflected multipath components of the transmitted signal also arrive at the receiver from different directions, and spatial processing can use these differences to attenuate the multipath, thereby reducing ISI and fading.[5-9] Since data rate and BER are degraded by these multipath effects, reduction in multipath through spatial processing can lead to higher data rates and better BER performance.

In a cellular system, omni-directional antennas have traditionally been used at base stations to enhance the coverage area of the base stations but it also leads a gross wastage of power that in fact is the main cause of co-channel interference at neighboring base stations. The sectoring concept with diversity system exploits space diversity and results in improve reception by counteracting with negative effects of multipath fading. Adaptive / smart antenna technology represents the most advanced smart antenna approach to date. Using a variety of new signal-processing algorithms, the adaptive system takes advantage of its ability to effectively locate and track various types of signals to dynamically minimize interference and maximize intended

signal reception. Both adaptive / smart systems attempt to increase gain according to the location of the user; however; only the adaptive system provides optimal gain while simultaneously identifying, tracking, and minimizing interfering signals.

Smart Antenna

Smart

The concept of using multiple antennas and innovative signal processing to serve cells more intelligently has existed for many years. In fact, varying degrees of relatively costly smart antenna systems have already been applied in defense systems. Until recent years, cost barriers have prevented their use in commercial systems. The advent of powerful low-cost digital signal processors (DSPs), general-purpose processors (and ASICs), as well as innovative software-based signal-processing techniques (algorithms) have made intelligent antennas practical for cellular communications systems. Today, when spectrally efficient solutions are increasingly a business imperative, these systems are providing greater coverage area for each cell site, higher rejection of interference, and substantial capacity improvements.

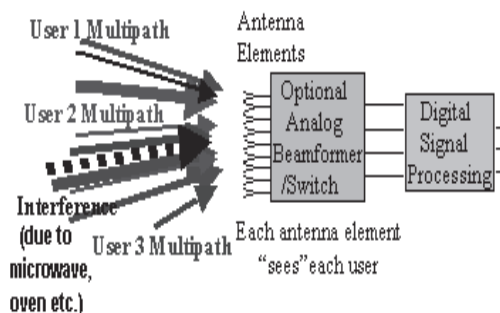


Fig. 1 : Smart Antenna System

Smart Antenna System

In truth, antennas are not smart antenna systems are smart. Generally co-located with a base station, a smart antenna system combines an antenna array with a digital signal-processing capability to transmit and receive in an adaptive, spatially sensitive manner. Such a configuration dramatically enhances the capacity of a wireless link through a combination of diversity gain, array gain and interference suppression. Increased capacity translates to higher data rates for a given number of users or more users for a given data rate per user. In other words, such a system can automatically change the directionality of its radiation patterns in response to its signal environment. This can dramatically increase the performance characteristics (such as capacity) of a wireless system.

Multipath of propagation are created by reflections and scattering. Also, interference signals such as that produced by the microwave oven in the picture fig (1) are superimposed on the desired signals. Measurements suggest that each path is really a bundle or cluster of paths, resulting from surface roughness or irregularities. The random gain of the bundle is called multipath fading [7], [8], [9], [10].

The Goals of a Smart Antenna System

The dual purpose of a smart antenna system is to augment the signal quality of the radio-based system through more focused transmission of radio signals while enhancing capacity through increased frequency reuse.

Feature	Benefit
Signal Gain—Inputs from multiple antennas are combined to optimize available power required to establish given level of coverage.	Better Range / Coverage—Focusing the energy sent out into the cell increases base station range and coverage. Lower power requirements also enable a greater battery life and smaller/lighter handset size.
Interference Rejection—Antenna pattern can be generated toward co-channel interference sources, improving the signal-to-interference ratio of the received signals.	Increased Capacity—Precise control of signal nulls quality and mitigation of interference combine to frequency reuse reduce distance (or cluster size), improving capacity. Certain adaptive technologies (such as space division multiple access) support the reuse of frequencies within the same cell.
Spatial Diversity Composite information from the array is used to minimize fading and other undesirable effects of multipath propagation.	Multipath Rejection can reduce the effective delay spread of the channel, allowing higher bit rates to be supported without the use of an equalizer, improved bit error rate (due to decreased amount of multipath and ISI).
SDMA- SDMA continually adapts to the radio environment through intelligent / smart antenna.	Providing each user with uplink and downlink signals of the highest possible quality and can adapt the frequency allocation to where the most users are located.
Power Efficiency combines the inputs to multiple elements to optimize available processing gain in the downlink (toward the user)	Reduced Expense Lower amplifier costs, power consumption, and higher reliability will result. Lower power consumption reduces not only interferences but also reduces RF pollution (ease health hazard). It will also result in reduction of scares energy resource (diesel consumption) and save foreign currency.

Table 1 : Features and Benefits of Smart Antenn Systems

The Architecture of Smart Antenna Systems

Traditional switched beam and adaptive array systems enable a base station to customize the beams they generate for each remote user effectively by means of internal feedback control. Generally speaking, each approach forms a main lobe toward individual users and attempts to reject interference or noise from outside of the main lobe.

Switched Beam Systems

In terms of radiation patterns, switched beam is an extension of the current microcellular or cellular sectorization method

of splitting a typical cell. The switched beam approach further sub-divides macrosectors into several microsectors as a means of improving range and capacity. Each microsector contains a predetermined fixed beam pattern with the greatest sensitivity located in the centre of the beam and less sensitivity elsewhere. The design of such systems involves high-gain, narrow azimuthal beam width antenna elements (fig 2).

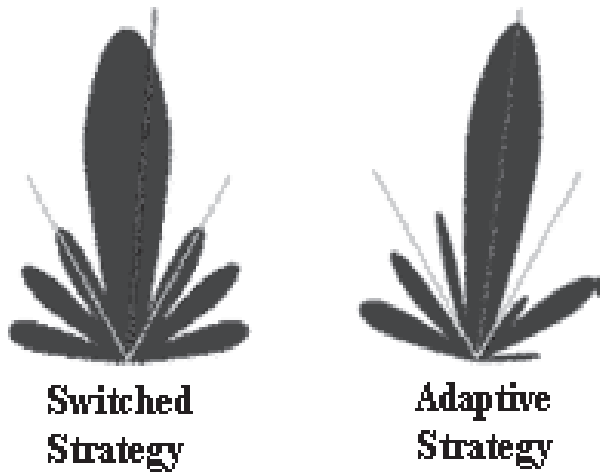


Fig. 2 : Beam forming Lobes and Nulls that Switched Beam and Adaptive Array Systems

Smart antenna systems communicate directionally by forming specific antenna beam patterns. When a smart antenna directs its main lobe with enhanced gain in the direction of the user, it naturally forms side lobes and nulls or areas of medium and minimal gain respectively in directions away from the main lobe. Different switched beam and adaptive smart antenna systems control the lobes and the nulls with varying degrees of accuracy and flexibility.

Conclusion

In conclusion to this paper "Smart Antenna" systems are the antennas with intelligence and the radiation pattern can be varied without being mechanically changed. With appropriate adaptive algorithms such as Recursive Least Square Algorithm (RLS) the beam forming can be obtained. As the system uses a DSP processor the signals can be processed digitally and the performance with a high data rate transmission and good reduction of mutual signal interference.

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Statistical Study of the Ultrasonic Attenuation and Critical Fluctuations in a Critical Binary Polymeric Mixture

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Abstract

In this paper we report the experimental studies of the ultrasonic attenuations in a critical binary polymeric mixture namely (Polypropylene glycol (M= 1000 gm / mol) and Polyethylene glycol (M= 400 gm / mol). The critical mixture is investigated by light scattering and ultrasonic experiments in the homogenous one – phase region. The ultrasonic measurements are interpreted by a dynamic scaling theory for low molecular weight binary critical mixtures of Bhattacharjee and Ferrell. The characteristic time scale of the dynamics of the critical concentration fluctuations is described by the critical frequency. The low mutual diffusion coefficient of the mixture allows us to study the high frequency behaviour of critical ultrasonic attenuation in the reduced frequency range. We employ dynamical scaling theory in this study. The light data are mainly intended to get the critical amplitudes of the correlation length (ξ) and diffusion coefficient (D).

Keywords: Critical concentration fluctuations, mutual diffusion coefficient critical exponent, critical amplitude, ultrasonic behaviour and critical dynamics

Introduction

In recent years the study of ultrasonic and dispersion has been an important tool in the investigations of dynamics of relaxations processes in theoretical as well as experimental realm. For the experimental purpose, the coupling of the sound wave with critical order parameter fluctuations will be required. The alternate adiabatic comparisons are expansion of the fluids allows change in pressure and critical temperature of the system. The energy dissipation will take place due to lagged response of the concentration fluctuations. A coupling constant g (Bhattacharjee and Ferrell) is introduced given by

$$g = \rho C_p \left[\frac{dT_c}{dP} - \left(\frac{\partial T}{\partial P} \right)_S \right] \quad \dots\dots\dots (1)$$

In relation 1, C_p is Heat capacity of the critical mixture at constant pressure, T_c is critical temperature, P is pressure, S is Entropy and ρ is density of the mixture.

In the present consideration, the deserved critical excess ultrasonic attenuation is chiefly attributed to the pressure dependence of the critical temperature of demixing is the mixture 2 – butoxy ethanol / water. The time scale of the decay of critical concentration fluctuations is characterised by a frequency $\omega_c = \frac{2D}{\xi^2}$ where D is mutual diffusion coefficient and ξ is correlation length of critical concentration fluctuation. Here we investigate the problem in

the following ways When polymeric compounds attenuate the sound wave by Rorse – Zimm – like modes.

When we study the high frequency behaviour ($\omega > \omega_c$) due to low mutual diffusion coefficients of polymeric mixtures.

In this regime the theories seems to be very sensitive to the incorporated mode coupling terms. The time scale of the critical dynamics of binary mixtures will be characterized experimentally by the critical amplitude ω_0 of ω_c . Where

$$\omega_c = \omega_0 \epsilon^v \quad \dots\dots\dots (2)$$

With $\epsilon = \left(\frac{T - T_c}{T_c} \right)$ as reduced temperature, ω_0 is critical amplitude and v is critical exponent. ω_0 lies in the range 1 – 50 GHz for low molecular components. Presently we study the ultrasonic attenuation in a critical binary polymeric mixture. The high frequency behaviour of critical ultrasound attenuation will be specially considered here. In order to obtain the critical amplitudes of the correlation length and diffusion co – efficient using light scattering data; conclusion will be drawn from the ultrasonic measurements.

Discussion of the Experiments related to Ultrasonic Measurement and Light Scattering Measurements

The composition of the critical mixture is determined by measuring the volume ratios of the separated phases in the inhomogeneous two – phase region (criterion of equal volumes). The samples employed are commercial products by Merck (Darmstadt). Their molecular weight distribution is

characterized by MALDI mass spectroscopy. For polypropylene glycol, $M_w = 1000$ gm / mol, $M_u = 970$ gm / mol, $M_w / M_u = 1.03$, for polyethylene glycol; $M_w = 406$ gm / mol, $M_u = 384$ gm / mol so $M_w / M_u = 1.06$.

Before mixing by weight, the probes were dried for several days in vacuum at 50°C . (i)

Light scattering experiments are performed with a commercial photometer (ALV, Langen) with He – Ne laser light source having 35 Mw laser power. The goniometer covers an angular range $30^\circ < \theta < 140^\circ$, where θ is scattering angle. The dynamic light scattering experiments were performed with a correlation model ALV 50 W / E. an external thermostat controls temperature with accuracy ± 0.05 k. static and dynamic light scattering experiments were carried out in the temperature range $0.07 \text{ k} \leq (T - T_c) \leq 1.42$ k. in this experiments the transmitted light is measured by a photodiode.

A resonator method (frequency range $0.4 \text{ MHz} < f < 1.5 \text{ MHz}$) is employed to measure the ultrasonic absorption and the same resonator cell by a pulse transmission technique (frequency range $6 \text{ MHz} < f < 30 \text{ MHz}$) is used similar to the method described by (Eggers and Funck). Thermostat controls the temperature with accuracy 0.02k . the ultrasonic cell is enclosed in a dry box filled with silica gel to avoid humidity. The temperature range for the said experiments range between 0.1 k to 21.3 k . the wide range of temperature is better for cell alignment.

Results of the performed Experiments

The plot of temperature in $^\circ\text{C}$ of the critical mixtures (Polyethylene glycol (PEG) and polypropylene glycol (PPG)) versus Y_{PPG} (Mass fraction of PPG in the mixture) is shown schematically in figure 1 with a dotted line because no sharp cloud points could be observed with the optical method. . we study the volume ratios of the phase separated samples to have the composition of the critical mixture. The used mixture in the experiments is characterised by the criteria of equal volume in the vicinity of its phase separation temperature. The experiments provide $Y_{\text{PPG}} = 0.46$ at phase separation temperature = $T_c = 30.5^\circ\text{C}$ (critical temperature).

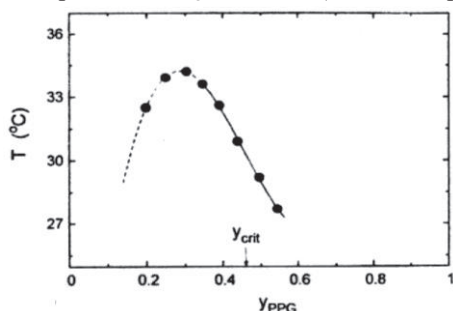


Figure1: Phase diagram of PEG and PPG

The study of this graph reveals that in contrast to binary mixture the critical mixture does not coincide with the mixture having the highest phase separation temperature. This result is justified by the polydispersity of the oligomeric compound.

Static Light Scattering

In the vicinity of the critical point of demixing, the scattered light intensity mainly results from fluctuations of the concentration. This enables us to determine structure function $S(\mathbf{k})$ by measuring the scattered intensity at various scattering angles. $S(\mathbf{k})$ is given by

$$S(\mathbf{k}) = \frac{I(\mathbf{k}) \cdot A^4 \cdot \bar{\rho} \cdot N_L \cdot R_{VV}^{ref}}{I_{ref} \cdot n_A^2 \cdot n_B^2 \cdot 4\pi^2 (n_A - n_B)^2 \cdot M_{mon}} \\ = A k_B \cdot T \cdot \chi(T) \cdot g(\mathbf{k}\xi) \quad \dots\dots\dots (3)$$

Where $I(\mathbf{k})$ == Scattered light intensity, $\bar{\rho}$ is mean density, n_A and n_B are the refractive indices of the components A and B. R_{VV}^{ref} is Rayleigh ratio of reference line, $\mathbf{k}$ = wave vector = $2\pi / \lambda$ and $c(\mathbf{k})$ is Fourier components of the concentration fluctuations.

Also $\mathbf{k} = \frac{4\pi n}{\lambda} \sin \frac{\theta}{2}$ where θ is scattering angle, M_{mon} is mean molecular mass of the moment units of A and B. $\chi(T)$ is osmotic compressibility at temp. T.

$\chi = \chi_0 e^{-v}$ Where χ is critical amplitude, v is critical exponent and $\varepsilon = (T - T_c) / T_c$ is reduced temperature.

$g(x)$ = Ornstein – Zernicke scaling factor = $1 / (1 + x^2)$

But according to Finner

$$g(x) = \frac{1}{(1 + x^2)^{1-\eta}} \text{ Where } \eta \text{ is critical exponents} = 0.037$$

nearly.

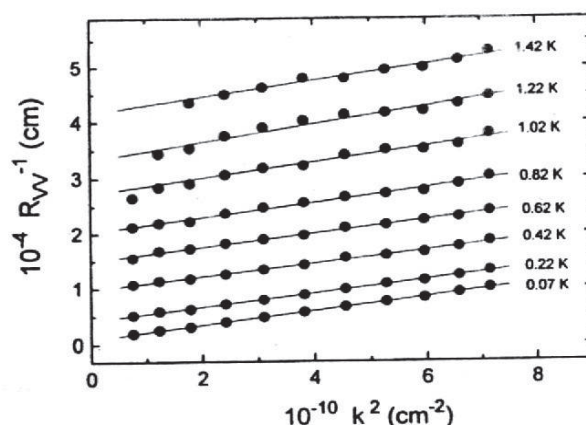


Figure 2: Plot of $1 / R_{VV}$ versus k^2 at different values of $(T - T_c)$ &

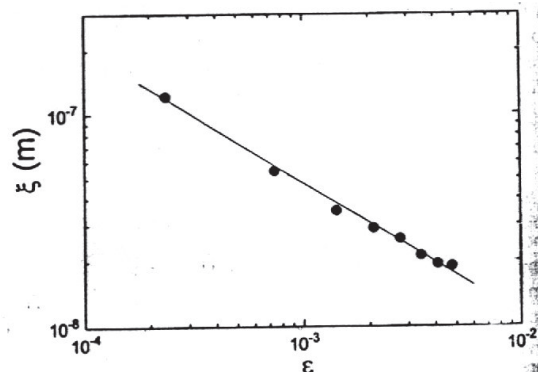


Figure 3: Plot of correlation length of critical concentration fluctuation ξ

Figure 2 shows a plot of the inverse Rayleigh ratio of the scattering of light intensity as a function of k^2 at different temperature. The data of this plot fit the Orustein - Zernicke form good. The slopes of the $1 / R_{vv}$ Vs k^2 curves are constant within the experimental error limit. We get $T_c = 23.33$ from the graph and hence the critical amplitude is $\xi_0 = (0.60 \pm 0.01)$ nm is evaluated. In contrast to light scattering cell the ultrasonic resonator could not be completely closed.

The decay of the order parameter is obtained by measuring $S(\mathbf{k}, t)$ of the scattered light intensity as

$$S(\mathbf{k}, t) = \langle c(\mathbf{k}, t) \exp(-i\mathbf{k} \cdot \mathbf{0}) \rangle \quad \text{..... (4)}$$

According to single exponential law, we have

$$S(\mathbf{k}, t) = S(\mathbf{k}) \exp[-\Gamma(\mathbf{k})t] \quad \text{..... (5)}$$

$$\text{Where } \Gamma \text{ is decay time} = D k^2 \quad \text{..... (6)}$$

$$\text{With } D = L / \chi(\mathbf{k}) \quad \text{..... (7)}$$

Where $\chi(\mathbf{k})$ is wave vector dependent Osmotic compressibility and L is mobility.

The decay of concentration fluctuation couples to the transverse local momentum fluctuations. As the decay rate contain a critical part along with a non - critical background.

L is expressed as $L = L_B + L_C$ so $D_M = D_B + D_C$ where $D_B = L_B / \chi(\mathbf{k})$ and $D_C = L_C(\mathbf{k}) / \chi(\mathbf{k})$ and D_C is given by

$$D_C = \frac{R(\mathbf{k}) \xi \cdot T}{6\pi\eta\xi} \times \Omega(\mathbf{k}\xi) \quad \text{..... (8)}$$

Where η is shear viscosity $= \eta_B \cdot (Q_0 \cdot \xi)^{x_\eta}$ where η_B is background viscosity, Q_0 is microscopic cut off wave number x_η is critical exponent, R is universal dynamic amplitude ratio.

$\Omega(x) = \Omega_k(x) [S(x)]^{x_\eta}$ Where Ω_k is Kawasaki function $= 1 + 3/5 x^2$ for $x \ll 1$.

Figure 4 expresses the mutual diffusion coefficient as a function of the reduced temperature ϵ .

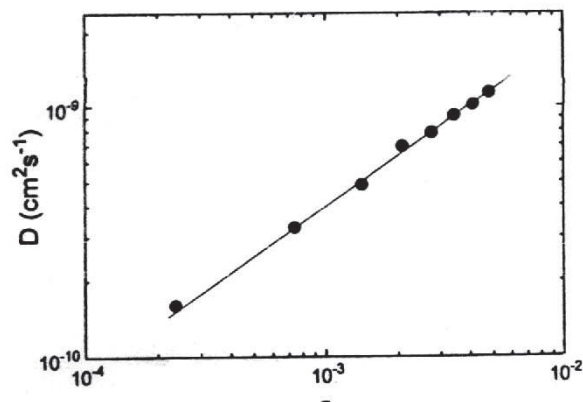


Figure 4: Plot of D versus ϵ

We can characterize the ultrasonic measurement by critical ultrasound theories for low molecular weight binary liquid mixtures. This is chiefly motivated by the light scattering results, which shows the presence of critical concentration fluctuations. The light scattering data is properly explained by common laws of static and dynamic critical phenomenon established earlier for binary liquid mixture. Critical ultrasound attenuation is founded in the lagged response of the critical concentration fluctuation to the pressure induced temperature variations.

Theoretical studies of critical ultrasound behaviour are based on calculation if the frequency dependent complex bulk viscosity or frequency dependent complex specific heat. In the present paper, only the comparison of the result of the dynamic scaling theory achieved are made with the experimentally observed results. This comparison is found very successful in low molecular weight components in the high frequency limit. These results resembles to the results of Kroll and Ruhland. In order to include the entire frequency and temperature range, a new parameter free relation for the reduced quantity was introduced by Bhattacharjee and Ferrell.

This quantity is $\frac{\alpha_\lambda}{\alpha_{\lambda_1} c}$ where α_λ is contribution to the attenuation coefficient per wavelength from critical order parameter fluctuation at a given temperature and $\alpha_{\lambda_1} c$ is contribution to the attenuation coefficient per wavelength from critical order parameter fluctuations at the critical temperature.

Conclusion

The measured excess ultrasonic attenuations of the critical mixture PEG 400 - PPG 1000 can be attributed to the presence of critical concentration fluctuations.

The achieved results reproduce the high frequency regime of theoretical predictions of Bhattacharjee and Ferrell and others.

Rouse – Zimm – like contribution of the pure components can be incorporated as an additive background term to the critical part of the ultrasound attenuations.]

The critical contribution of high molecular weight polymeric mixture would give a nearly temperature independent contribution to the observed attenuation.

On the basis of these observations we can proceed further to deal the case of noncritical mixture too.

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A Review on Knowledge Discovery in Databases (KDD) System

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Abstract

Reconfigurable systems use adaptive hardware to address the varying needs of different applications. The reconfigurable logic, generally a field-programmable gate array (FPGA), augments the functionality of a general-purpose processor (GPP). The current trend is to incorporate the reconfigurable logic fabric (RF) on the same die as the GPP, to alleviate the problem of communication overhead between the GPP and the RF. Reconfigurable hybrid processor systems provide a flexible platform for mapping data - parallel applications, while providing considerable speedup over software implementations. However, the overhead for reconfiguration presents a significant deterrent in mapping applications onto reconfigurable hardware. Presently, a very large amount of data stored in databases is increasing at a tremendous speed. This requires a need for new techniques and tools to aid humans in automatically and intelligently analyzing large data sets to acquire useful information. This growing need gives a view for a new research field called Knowledge Discovery in Databases (KDD) or Data Mining, which attract an attention from researchers in many different fields including database design, statistics, pattern recognition, machine learning, and data visualization. Data mining is the process of discovering insightful, interesting, and novel patterns, as well as descriptive, understandable and predictive models from large-scale data. In this paper we overviewed different tasks includes in Data mining.

Keywords: Reconfigurable system, field - programmable gate array (FPGA), data mapping, data visualization, data mining and data learning.

Introduction

Reconfigurable hybrid processor systems provide a flexible platform for mapping data-parallel applications, while providing considerable speedup over software implementations. However, the overhead for reconfiguration presents a significant deterrent in mapping applications onto reconfigurable hardware. Partial runtime reconfiguration is one approach to reduce the reconfiguration overhead.

Presently, a very large amount of data stored in databases is increasing at a tremendous speed. This requires a need for new techniques and tools to aid humans in automatically and intelligently analyzing large data sets to acquire useful information. This growing need gives a view for a new research field called Knowledge Discovery in Databases (KDD) or Data Mining, which attract an attention from researchers in many different fields including database design, statistics, pattern recognition, machine learning, and data visualization. Data mining is the process of discovering insightful, interesting, and novel patterns, as well as descriptive, understandable and predictive models from large-scale data.

The last decade has experienced a revolution in information availability and exchange of it through internet. In the same strength more business as well as organizations began to collect data related to their own operations, while the database technologists have been seeking efficient means of storing, retrieving and manipulating data, the machine learning community focused on techniques which used for developing, learning and acquiring knowledge from the data. Data Mining is the process of analysing data from different perspectives and summarizing it into useful information. Data mining consists of extract, transform, and load transaction data onto the data warehouse system, store and manage the data in a multidimensional database system, by using application software analyse the data, provide data access to business analysts and information technology professionals, present the data in a useful format, like a graph or table. Data mining involves the anomaly detection, association, classification, regression, rule learning, summarization and clustering.

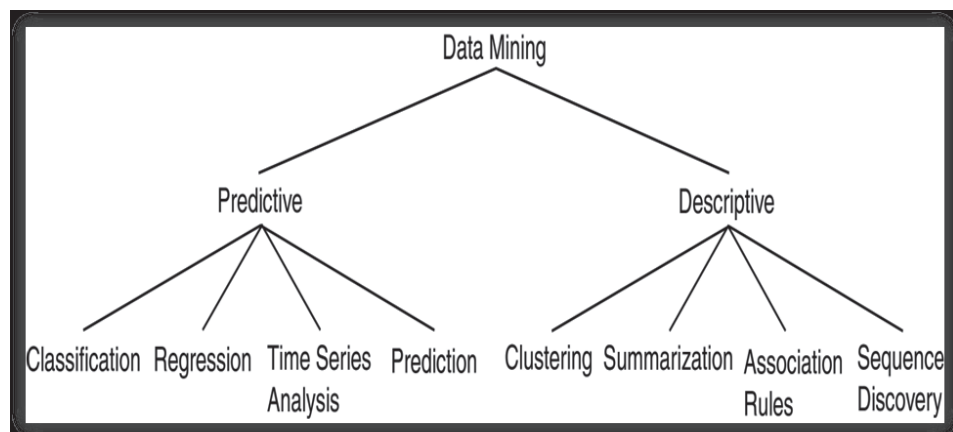


Figure 1(a): Data mining schematic diagram

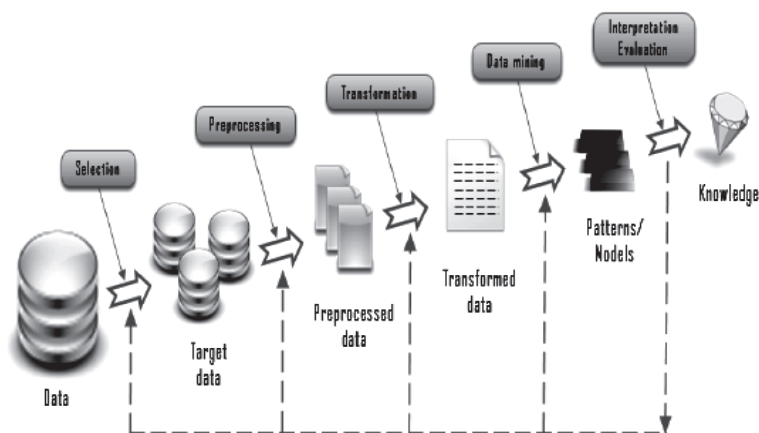


Figure 1(b): Design of data mining process

Data Mining

Data mining is the exploration and analysis of large data sets, in order to discover meaningful pattern and rules. The key idea is to find effective way to combine the computer's power to process the data with the human eye's ability to

detect patterns. The objective of data mining is to design and work efficiently with large data sets. Data mining is the component of wider process called knowledge discovery from database.

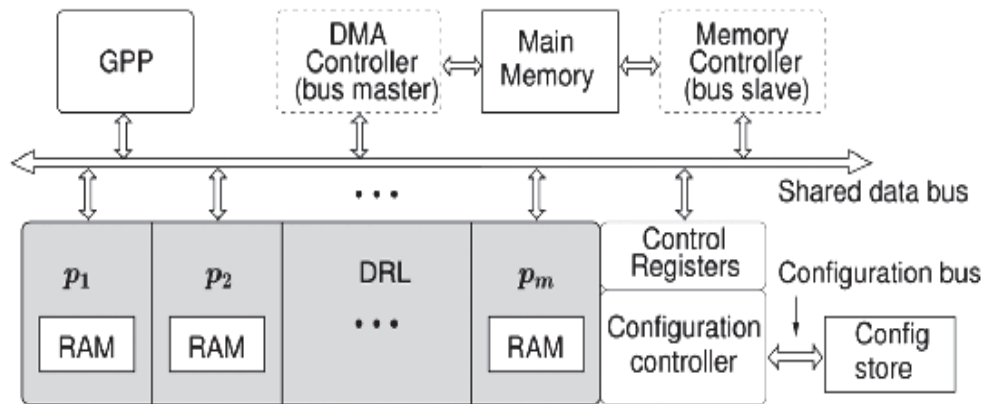


Figure 2: Process of Data Mining

Data Mining is the process of analysing data from different perspectives and summarizing the results as useful information (Figure 1(a) and 1(b)). It has been defined as "the nontrivial process of identifying valid, novel, potentially useful, and ultimately understandable patterns in data" The definition of data mining is closely related to another commonly used term knowledge discovery fig. (2). Data mining is an interdisciplinary, integrated database, artificial intelligence, machine learning, statistics, etc. Many areas of theory and technology in current era are databases, artificial intelligence, data mining and statistics is a study of three strong large technology pillars. Data mining is a multi - step process, requires accessing and preparing data for a mining the data, data mining algorithm, analysing results and taking appropriate action. The data, which is accessed, can be stored in one or more operational databases. In data mining the data can be mined by passing various processes.

In data mining the data is mined using two learning approaches i.e. supervised learning or unsupervised learning.

A. Supervised Learning

In supervised learning (often also called directed data mining) the variables under investigation can be split into two groups: explanatory variables and one (or more) dependent variables. The goal of the analysis is to specify a relationship between the dependent variable and explanatory variables as it is done in regression analysis. To proceed with directed data mining techniques the values of the dependent variable must be known for a sufficiently large part of the data set.

B. Unsupervised Learning

In unsupervised learning, all the variables are treated in same way; there is no distinction between dependent and explanatory variables. However, in contrast to the name undirected data mining, still there is some target to achieve. This target might be as data reduction as general or more specific like clustering. The dividing line between unsupervised learning and supervised learning is the same that distinguishes discriminate analysis from cluster analysis. Supervised learning requires, target variable should be well defined and that a sufficient number of its values are given. Unsupervised learning typically either the target variable has only been recorded for too small a number of cases or the target variable is unknown.

Data mining has evolved into an important and active area of research because of the theoretical challenges and practical applications associated with the problem of discovering interesting and previously unknown knowledge from real -

world databases. The main challenges to the data mining and the corresponding considerations in designing the algorithms are as follows:

1. Massive datasets and high dimensionality.
2. Over fitting and assessing the statistical significance.
3. Understandability of patterns.
4. Non - standard incomplete data and data integration.
5. Mixed changing and redundant data.

C. Data Mining Task

The three tasks - classification, estimation and prediction rules are examples of directed data mining or supervised learning. In directed data mining, the goal is to use the available data to build a model that describes one or more particular attribute(s) of interest in terms of the rest of the available attributes. The other three tasks – association rules, clustering and description are examples of undirected data mining i.e. no attribute is singled out as the target, the main goal is to establish some relationship among all attributes.

Classification consists of examining the features of a newly presented object and assigning to it a predefined class. The classification task is characterized by the well - defined classes, and a training set consisting of reclassified examples. The task is to build a model that can be applied to unclassified data in order to classify it. Estimation deals with continuously valued outcomes. Given some input data, we use estimation to come up with a value for some unknown continuous variables such as income, height or credit card balance. Any prediction can be thought of as classification or estimation. The difference is one of emphasis. When data mining is used to classify a phone line as primarily used for internet access or a credit card transaction as fraudulent, we do not expect to be able to go back later to see if the classification was correct. Our classification may be correct or incorrect, but the uncertainty is due to incomplete knowledge only: out in the real world, the relevant actions have already taken place. The phone is or is not used primarily to dial the local ISP. The credit card transaction is or is not fraudulent. With enough efforts, it is possible to check. Predictive tasks feel different because the records are classified according to some predicted future behaviour or estimated future value. With prediction, the only way to check the accuracy of the classification is to wait and see. Any of the techniques used for classification and estimation can be adopted for use in prediction by using training examples where the value of the variable to be predicted is already known, along with historical data for those examples. The historical data is used to build a model that explains the

current observed behaviour. When this model is applied to current inputs, the result is a prediction of future behaviour. An association rule is a rule which implies certain association relationships among a set of objects (such as “occur together” or “one implies the other”) in a database. Given a set of transactions, where each transaction is a set of literals (called items), an association rule is an expression of the form $X \rightarrow Y$, where X and Y are sets of items. The intuitive meaning of such a rule is that transactions of the database which contain X tend to contain Y . Cluster analysis can be used as a standalone data mining tool to gain insight into the data distribution, or as a pre-processing step for other data mining algorithms operating on the detected clusters. Many clustering algorithms have been developed and are categorized from several aspects such as partitioning methods, hierarchical methods, density-based methods, and grid-based methods. Further data set can be numeric or categorical. Clustering is the task of segmenting a diverse group into a number of similar subgroups or clusters. What distinguishes clustering from classification is that clustering does not rely on predefined classes. In clustering, there are no predefined classes. The records are grouped together on the basis of self similarity. Clustering is often done as a prelude to some other form of data mining or modelling. Data visualization is a powerful form of descriptive data mining. It is not always easy to come up with meaningful visualizations, but the right picture really can be worth a thousand association rules since the human beings are extremely practiced at extracting meaning from visual scenes. Knowledge discovery goals are defined by the intended use of the system. There are two types of goals: (1) verification and (2) discovery. With verification, the system is limited to verifying the user’s hypothesis. With discovery, the system autonomously finds new patterns. The discovery goal is further divided into prediction, where the system

finds patterns for predicting the future behaviour of some entities and description, where the system finds patterns for presentation to a user in human understandable form.

Conclusion

Data mining is an interdisciplinary, artificial intelligence, integrated database, machine learning, statistics, etc. Data mining involves extracting useful rules or interesting patterns from huge historical data. Many data mining tasks are available and each of them further has many techniques. . Data mining is a large number of incomplete, noisy, fuzzy, random application of the data found in hidden, regularity which are not known by people in advance, but is potentially useful and ultimately understandable information and knowledge of non-trivial process. In this paper we discuss some issues in Data Mining and activities used for Data mining task.

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Study on Switch Reduction Technique and Real - Time Optimization of the Pixelled Reconfigurable Antenna

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Abstract

Reconfigurable antennas are radiating structures that integrate an inner mechanism that allows the modification of its properties (resonant frequency, radiation diagram or polarization). Pixelled antennas are a particular type of reconfigurable antennas which are composed of a grid of electrically small metallic patches, where each pair of adjacent patches is interconnected by a RF - switch. The activation or deactivation of these switches modifies the antenna properties. Pixelled antennas can create a breakthrough in reconfigurable antennas field. Their increased reconfiguration capability offers a new set of possibilities but introduces fundamental differences due to the high number of necessary switches. This work focuses in three points: the study of techniques to minimize the number of switches, the development of monolithical integrated switches which are the more realistic approach in densely switched designs, and finally, the implementation of a real - time reconfiguration system for taking full advantage of the enhanced reconfiguration properties of pixelled antennas.

Keywords: Reconfigurable antenna, pixelled antenna, switches, monolithical integrated switches, mat lab and HFSS software

Introduction

An antenna is a part of a transmitting or receiving system which is designed to radiate or to receive electromagnetic waves. Which have been regarded as static elements considering their properties. However, during the last decade a new field of research has appeared under the name of antenna reconfigurability. A reconfigurable antenna is an antenna which integrates an inner mechanism that allows an intentional redistribution of the currents or equivalently, the electromagnetic fields of the antenna's effective aperture, resulting in reversible changes in the antenna impedance and/or radiation properties. Thus a reconfigurable antenna provides a trade - off in increased complexity for enhanced performance. From a practical perspective, the additional performance provided by a reconfigurable antenna should therefore offset the costs and complexity of reconfiguration.

There is a broad range of mechanisms for providing antenna reconfigurability employing basically mechanical, thermal, electrostatic or electric actuation. Among them we find the use of RF - switches, tunable electronic devices like varactors, structural modifications or material tuning. Reconfigurable antenna applications depend basically on the parameter under reconfiguration. It can be roughly said that impedance reconfigurability permits the antenna integration of different systems while radiation reconfigurability enhances the channel utilization. Impedance reconfigurability

is of particular interest in situations where several communication/sensing systems converge. In this case, instead of using a multiple of single - function legacy antennas, where each antenna is optimized for one of the different requirements, a single reconfigurable antenna can be employed. The behaviour of a reconfigurable antenna can be dynamically changed, thus a single antenna can accommodate multiple systems requirements. This is a very attractive approach for actual wireless devices which integrate multiple communication systems like GSM, UMTS, GPS, Bluetooth or WLAN. Radiation reconfigurability is conceptually different from the previous. The most classical application is beam - steering, which consists on modifying the direction of maximum radiation in order to maximize the gain of the antenna in a certain specific direction. A recent application of radiation reconfigurability are adaptive MIMO (Multiple Input Multiple Output) systems which result from the combination of reconfigurable antennas and Multiple Input Multiple Output (MIMO) concepts, that is to substitute the multiple static antennas of the standard MIMO systems by multiple reconfigurable antennas each one capable of synthesizing several different radiation patterns. The previous reconfiguration techniques are very appropriate in situations where a single specific parameter must be reconfigured and over reduced ranges. However the natural tendency of reconfigurable antennas is to evolve towards higher

reconfiguration capabilities, being able to reconfigure simultaneously impedance and radiation properties. In these situations the previous techniques present an insufficient reconfiguration capability and the development of more complex geometries is necessary. In these situations the high degree of redundancy of pixelled antennas can help to meet requirements where the other methods fail. Pixelation concepts were first applied in antenna's field as a design procedure completely independent of reconfiguration techniques. It basically consisted on the discretization of the antenna surface on electrically small rectangular patches, and by using an optimization algorithm it was determined whose rectangular must be filled with metal and whose should be left unfilled. By this procedure an optimized shape was developed according to the design criteria.

A reconfigurable pixelled antenna consists on a grid of electrically small metallic patches, where each pair of adjacent patches is interconnected by a RF - switch. The activation or deactivation of these switches modifies the current distribution over the antenna surface providing reconfigurability to the antenna structure. By conveniently choosing the status of each switch the properties of the antenna are modified in order to improve its performance. Pixelled antennas can create a breakthrough in reconfigurable antennas field. Their increased reconfiguration capability offers a new set of possibilities but introduces fundamental differences with common switched reconfigurability. The main difference is the total number of necessary switches. Common switched reconfigurable antennas require a low number of switches, usually between one and ten depending on the structure and the reconfiguration capabilities, while pixelled antennas usually require around a hundred switches to operate properly. These large amounts of switches have some negative impacts on the antenna performance. The efficiency of the antenna is significantly reduced due to the serial concatenation of lossy switches, the feasibility of the antenna is also drastically damaged due to the increased biasing circuitry complexity and the total cost of the antenna increases due to the addition of the large amount of switches. In this paper, we discussed switch reduction, switch integration and real - time optimization of the proposed structure.

Geometry of the Proposed Structure and Optimization Procedure

The overall performance of a pixelled antenna is highly dependent on the fundamental structures that compose it. Switches are the basic elements which provide reconfigurability to a pixelled antenna. The switch individual

specifications are directly translated to the antenna properties and for that reason switching technology is an essential element in a pixelled design. The main RF - switching structures are based either on semiconductor technology or electro - mechanical devices. The high number of necessary switches conforming pixelled antennas causes small degradations of the switch insertion losses to be translated in severe reductions of the antenna efficiency. This situation makes insertion loss the most significant parameter in switching technology for pixelled antennas. This large amount of switches also intensifies the importance of very low switch power consumption. These two requirements make RF - MEMS switch the most suitable candidate for pixelled antenna designs. A pixelled monopole antenna has been designed at 5.5 GHz which is depicted in Fig 1. .

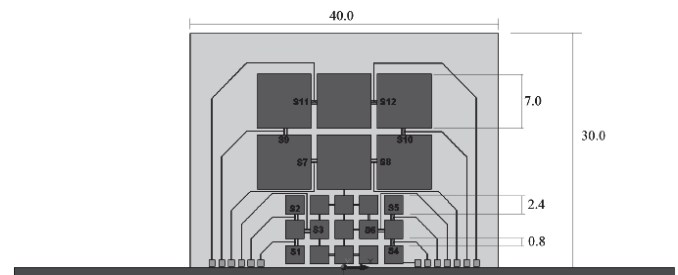


Figure 1: Pixelled monopole design

The overall size of the antenna is 30 mm by 40 mm (approximately $0.6\lambda - 0.7\lambda$, which are the dimensions of the substrate supporting the structure. The substrate material is synthetic quartz glass and its thickness is 1mm. To determine the position of the located switches, the contribution of each switch to the antenna reconfigurability has been analyzed by examining, each switch status variability when a fully switched model is optimized to synthesize different radiation patterns and resonant frequencies. Switches having a low variability in their status have been substituted by gaps or wires depending on their predominant status. This is the second switch minimization technique introduced in this design and it leads to only 12 switches which is a significant reduction. In fact, if the area occupied by the antenna would be filled with small patches (2.4 x 2.4 mm) and a switch was located between each pair of adjacent patches, the number of necessary switches would be larger than 80. Due to the low number of switches in the resulting design, it is possible to fit all the bias lines in the same side of the substrate which avoids the necessity of making vias in the quartz substrate. The whole structure is interconnected with high resistance lines which set the same ground potential over the structure

and thus only 1 bias line is required to control the status of each switch.

The optimization procedure is the sequence of automatic steps for the dynamically adjustment of the switch status configuration in order to maximize the performance of the antenna, which is quantified by the fitness function. The optimization parametric space corresponds to the switch status configuration and thus only discrete optimization algorithms can be used. The main discrete optimization algorithms are Random Walk, Simulated Annealing (SA), Genetic Algorithms (GA), Particle Swarm Optimization (PSO) and Ant Colony Optimization (Ant - Colony Optimization (ACO)). A convergence example of the main methods is presented in Fig. 2. The performance of each method has a strong dependence on the parameters geometry. One of the methods which has demonstrated a greater performance in electromagnetic optimization are genetic algorithms and one of their several variants is the optimization algorithm applied in this case.

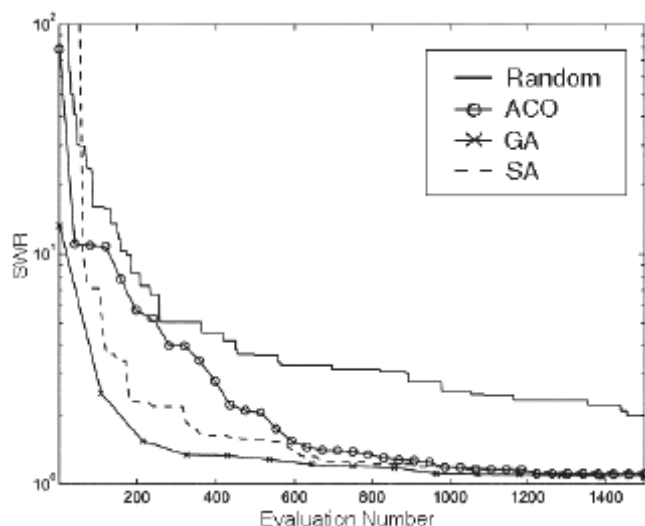


Figure 2: Convergence example of optimization algorithms

The genetic algorithm has been implemented by Mat lab using a homemade code developed. The fitness computation involves some parameters of the antenna, and for that reason requires the simulation of the structure. This electromagnetic simulation has been done using the software HFSS,

Result and Discussion

The presented design has been optimized to resonate at several frequencies and to synthesize several radiation diagrams by using the corresponding fitness functions described. The reconfiguration parameters and the resulting switch configurations is shown in Table 1.

	Resonant Frequency			Radiation Pattern			
	4GHz	5 GHz	6GHz	=60°	-30°	30°	60
S ₁	0	1	0	0	1	0	0
S ₂	0	0	1	0	0	1	1
S ₃	0	1	0	1	0	0	0
S ₄	0	0	0	0	0	1	0
S ₅	1	0	1	0	1	0	1
S ₆	1	0	0	1	1	1	1
S ₇	1	1	1	0	1	0	0
S ₈	1	1	0	0	0	0	1
S ₉	1	1	1	1	1	1	1
S ₁₀	1	0	1	0	1	1	1
S ₁₁	1	0	1	0	0	1	1
S ₁₂	0	1	1	1	1	0	0

Table 1: Switch configuration

The reconfigurable pixelled monopole has been designed to operate at 5.5 GHz. The antenna has been optimized for three different resonant frequencies: 4 GHz, 5 GHz and 6 GHz. The simulated return loss curves are presented in Fig. 3 showing that the antenna is able to use reconfigurability to cover from 3GHz to more than 6GHz.

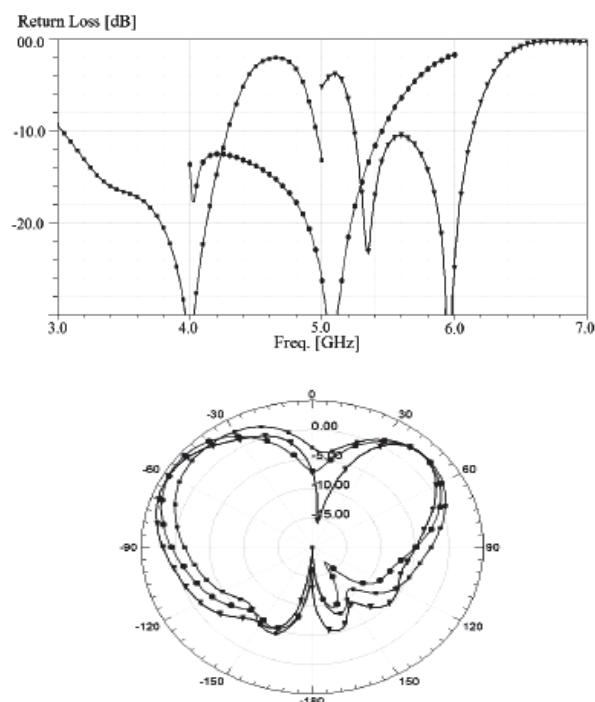


Figure 4(a): Return loss1 **Fig. 4(b) :** Radiation pattern

The radiation pattern of the configurations is presented in fig. 4. Although the radiation diagram has not been taken into account in the fitness function, the three configurations have

a similar monopole like pattern. The reason behind is that due to the antenna monopole structure, configurations having a monopole - like behaviour can achieve matching more easily than other configurations. The pixelled monopole has been optimized to maximize gain in -60° , -30° , 30° and 60° directions. Since the antenna has a planar geometry, directions of arrival (DOA) have been chosen over the antenna plane, otherwise the synthesis of a high gain beam out of this plane would generate a beam towards the symmetric direction that would strongly reduce the maximum directivity of the antenna. In fact, this DOA choice has an extra advantage related with polarization. While the antenna polarization is in general elliptical, due to the planarity of the structure, the polarization over the selected plane is always linear and no extra losses due to polarization mismatch must be considered.

Conclusion

This paper has been focused in the development of reconfigurable pixelled antenna, developing techniques and systems for taking advantage of their favorable characteristics (increased reconfiguration capability) and mitigate their drawbacks (high number of necessary switches). A novel switch number reduction technique has been presented and experimentally validated. This technique has been based in the introduction of a hybrid pixelled geometry composed of multiple sized patches, which makes a more efficient use of the located switches than in common uniform pixelled antennas. A real - time reconfiguration system based in the antenna performance sensing has been implemented, exhibiting excellent performance.

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Study of Properties and Morphology of a Ternary Semiconductor Alloy

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Abstract

Ternary and quaternary semiconductors provide a natural way of tuning the desired band gap and flexibility to control other material parameters, by changing the relative composition of the pure elements in the alloy. In this paper we studied growth and properties of quaternary alloy magnetic semiconductor (InGaMn)As grown both on GaAs substrates and on InP substrates by low-temperature molecular - beam epitaxy (LT - MBE). (InGaMn)As thin films shows ferromagnetism below ~30 K, exhibiting strong magneto - optical effect.

Keywords: Ternary semiconductor, quaternary semiconductor, band gap, alloy, magneto – optical effect and substrate.

Introduction

In recent years, there is much interest in the characterization of GaAs crystal because of its potential optoelectronics applications in the various fields. For instance, GaAs has caused widespread concern as a semiconductor saturable absorber for its advantages of stable photochemical property, good thermal conductivity, non - degradability and high damage threshold. GaAs - based solar cells also have many advantages, such as high photoelectric conversion efficiency, fine radiation resistance and good performance at high temperatures. As science and technology develops, people always expect to broaden the applications of GaAs. Doping other elements is one of the most commonly used methods, and is performed by replacing a small amount of anion species in GaAs with isovalent impurities, such as N, P, Sb and Bi. The ternary alloy GaAs_{1-x}Bix has been successfully grown by metal organic chemical vapour deposition (MOCVD) and quite recently by molecular beam epitaxy (MBE). Existing research shows that GaAs_{1-x}Bix alloy are strong potential candidate materials for long wavelengths emitters and detectors as well as spintronic - related devices.

Theoretical Studies

On the theoretical side, the calculated electronic properties of GaAs_{1-x}Bix show that Bi doping can reduce the band gap of GaAs, and the effect is more significant than alloying other III-V compound semiconductors at the same concentration. Recent research of the GaAsBi system is mainly concerned with the lattice constant, alloy formation energy, energy level position and composition evolution of the system, however, the optical properties have not yet been reported in detail, and the relationship between optical properties and electronic

structure are scarcely investigated. It is important to study theoretically the electronic structure and optical properties of GaAs_{1-x}Bix alloy, which can be applied in more tech areas. For instance, it can make GaAs_{1-x}Bix a promising new semiconductor saturable absorber that has a very high possibility to be used in all-solid-state Q-switched and mode-locked lasers. A variety of II-VI and III-V based magnetic semiconductors, such as (CdMn)Te,¹⁾ (InMn)As²⁾ and (GaMn)As,^{3,4)} have been extensively studied in the past years, because they have both properties of semiconductors and of magnetic materials and their spin - related functions could be used for novel “spintronic” devices. Among them, feasibility of growing magnetic quantum heterostructures based on (GaMn)As can give a new degree of freedom in the design of semiconductor materials and devices. One of the recent achievements are, for example, very high tunnelling magneto-resistance (TMR) ratios over 70% (maximum 75%) observed in (GaMn)As/AlAs/(GaMn)As heterostructures.⁷⁾ All of these studies mentioned above were done on ternary alloy magnetic semiconductors. In this paper, we study growth and properties of quaternary alloy magnetic semiconductor [(In_yGa_{1-y})_{1-x}Mnx]As which is based on In_yGa_{1-y}As. This quaternary alloy system has many potential advantages. First, the bandgap energy can be changed over a wide range, roughly from 0.4 eV to 1.4 eV, by changing the In content *y*. Especially, the bandgap of (InGaMn)As can be tuned at the wavelength of 1.3 or 1.55 μ m for optical communications, which cannot be realized by ternary alloy magnetic semiconductors. Another possibility is controllability of band structures and of easy magnetization axis (which is sensitive to the lattice strain) by changing the

In content. Therefore, introduction of quaternary alloy magnetic semiconductors will broaden the possibility of bandgap engineering using semiconductor alloys and heterostructures. In this paper, we present successful growth of ferromagnetic (InGaMn)As thin films by low-temperature MBE (LT-MBE).

Result and Discussion

In order to estimate the lattice constant a of (InGaMn)As, we have done X-ray diffraction measurements on 200 nm-thick $[(\text{In}_{0.53}\text{Ga}_{1-y})_{1-x}\text{Mn}_x]\text{As}$ ($y = 0.53$) thin film samples grown on InP substrates whose Mn concentration x was 0, 0.02, 0.04, 0.06 and 0.08. Here the In content y was fixed at 0.53, thus the films are nearly lattice matched to InP. The intrinsic lattice constant a of (InGaMn)As was calculated from the diffraction peak in X-ray spectra, assuming that (InGaMn)As was fully strained and that (InGaMn)As has the same elastic constant as InGaAs.

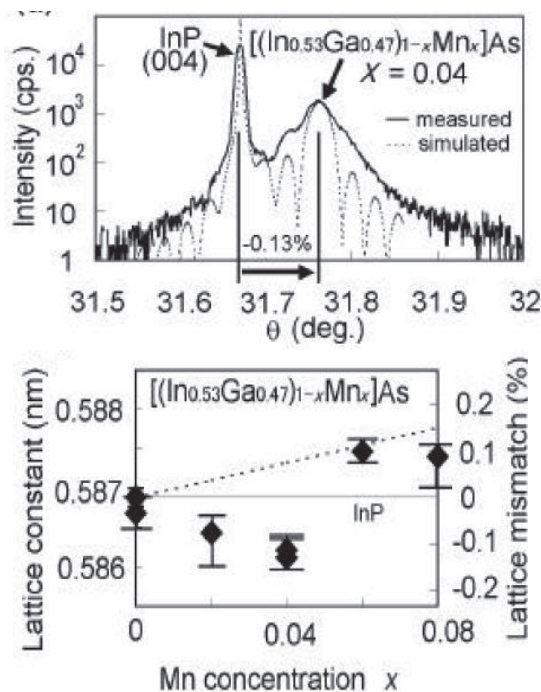


Figure 1: X-ray diffraction spectra of $[(\text{In}_{0.53}\text{Ga}_{1-y})_{1-x}\text{Mn}_x]\text{As}$, **Figure 2:** x dependence of the lattice constant a of (InGaMn)As

Figure 1 shows (a) X-ray diffraction spectra of $[(\text{In}_{0.53}\text{Ga}_{1-y})_{1-x}\text{Mn}_x]\text{As}$ in the case of $x = 0.04$, $y = 0.53$, and (b) x dependence of the lattice constant a of (InGaMn)As. The solid and broken lines in Fig. 1 are simulated X-ray spectra. The broken line in Fig. 2 indicates the lattice constant derived by Vegard's law, where we used the value of 0.598 nm as the lattice constant of hypothetical zinc-blende

MnAs.3). From Fig. 1, we can see that the experimental lattice constant slightly decreases with increasing x when $x = 0.04$. When x reached 0.06, the lattice constant became close to the value predicted by Vegard's law. The reason why the lattice constant does not follow Vegard's law is not clear at present. There is no indication of MnAs clustering in the RHEED and X-ray results of all the samples examined here. One practically important point is that the lattice mismatch of $[(\text{In}_{0.53}\text{Ga}_{0.47})_{1-x}\text{Mn}_x]\text{As}$ films to InP is very small, only $\pm 0.1\%$. In order to study the magnetic and magneto-optical properties, we have simulated magnetic circular dichroism (MCD) of $[(\text{In}_{0.53}\text{Ga}_{1-y})_{1-x}\text{Mn}_x]\text{As}$ films with the Mn concentration x fixed at 0.04 and the In content y varied from 0 to 0.53. All the samples were found to be ferromagnetic when prepared under appropriate conditions. In some (not all) as-grown samples, we have seen no ferromagnetic order. However, after annealing these samples at 280 - 300 °C for 10 minutes, ferromagnetic ordering was obtained. This improvement of magnetic property by annealing is due to the increase of hole concentration caused by the reduction of excess As incorporated during the non-equilibrium low-temperature MBE growth. Figure 4 shows MCD spectra of (a) $[(\text{In}_{0.53}\text{Ga}_{1-y})_{1-x}\text{Mn}_x]\text{As}$ ($x = 0.04$, $y = 0.26$) simulated at 8 K and of (b) (GaMn)As as a reference measured at 5 K, in a reflection set up with a magnetic field of 1 T applied perpendicular to the film. Both samples were grown on GaAs substrates. The result of $[(\text{In}_{0.53}\text{Ga}_{1-y})_{1-x}\text{Mn}_x]\text{As}$ ($x = 0.04$, $y = 0.53$) grown on InP was similar (not shown here). For (InGaMn)As, strong negative peaks were observed at 1.46 eV (near E_0) and at 2.7 - 2.8 eV (near E_1). Here, E_0 and E_1 correspond to the optical transitions at G and at critical points, respectively. The MCD intensity of (InGaMn)As was very strong, and comparable to or even stronger than that of (GaMn)As.¹⁰ The spectral features of (InGaMn)As were similar to those of (GaMn)As, indicating that the band structure of (InGaMn)As is of zinc-blende semiconductor type. It was seen from the spectra that E_0 and E_1 critical point energy values were shifted to lower energy due to the reduction of bandgap. For example, in Fig. 5 and 6, the largest negative peak energy coming from E_0 shifted from 1.62 eV to 1.46 eV. The amount of this peak energy shift in MCD was 0.16 eV, which is smaller than the E_0 shift ~ 0.36 eV expected from the bandgap reduction by adding the same content of In ($y = 0.26$) to GaAs. In MCD spectra a peak does not directly indicate critical point energy but a value of energy near the critical point. In this experiment, we simulated reflection MCD whose peaks do not exactly indicate critical points, and also the peak energy depends on the carrier concentration especially in the heavily doped

semiconductors such as (GaMn)As and (InGaMn)As. This red shift of the MCD peak, however, was observed in all the (InGaMn)As samples and tend to increase with increasing the In content. This result indicates that the bandgap energy can be changed in principle by changing the In content of (InGaMn)As films. We can see a clear hysteresis loop, indicating a ferromagnetic order up to 30 K. The reason why the Curie temperature of this (InGaMn)As sample is lower than those of (InMn)As and (GaMn)As is probably that the growth condition of (InGaMn)As is not fully optimized and there are many excess As in the (InGaMn)As lattice shown in figure 3, 4 and 5.

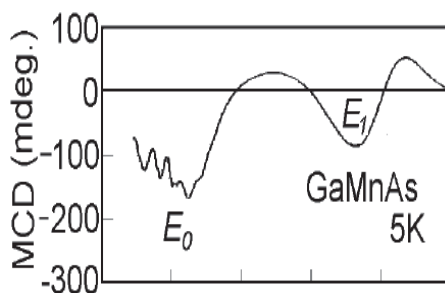
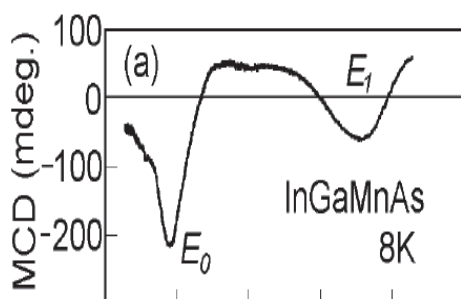


Figure 3: MCD spectra of (a)

$[(\text{In}_{0.26}\text{Ga}_{0.74})_{0.96}\text{Mn}_{0.04}]\text{As}$ grown on a GaAs at 8K,

Figure 4: (GaMn)As at 5 K

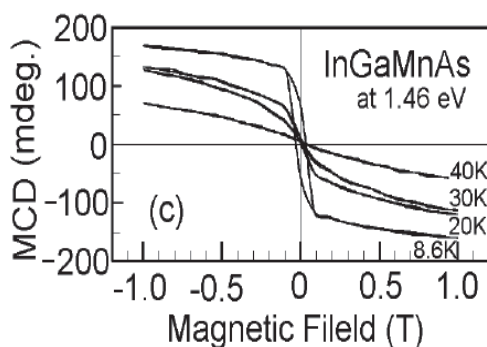


Figure 5: Magnetic field dependence of MCD on (InGaMn)As

We expect that higher Curie temperature will be obtained by increasing the growth temperature or by reducing the As pressure during the growth. It was also found that the easy magnetization axis of the (InGaMn)As film is perpendicular to the film plane. This is consistent with the fact that this $[(\text{In}_y\text{Ga}_{1-y})_{1-x}\text{Mn}_x]\text{As}$ ($y = 0.26$, $x = 0.04$) film receives tensile strain and its lattice constant is smaller than that of $\text{In}_y\text{Ga}_{1-y}\text{As}$ ($y = 0.26$), as in the case of $y = 0.53$ described earlier. Also this result indicates that the easy magnetization axis is controllable by changing the strain generated in the (InGaMn)As layer by changing the Mn concentration x or In content y .

Conclusion

Ferromagnetic $[(\text{In}_y\text{Ga}_{1-y})_{1-x}\text{Mn}_x]\text{As}$ quaternary alloy thin films were successfully grown by LT - MBE both on GaAs (001) substrates and on InP(001) substrates. The surface morphology of (InGaMn)As can be very smooth when nearly lattice matched films were grown on InP substrates. The lattice constants of $[(\text{In}_y\text{Ga}_{1-y})_{1-x}\text{Mn}_x]\text{As}$ thin films were slightly smaller (less than 0.2 %) than that of $\text{In}_y\text{Ga}_{1-y}\text{As}$ with the same In content y when the Mn concentration x is less than 0.04. MCD studies show that (InGaMn)As has zinc-blende type band structure, strong magneto - optical effect at E_0 and E_1 critical point energies, and perpendicular magnetization due to the tensile strain.

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Study of Bandwidth Enhanced Microstrip Patch Antenna with Metamaterials

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Abstract

In this paper, a metamaterial antenna based on Split Ring Resonator (SRR) and a Thin Wire (TW) is proposed and studied. The proposed antenna consists of six unit cells and a thin wire of Left Handed Materials (LHM) arranged in 2×3 printed on same side of substrate. The TW stub of the unit cell is used to connect the LHM's to the radiating edge of a microstrip patch antenna. The constitutive parameter of the LHM is retrieve to reveal the left handed region. By using these techniques, the bandwidth of the metamaterial antenna is twice that of the ordinary patch antenna at the same operating frequency.

Keywords: Thin wire, Microstrip patch antenna and Left handed materials.

Introduction

Metamaterials are artificially engineered materials with properties not readily found in naturally occurring materials. When the structural properties of a substance were manipulated, new substances were formed with unusual properties which could be desirable for a particular application. Therefore metamaterials are materials organised at structural level. Veselago in his early work put forward a theoretical speculation of materials having both negative electric permittivity and magnetic permeability negative in the same frequency band. Smith et al and Pendry in 2001 experimentally prove the work of Veselago. Metamaterials are characterized as having backward wave, reversal of Cherenkov radiation, negative velocity, negative refractive index and inverse Snell's law of refraction. This unusual electromagnetic phenomenon has attracted many researchers to turn their attention to this novel direction.

Among the numerous applications of metamaterials are perfect lens, wave retarder, absorber, cloak and antennas. Therefore in this paper, application of metamaterials for antenna design will be discuss. Several research works on antenna has been reported by many researchers with a great interest in the enhancement of the antennas properties such as bandwidth, gain, efficiency and small size.

In the related works, Circular Patch Antenna in the presence of a left handed medium is investigated, results shows that the antenna is more directive and has high gain where bandwidth is the major challenges, Microstrip Antenna's gain enhancement using metamaterial structure and patch antenna bandwidth enhancement through the use of metamaterials were studied and investigated. Result shows bandwidth enhancement of less than 5.8% was obtained.

LHM Structure

The LHM unit cell structure is shown in Figure 1. The structure composes of combination of rectangular split ring resonator (SRR) and thin wire (TW). The SRR and TW are printed on same side of FR4 substrate. The permittivity of the substrate is 4.4 and height h of 1.6mm. The length of the outer SRR is $L1=7.3\text{mm}$, and the inner SRR is $L2=5.4\text{mm}$, the width of the SRR and the thin wire is $W1=0.5\text{mm}$, the gap between the outer and inner ring is $W2=0.5\text{mm}$, the cut on both rings is $G1=1.6\text{mm}$ and $L3$ is the protrusion of the TW which connect to the patch.

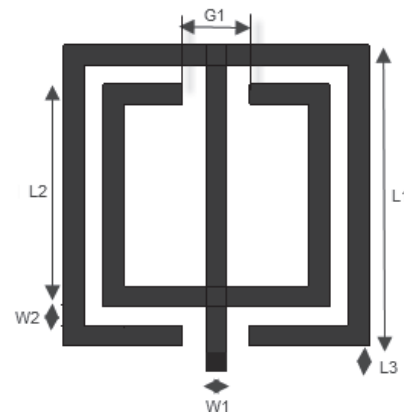


Fig 1: LHM Structure

Constitutive Parameters Retrieval

Usually the constitutive parameters of a unit cell are always retrieved using an algorithm to revealed the region where both permittivity and permeability negative. Nicolson Wier - Ross (NRW) algorithm as reported by used for the retrieval. The approach is to place a unit cell in a waveguide with boundary conditions set as shown in Fig 2. Perfect magnetic

condition (PMC) is place on z-axis, Perfect electric condition (PEC) on y- axis and waveguide ports on x-axis with open boundary conditions. Full electromagnetic Simulator software CST Microwave studio is used to obtain the return loss (S11) and the insertion loss (S21). The E - field of the incident wave is polarized along y-axis while the H - field of the incident wave is polarized along z-axis. The retrieved results from S - parameter is shown in Fig. 3

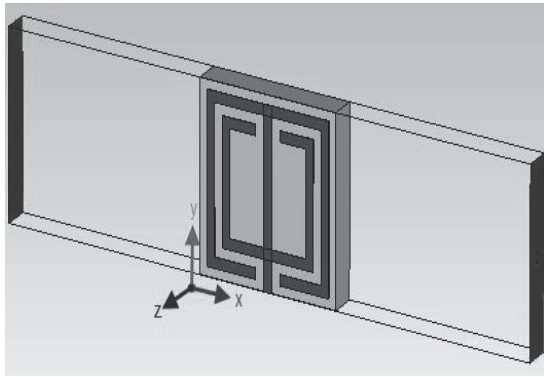


Figure 2: Unit cell in a waveguide

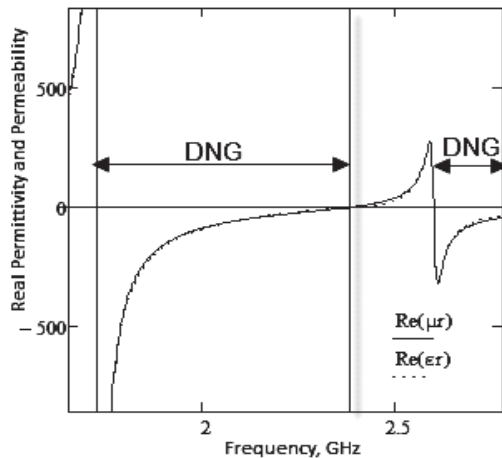


Figure 3: Retrieved effective parameters of the proposed LHM

Figure 3 shows the results of extracted constitutive parameters, two bands of double negative (DNG) region are revealed: 1.7- 2.44 GHz and 2.6- 3.2 GHz. This actually shows the left handed properties of the DNG structure.

Proposed Metamaterials Antenna Design

A microstrip patch antenna is design to resonate within the DNG region of the LHM structure. Here a centre frequency of 2.42 GHz for WLAN/ISM is chosen. The dielectric material used for both the LHM and antenna design is the FR4 substrate with the relative permittivity of 4.4, loss tangent of 0.02 and height of 1.6mm. Six unit cells of the LHM arranged in 2 by 3 were connected to the upper

radiating edge of the patch antenna. Also a slot is created on the ground plane of length $0.288\lambda_g$ and width $1.254\lambda_g$ which covers the LHM as shown in Fig. 6 while the proposed MTM antenna is shown is Fig.5. The designs parameters for the antenna are shown in Table 1.

Parameter	Dimension (mm)	Parameter	Dimension (mm)
L_P	29.2	L_{Sub}	62
w_P	38.2	$W_{Sub} = W_{slot}$	74
L_S	16.5	L_{slot}	17
W_S	3	L_n	15
y	6.1	L_m	22.9

Table 1: Antenna design parameter

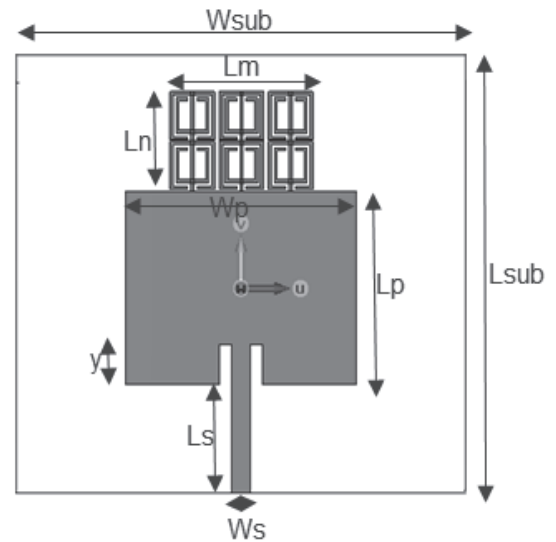


Figure 4: Patch Antenna with LHM (Front view)

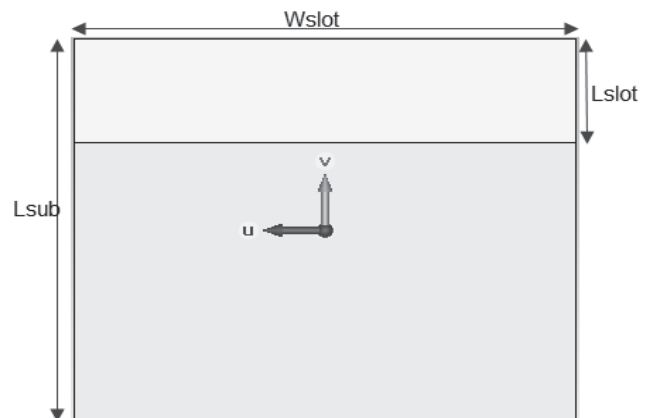


Figure 5: Slot on the ground plane (Back view)

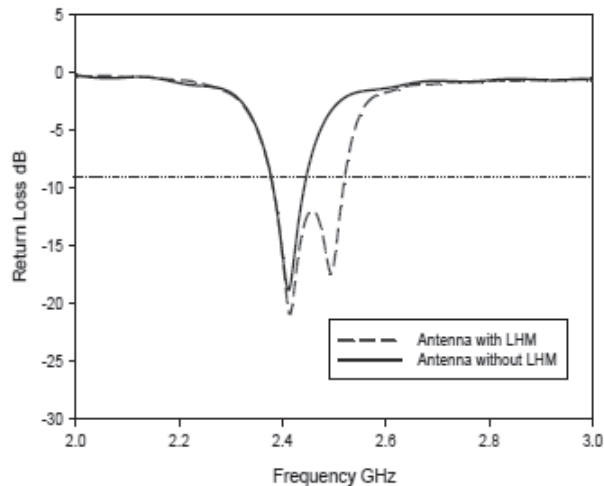


Figure 6: Simulated return loss

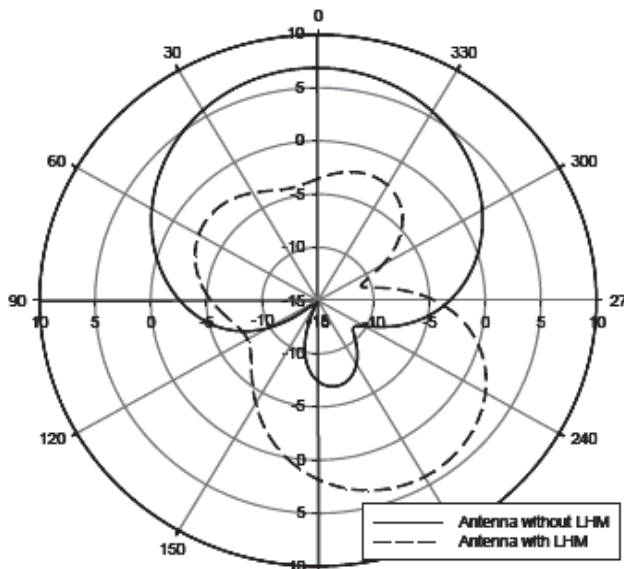


Figure 7: Simulated E-plane radiation Pattern at 2.42 GHz

Antenna	Gain (dBi)	Rad. Efficiency(dB)	Bandwidth GHz	% Bandwidth
Without LHM	6.847	-2.716	0.064	2.65
With LHM	4.148	-2.005	0.135	5.58

Table 2: Antenna gain, efficiency and bandwidth analysis

Conclusions

Microstrip patch antenna with connected LHM is presented in this paper; DNG region for the LHM structure is retrieved to validate the left-handedness of the structures. The simulated results obtained from CST microwave studio shows a bandwidth enhancement with antenna with LHM compare to the antenna without LHM structures.

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Study the Characteristics of Empirical Pseudo – Potential Method for Electronic Structure Calculations

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Abstract

Quantum dots serve as very attractive and interesting objects for scientific research of three-dimensionally confined systems. The experimental and theoretical progress has opened up a new realm of fundamental physics and chemistry as researchers can make and study artificial analogues of atoms, molecules and crystals. Theoretical investigations of electronic structure of quantum dots are of current interest in nano - materials. Empirical theories such as effective mass approximation, tight binding methods and empirical pseudo - potential method are capable of explaining the experimentally observed optical properties. We employ the empirical pseudo - potential to calculate the gap between the highest occupied molecular orbital (HOMO) and the lowest unoccupied molecular orbital (LUMO) as a function of shape and size of the quantum dots. In this paper we study the HOMO – LUMO gap variation with size for CdSe, ZnSe and GaAs quantum dots.

Keywords: Quantum dots, nanophase materials, pseudo – potential, HOMO and LUMO.

Introduction

Research of more than two decades has given a way to tailor materials having desired optical and electronic properties. The main focus of the research in this field is to investigate, experimentally as well as theoretically, the effect of three - dimensional confinement on the electronic structure which is affected by quantum size effects, dielectric confinement effects and high surface to volume ratio for semi - conducting quantum dot. Variation in the optical and electronic properties is substantial for semiconductors as compared to that for the metals. The clusters of metals start showing bulk - like properties for comparatively small sizes (with few hundred atoms) as against semiconductors (where even with few thousand atoms, the bulk like states do not show up). This can be explained by considering the bands of solids which are centred about the atomic energies with the width of the band related to the strength of the nearest neighbour interactions.

As the size of the clusters increases, centre of the band develops first and the edges develop last. In metals, the Fermi energy lies amidst the bands and the relevant energy spacing is small at temperatures above a few Kelvin. As opposed to this, in semiconductors, Fermi energy lies between two bands, hence the edges of the band determine the low-energy optical properties and across the gap, the states are strongly size - dependent even when the crystallites contain few thousand atoms.

Various empirical theoretical methods are being used for understanding the size dependent optical properties of semi - conducting quantum dots. Empirical tight binding methods (TB), effective mass approximation (EMA), empirical pseudo - potential method (EPM) and effective bond orbital model (EBOM) are some of the important methods that are extensively applied in this field to understand the electronic structure and related properties, the optical properties, especially the bandgap variation with size, etc. It may be noticed that the first principles methods have been restricted to work for clusters containing few tens of atoms only. On the other hand, experimental observations are available for nano - particles in the size range of a few nanometers. Thus, to compare theoretical results with experimental data, one has to take recourse of empirical methods.

In simple empirical methods, experimental data is used to adjust and fix electronic structure parameters whereas first principles methods need no experimental data but attempt to support the experimental observations by self consistent calculations. In the present paper, we present our results of electronic structure calculations for CdSe, ZnSe and GaAs, based on empirical pseudo - potential method (EPM) modified for quantum dots.

Method of Finding and Mathematical Calculation

The electrons in a solid can be separated into two categories: itinerant or valence electrons and core electrons. The itinerant electrons in the solids are responsible for most of

their chemical properties. The core electrons are strongly localized near the nucleus and have negligible overlap with the electrons from the neighbouring atoms.

They are thus chemically inert and do not affect the chemical bonding. They can, therefore, be excluded from the electronic structure calculations performed to study the bonding properties and properties governed by the states near the valence band maximum and the conduction band minimum. This fact can be extended for calculations of electronic properties of quantum dots. In the 'frozen core approximation', the core states of the bulk or quantum dot are taken to be identical to those of the isolated atoms.

Pseudo - potentials eliminate the atomic core states and the strong potentials that are used to bind them (Harrison 1966). Hence, the computational efforts to solve the Hamiltonian corresponding to many electron systems are significantly reduced for the valence electrons. The effect of the core electrons and nuclei is replaced by an effective 'pseudo - potential', as shown in figure 1.

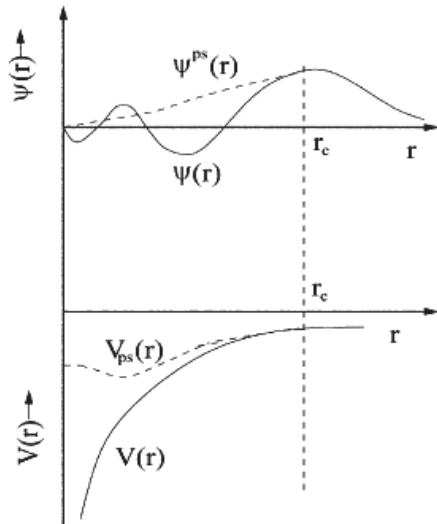


Figure 1: Diagram showing effect of the core electrons and nuclei is replaced by an effective pseudo - potential

This potential is chosen as soft as possible and the nodeless pseudo wave functions are matched with the actual electronic orbitals beyond a certain distance, r_c , often referred to as the critical radius. Pseudo - potentials can be constructed either by ab initio approach (Pickett 1989) or by empirical approach (Cohen and Heine 1970; Cohen and Chelikowsky 1989). We have employed the elegant yet simple empirical approach and we briefly discuss the modifications to employ them for quantum dots (Rama Krishna and Friesner 1991a, b, 1992). In pseudo - potential method, we solve the following

one - electron equation for a valence electron in the state with wave-vector $\vec{k}$ and band index n

$$\left[-\frac{1}{2} \nabla^2 + V_{ps}(\vec{r})\right] \psi_{\vec{k}}^n(\vec{r}) = \varepsilon_{\vec{k},n} \psi_{\vec{k}}^n(\vec{r}) \quad \text{..... (1)}$$

Where atomic units are used throughout; energy is measured in Hartree in a.u. $V_{ps}(\vec{r})$ is the crystal pseudo - potential and contains contributions from all the atoms in the crystal. Expanding the electronic wave - function, $\psi_{\vec{k}}^n$ in terms of plane waves

$$\psi_{\vec{k}}^n(\vec{r}) = \sum_{\vec{G}} C_{\vec{k}}^n + \vec{G}^{\dagger}(\vec{k} + \vec{G}) \cdot \vec{r} \quad \text{..... (2)}$$

Where $\vec{G}$ being the reciprocal lattice vector, then

$$\left[H_{\vec{k}}^{\vec{G}\vec{G}'} - \varepsilon_{\vec{k},n}\right] C_{\vec{k}+\vec{G}}^n = 0 \quad \text{..... (3)}$$

Where

$$H_{\vec{k}}^{\vec{G}\vec{G}'} = \Omega |\vec{k} + \vec{G}|^2 + V_{ps}^{\vec{G}\vec{G}'} \quad \text{..... (4)}$$

Here $V_{ps}^{\vec{G}\vec{G}'}$ is the Fourier transforms of the pseudo - potential $V_{ps}(\vec{r})$, then

$$V_{ps}(\vec{r}) = \sum_{\vec{R},j} V_p^j(\vec{r} - \vec{R} - \delta_j) \quad \text{..... (5)}$$

Where V_p^j is the atomic pseudo-potential of the j^{th} basis atoms situated at δ_j from the corner of the unit cell, and $\vec{R}$ is the position vector of the corner of the unit cell from arbitrary origin. Since the atomic pseudo - potentials are spherically symmetric, it turns out that the form factors, $V_j(\vec{G})$ of the atomic pseudo - potentials depend only on the magnitude of the reciprocal lattice vector, $\vec{G}$ and the number of form factors needed to represent the potential correctly is reduced considerably. It is well known that most of the semiconductors exist in zinc blende or wurtzite structure. The zinc blende structure is composed of two interpenetrating fcc structures with different basis atoms on each lattice. If we assume the origin in between these two atoms, then for a binary compound, the symmetric and anti-symmetric parts of the Fourier components of potential may be written as

$$V_s(\vec{G}) = \frac{1}{2} [v_1(\vec{G}) + v_2(\vec{G})] \quad \text{..... (6)}$$

$$V_A(\vec{G}) = \frac{1}{2} [v_1(\vec{G}) - v_2(\vec{G})] \quad \text{..... (7)}$$

Where $v_1(\vec{G})$ and $v_2(\vec{G})$ are the pseudo - potential form factors for individual atoms. The symmetric and anti symmetric parts of the structure factors for each atom may be written as

$$S_5(\vec{G}) = \cos(\vec{G} \cdot \vec{t}_1) \quad \text{..... (8)}$$

$$S_A(\vec{G}) = \sin(\vec{G} \cdot \vec{r}_1) \quad \dots\dots\dots (9)$$

The reciprocal vectors are defined, in terms of the lattice constant, a_0 , of the material as

$$\vec{G} = \frac{2\pi}{a_0} [G_x, G_y, G_z] \quad \dots\dots\dots (10)$$

Substituting eqn. 10 into eqn.8 and eqn.9, we get

$$S_5(\vec{G}) = \cos\left[\frac{\pi}{4} (G_x + G_y + G_z)\right] \quad \dots\dots\dots (11)$$

$$S_A(\vec{G}) = \sin\left[\frac{\pi}{4} (G_x + G_y + G_z)\right] \quad \dots\dots\dots (12)$$

Thus knowing the Fourier transform of atomic pseudo - potentials for a given $\vec{G}$ shell, it is quite straight forward to generate the pseudo - potential matrix elements in the basis of plane waves. The best choice for expanding electronic wave - functions using pseudo - potentials is plane waves.

Result and Discussion

The reduced dimensionality of the semiconducting quantum dots often results into discretization of energy states giving rise to molecular like orbital energies instead of band structure. We can obtain these energy levels by imposing proper boundary conditions on the wave functions.

Empirical pseudo - potentials which work for bulk materials are used for predicting size quantization effects in quantum dots. This is possible because experimental evidences such as transmission electron microscopy and X - ray diffraction studies have shown that nano-crystallites have same geometric structure as that of the bulk. The bulk pseudo - potentials are generated by using form factors reported by Cohen and Bergstresser (1966) that are accurate to 0.01. We know from elementary quantum mechanics that the wave vector $\vec{k}$ and the corresponding energies are quantized for a particle in a rigid box (Schiff 1968). However, the allowed $\vec{k}$ values are different for different shapes because of the boundary conditions. Also, the magnitude of $\vec{k}$ depends on the size of the box. We have calculated the allowed $\vec{k}$ values by imposing proper boundary conditions, depending on the shape and size of the dots. To assure the accuracy of the pseudo-potentials, we have calculated the complete band structure of the bulk semiconductors using the form factors given by Cohen and Bergstresser (1966). Here, the bulk bandgap is underestimated by 0.03 eV for ZnSe and overestimated by 0.19 for CdSe, respectively. We rigidly add these corresponding differences to the HOMO – LUMO gap of each cluster. Calculations for clusters are carried out till the bulk bandgap is achieved for larger clusters.

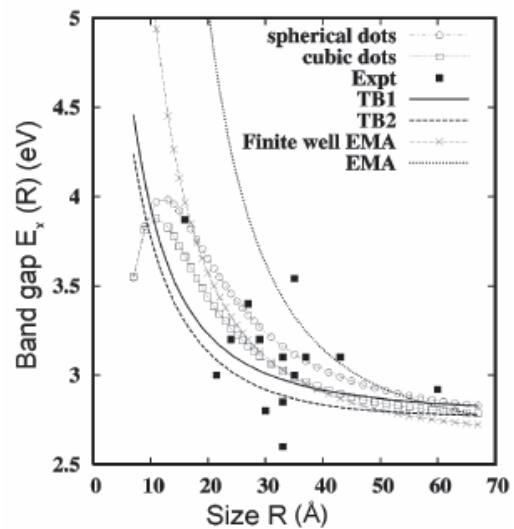
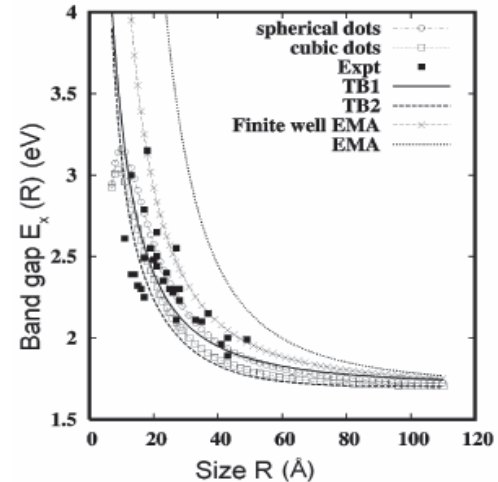


Figure 2: Variation of HOMO – LUMO gap with size for CdSe quantum dots and **Figure 3:** Variation of HOMO – LUMO gap with size for ZnSe quantum dots.

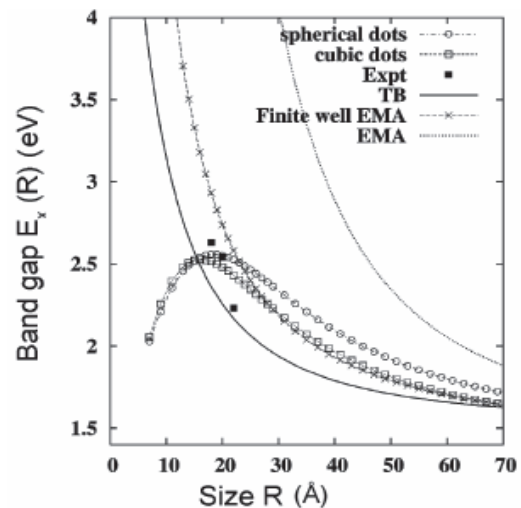


Figure 4: Variation of HOMO – LUMO gap with size for GaAs quantum dots.

Figures 2, 3 and 4 show the excitonic energies calculated within effective mass approximation with infinite potential. All the curves qualitatively show the correct trend viz. an increase in the HOMO – LUMO gap with decreasing dot size. However, as expected EMA highly overestimates excitonic energies in the region of interest, inclusion of Coulomb energy and finite depth potential considerably lowers the excitonic energy and the results are found to be in good agreement with experiments for medium sized clusters. The HOMO – LUMO gap clearly has a tendency to decrease with increasing size and it finally converges to bulk bandgap. A gradual upward shift in HOMO and downward shift in LUMO energies can be clearly seen with increase in size for CdSe and ZnSe.

Conclusion

We have successfully shown that empirical pseudo - potential method which can predict the trends in electronic structure evolution from small clusters to bulk. Small clusters have molecular - like energy levels and with substantial increase in size, the band forming tendency is clearly seen. LUMO energy decreases whereas HOMO energy increases,

comparatively slowly, as size increases. We have also shown results of electronic structure calculations for CdSe, ZnSe and GaAs based on empirical pseudo - potential method (EPM), modified for quantum dots.

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Study on Circuit Partitioning using Clustering Technique in VLSI Circuit

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Abstract

In the earlier period, the chief concern of the VLSI designer was area, performance, cost and reliability; power considerations were mostly of only secondary significance. The benefit of utilizing an arrangement of low-power components in conjunction with low-power design techniques is more important now than ever before. This paper aims at circuit partitioning using clustering technique by applying two clustering algorithms NNA (Nearest Neighbour) and PAM (Partitioning around mediods). These two algorithms were tested on a BCD to Seven Segment Code Converter circuit consisting of eight nodes and were implemented on VHDL. The tested results show that PAM yield better subcircuits than NNA.

Keywords: VLSI, NNA (Nearest neighbour), PAM (Partitioning around mediods), BCD (Binary coded Decimal) and AND gates.

Introduction

Necessities for lower power consumption continue to increase significantly as components become battery - powered, smaller and needed more functionality. In the earlier period the main concerns for the VLSI designers was area, performance and cost. Power consideration was the secondary concerned. Now a day's power is the prime concerned due to the significant growth and achievement in the field of portable. The relevance of VLSI in performance computing, telecommunications, and consumer electronics has been expanding progressively and at a very hasty pace. In order to build complex digital logic circuits it is often essential to subdivide multi – million transistors design into manageable pieces. Circuit partitioning is a general approach used to solve problems that are too large and complex to be handled at once. In partitioning, the problem is divided into small and manageable parts recursively, until the required complexity level is reached. In the area of VLSI, circuit complexity is rapidly multiplying, together with the reducing chip sizes; the integrated chips being produced today are highly sophisticated.

There are many diverse problems that occur during the development phase of an IC that can be solved by using circuit partitioning which aims at obtaining the sub circuits with minimum interconnections between them. Advances in semiconductor technology and in the integration level of integrated circuits have enhanced many features, increased the performance; improved reliability of electronic

equipment, and at the same time reduced the cost, power consumption and system size. As size and complexity of digital system has increased, more computer aided design tools are introduced into hardware design processes. VLSI design automation has attracted a great deal of interest. A recent survey lists almost 200 papers on the subject. Circuit partitioning in VLSI design is key role in physical design. The objective of circuit partitioning is to divide the circuit into number of subcircuits with minimum interconnections between them.

In past two decade, partitioning problems has been studied by researchers and various heuristic algorithms have been developed. The circuit partitioning is also achieved by using some of the clustering algorithm. In this paper to achieve circuit partitioning, with minimum interconnections is obtained by applications of two clustering algorithms PAM and NNA. These two clustering algorithms are applied to BCD to Seven segment circuit consisting of eight nodes. Results show that PAM gives better clusters with minimum interconnections. Both the clustering algorithms are run on VHDL code and are tested.

Implementation of the Theory

The implementation consists of three stages consisting of data extraction, partitioning and result. In data extraction, a VLSI circuit of BCD to 8 segment converter is considered. The circuit considered is shown in figure 1(a).

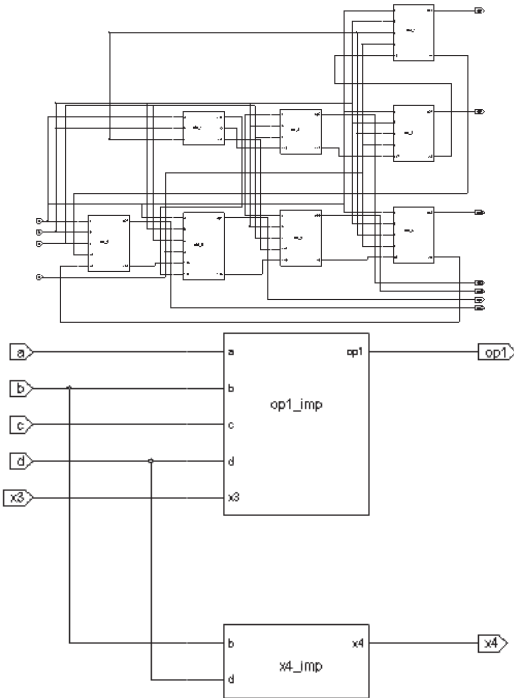


Figure 1(a): Circuit diagram and **1(b):** Sub circuit

In the above circuit each rectangular box is considered as a node. The circuit consists of 8 nodes with interconnection between them. The sub circuits within one the node is shown in figure 1(b). The input to this circuit is a BCD (Binary coded Decimal) code and output is seven segment codes. Figure 2(a) shows the internal view which consists some of the basic gates like AND, OR and inverter gates. Similarly each node consists of all the basic gates.

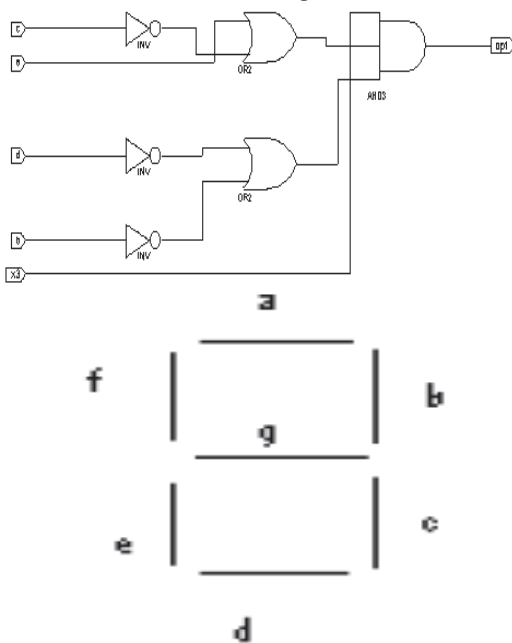


Figure 2(a) and 2(b)

The output is the seven segment display which is depicted in fig 4. The input to the circuit and the corresponding output obtained is shown in table 1.

Input 1	Input 2	Input 3	Input 4	a	b	c	d	e	f	g
0	0	0	0	1	1	1	1	1	1	0
0	0	0	1	0	1	1	0	0	0	0
0	0	1	0	1	1	0	1	1	0	1
0	0	1	1	1	1	1	1	0	0	1
0	1	0	0	0	1	1	0	0	1	1
0	1	0	1	1	0	1	1	0	1	1
0	1	1	0	0	0	1	1	1	1	1
0	1	1	1	1	1	1	0	0	0	0
1	0	0	0	1	1	1	1	1	1	1
1	0	0	1	1	1	1	0	0	1	1

Table 1: Input and Output for the BCD to Seven segment Circuit

Results and Discussion

A. Summary of Clustering Algorithm

Clustering is referred as unsupervised learning segmentation. The clusters are formed by finding the similarities between data according to characteristics found in the data. It can be thought of as partitioning or segmenting the data into groups that might or might not be disjointed. In this paper we are using Nearest Neighbour Algorithm and PAM (Partitioning around medoids). This is serial algorithm in which the items are iteratively merged into the existing clusters that are closest. In this algorithm a threshold, t is used to determine if items will be added to existing clusters or if a new cluster is created the complexity of the algorithm depends on the number of items. For each loop each item must be compared to each item already in a cluster. The time complexity is $O(n^2)$.

B. PAM Algorithm

These algorithms represent a cluster by a medoid. Using this algorithm handles outliers well. Quality of clusters is measured by the sum of all distances from a nonmedoid object to the medoid for the cluster it is in. We use C_{jih} to be the cost change for an item t_j associated with swapping medoid t_i with nonmedoid t_h . The total complexity per iteration is $n(nk)^2$. The cost is the change to the sum of all distances from items to their cluster medoids. The total impact to quality by a medoid change TC_{ih} is given by

$$TC_{ih} = \sum_{j=1}^n C_{jih}$$

In data extraction, the circuit which is a BCD to seven segment code converter consisting of 8 nodes is considered. The circuit is represented by an adjacency matrix where the values represent the distances between the nodes. The adjacency matrix is fed into algorithm which gives the subcircuits with lowest amount of interconnections between them.

Input	No. of Cluster	
	PAM	NNA
BCD to Seven segment	2	4

Table 2: Performance of PAM and NNA

The partition is tested on the circuit. NNA. Experimental results shows that NNA gives four clusters where are PAM gives two clusters for the same input as shown in table 2. The algorithm is implemented on Xilinx 6.1 using project Navigator using VHDL.

Conclusion

In this paper the Circuit Partitioning which plays a critical role in physical design of an integrated circuit with minimum interconnections is obtained by implemented using two

clustering algorithm PAM and NNA. Experimental results shows that NNA gives four clusters where PAM gives two clusters for the same input. The circuit also shows the minimum interconnections between them.

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Study of a Broadband Planar Sierpinski Fractal Antenna for Multiband Applications

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Abstract

Fractal antenna engineering focuses on two areas one is the design and analysis of fractal elements and other concerns to the applications of the fractal theory to the design of antenna arrays. Both the antenna types are highly desired in various commercial and military sectors. In this paper I shaped fractal antenna is discussed. This antenna finds applications for defence and secures communication C band (4 GHz to 8 GHz) and X band (8 GHz to 12 GHz) applications where antenna can be used for satellite communication and RADAR application.

Keywords: Fractal antenna, multiband application, antenna array and space filling properties.

Introduction

With the huge advancement in wireless communication systems and increasing its applications, multiband antenna with various shapes and design have become a great demand and desirable for various wireless applications, which gives raise to different shapes and types of antenna for the variations in antenna characteristic. There are different shapes and types of antennas designed for the variations in antenna characteristics. There are various shapes such as Sierpinski, Gasket, Minskowski, Hilbert curve; Koch curve, tree structure, etc are considered for the design of Fractal technique has got numerous applications in the field of science which includes fractal electro dynamics where the fractal concepts are combined with electromagnetic theory. Fractal antenna engineering focuses on two areas one is the design and analysis of fractal elements and other concerns to the application of the fractal theory to the design of antenna arrays. Both the antenna types are highly desired in various commercial and military sectors. Most of the fractal antenna elements possess compact size, low profile, cost effective, multi-band operations, easy feeding and their shape can be modified for better optimization. The fractal geometries feature two properties self - similarity and space filling properties. Both these properties become a reason attractive to consider fractal geometry. Similarity of antenna holds the duplication of it at several scales and can operate at several wavelengths in similar fashion, which reduces the antenna size and occupies lesser space. The space filling property fill the area occupied by the fractal antenna as the number of iteration is increased. Higher the order of fractal antenna, lesser the antenna size .This reason makes the fractal antennas compact, multiband and use full for cellular applications and microwave applications. A dual frequency antenna was obtained by replacing a segment of a square ring microstrip antenna with fractal Minkowski curve. To control resonant characteristics, width of two sides had been increased. Stacked fractal antenna for reducing radio interference had been proposed for GPS application using

fractal geometry. In this two antennas with different radiation pattern were stacked one over other. A novel and compact printed monopole antenna was presented with simple but radiating patch for multiple wireless communication systems and mobile devices. Multiband operation with circular polarization of radiation was obtained by combining square and Giuseppe Peano fractal geometries, which were realized on two layer microstrip antenna. A novel CPW feed inscribed triangle circular fractal patch antenna had been proposed. Proposed monopole fractal antenna offered UWB (Ultra Wideband) characteristics from 2.25 GHz to 15 GHz at VSWR 2:1. A compact, low profile and low cost self-affine antenna with wideband characteristics for wireless applications had been proposed using cantor fractal geometry. Proposed antenna had a gain of 5.04 dB at 2.5 GHz. A novel microstrip fed antenna design based on monopole patch for triple frequency operation had been presented using DGS (Defected Ground Structure). A novel microstrip fed ultra wideband printed Pythagorean tree fractal monopole antenna had been presented. By only increasing tree fractal iterations, new frequency bands were obtained. A monopole square patch antenna was introduced by shaping its edges in form of Giuseppe Peano fractal and its surface area in form of Sierpinski carpet fractal. It found application for UWB and WLAN (Wireless Local Area Network). New design to suppress harmonic radiation from microstrip antenna with proximity coupled feeding line had been implemented in multilayer substrate. Microstrip line fed printed wide slot antenna with a fractal shaped slot for bandwidth enhancement by etching wide slot was proposed and experimentally studied. A Novel technique to reduce the size of microstrip patch antenna was proposed. By etching the patch edges according to Koch curve and inserting Sierpinski carpets into patch, it was found that resonant frequency of antenna get reduced. A wideband vertical patch antenna was depicted from fractal antenna concept. By using a dual Koch loop structure, a wideband antenna with 42%

bandwidth and 8 dBi gain at center frequency was obtained. Koch iteration technique had been applied to obtain two fractal versions of folded slot antenna. A broadband planar Sierpinski fractal antenna for multiband application was proposed. Fractal geometry and slotted ground had been used to obtain broadband characteristics.

Proposed Antenna Geometry

By applying fractal geometry on the patch, area of patch decreases and resonant length increases. Algorithm has been applied to square. By Dividing square of length 37.6218 mm into three equal parts of 12.5406 mm each and by cutting 2 squares of 12.5406 mm length from parallel sides, I shape patch as shown in figure 1 has been formed. The dimensions of proposed I shape fractal antenna has length and width of patch is 37 mm each, thickness of the substrate is 2.6 mm with dielectric constant 2.2 mm. A Coaxial feed with feed point coordinates are selected in such way that impedance matching takes place. By applying 3rd iteration, the patch has been modified by removing 2 squares of length 1.3934 mm from each square of length 4.1802 mm. Hence geometry as shown in figure 1 has been obtained.

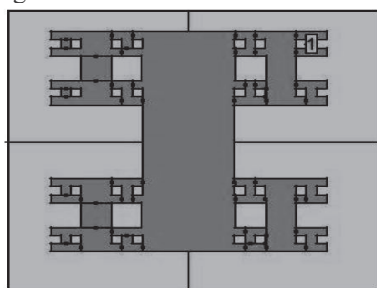


Figure1: Proposed Structure

Result and Discussion

After application of three iterations of fractal geometry on square patch, I - shaped fractal microstrip patch geometry has been formed. Antenna as shown in figure 1 resonates at 5.89 GHz, with a return loss of -13.87 dB, VSWR of 1.5 and 245 MHz bandwidth. Different iterations of fractal geometries have been applied one by one on square patch and results are analyzed as shown in table II. Figure 2 and 3 shows return loss Vs frequency graph and VSWR Vs frequency for 3rd iterated fractal patch geometry respectively. In this study bandwidth values are obtained between points where - 3 dB return loss line cut the curve between return loss and frequency.

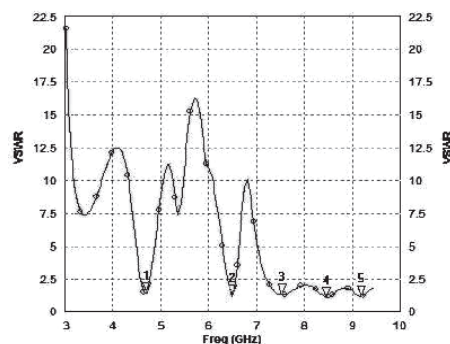
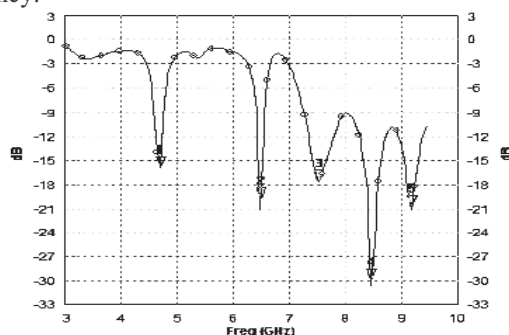


Figure 2: Return loss and 3: VSWR

From radiation patterns, it is clear that this antenna has maximum energy radiated in main lobes and small amount of energy is radiated in minor or side lobes. These curves show the return loss and VSWR curve and simulation of this theory has been done by HFSS software. The simulation result is in good agreement with the proposed structure.

Conclusion

By applying different iterations of fractal geometry, characteristics of antenna improve a lot. By applying three iterations of fractal geometry, number of bands increase from one to five. Further by applying fractal geometry effective area has been decreased and gain and bandwidth gets increased. Proposed antenna finds application for defence and secure communication, C band and X band applications. In these bands antenna can be used for satellite communication and RADAR applications.

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Study of Metamaterials used as a Superstrate to Enhance the Antenna Characteristics

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Abstract

Microstrip antennas (MAs) have been widely used in microwave systems because of their attractive features, such as low profile and low cost. Although MAs have been considered significant advances in antenna technology, they suffer from a number of serious drawbacks such as very narrow bandwidth and low gain. In order to improve the antenna characteristic, metamaterials (MTM) are used as a superstrate. In this paper, Multi - band antenna along with the innovative metamaterial structure is proposed which consists of a circular geometry incorporated with c shaped structure. This work is mainly focused on increasing the potential parameters of planar antennas and analyzing the multi band operation of proposed antenna. For verification of the proposed metamaterial structure possesses negative values of permeability and permittivity within the operating frequency ranges, Nicolson – Ross - Weir Method (NRW) has been employed. Simulation has been done by HFSS Software.

Keywords: Microstrip antenna, bandwidth, metamaterials, superstrate and HFSS software.

Introduction

Wireless communication systems have grown dramatically in the last few years. These systems have a great penetration on the society. Several examples can be mentioned: mobile communications (GSM, DCS, UMTS), personal area networks (Bluetooth), local area wireless networks (Wi Fi), radio navigation systems (GPS, Galileo), etc. Related to this fact, the popularity and sales of wireless terminals have also grown. The first user terminals which become so popular were mobile phones. However, mobile phones are not the only user terminals extended in the society. We have several examples: PDAs, laptops and more recently terminals embedded inside cars. Development in communication systems requires the development of low cost, minimal weight, low profile antennas that are capable of maintaining high performance over a wide spectrum of frequencies. The future development of the personal communication devices will aim to provide image, speech and data communications at any time, and anywhere around the world. This indicates that the future communication terminal antennas must meet the requirements of multi-band or wideband operations to sufficiently cover the possible operating bands. The performance of the fabricated antenna was measured and compared with simulation results. Moreover, we have also indicated the appropriate choice of particular metamaterial for different specific purposes like antenna size reduction and other mode modification-related applications. The performance of a rectangular patch antenna array on a metamaterial substrate was studied relative to a similar array constructed on a conventional FR4 substrate. In modern wireless communication systems, the microstrip patch antennas are commonly used in the wireless devices. Therefore, the miniaturization of the antenna has become an important issue in reducing the volume of entire communication system.

In modern wireless communication systems, the microstrip patch antennas are commonly used in the wireless devices. The demand in commercial and military wireless systems is due to capabilities of proposed Antenna such as low weight, low profile, low cost, easily combined with design and technology, and relatively simple fabrication. All these antennas can also be fabricated by using HFSS simulation software and get very sharp characteristics. Proposed Antenna can be largely used in many wireless communication systems because of their low profile and light weight. The theoretical concept of metamaterials to increase the transmitting power was given by Victor Veselago (1968), Engheta and Ziolkowski (2006), Pendry (2000) Garg et al. An array of metallic wires can be used to obtain negative permittivity and split ring resonators for negative permeability (Pendry et al., (1999)). By using the available information a structure composed of split ring resonator and thin wire fabricated by Smith et al. (2001) possessed the negative permittivity and permeability simultaneously. This structure was named as LHM (Wu et al., 2005) (Burokur et al., 2005). Metamaterial is an artificial material that exhibits unique properties in electromagnetic spectrum unlike those of conventional materials available in nature. The conventional material available in nature have positive magnetic permeability and almost positive dielectric permittivity, whereas metamaterial exhibits negative permeability, μ and/or negative permittivity, ϵ . Due to these significant features metamaterial is also referred to as single negative(SNG) or double negative(DNG) material. In metamaterial, the refractive index is negative i.e. the group and phase velocity of the electromagnetic wave appears in opposite direction. The reversed direction of propagation with respect to direction of energy flow leads to reversed Doppler shift. These materials are also referred to as negative index material (NIM) because the refractive index is negative

hence both refracted and incident wave remains on same side.

The potential applications of metamaterials at microwave frequencies include (a) substrate materials for antenna and microwave component designs and fabrications, and (b) absorbers for engineering and radar applications. Split ring resonators (SRRs) and some other planar structures were applied to enhance the return loss.

In this paper a novel shape of S R R metamaterial structure is used to antenna patch that provide an antenna with multiband application.

Result and Discussion

In order to calculate the S - Parameters, the proposed metamaterial structure is placed between the two waveguide ports at the left and right hand side of the X axis.

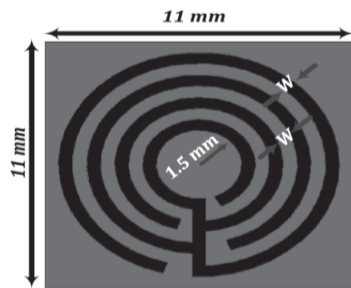


Figure 1: Structure of Metamaterial

In Fig.1, Y-Plane is defined as Perfect Electric Boundary (PEB) and Z - plane is defined as the perfect magnetic boundary (PMB) to create internal environment of waveguide. The simulated S - Parameters are then exported to Microsoft Excel Program for verifying the Double - Negative properties of the proposed metamaterial structure. Nicolson – Ross - Weir (NRW) technique has been used for obtaining the values of permittivity and permeability.

$$\mu_r = (2 \times c \times (1 - v_2)) / (\omega \times d \times i \times (1 + v_2)) \quad (1)$$

$$\epsilon_r = \mu_r + (2 \times S_{11} \times c \times i) / (\omega \times d) \quad (2)$$

$$v_2 = S_{21} - S_{11} \quad (3)$$

Where ϵ_r = Permittivity, μ_r = Permeability, ω = Frequency in Radian, d = Thickness of the Substrate, c = Speed of Light, and v_2 = Voltage Minima

The configuration of the novel metamaterial model that is investigated in this paper is shown in Fig. 1. The results of the return loss and relative constitutive parameters of the unit cell (permittivity and permeability) are shown in Fig. 2 a, b and c respectively.

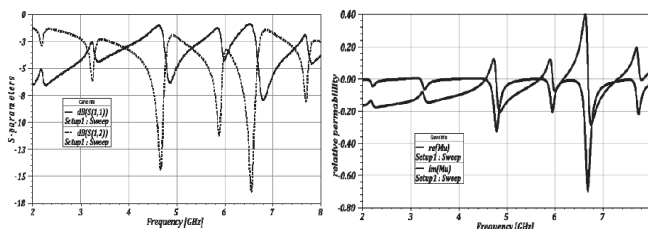


Figure 2(a): S parameter

Figure 2(b): μ

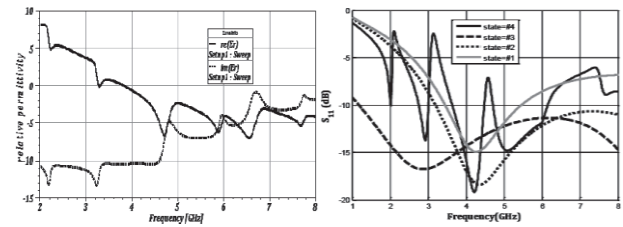


Figure 2(c): ϵ

Figure 3: Simulated return loss

The constitutive parameters are calculated from S - parameters by using retrieval method. As it is seen, both of the permittivity and permeability parameters are negative in a specific frequency band. So this structure shows double negative property and it can be used as a DNG medium.

Fig. 1 shows the geometry of the proposed antenna. The substrate width is 20 mm and its length is 22 mm. This antenna is fed by a CPW line inset feed.

The width of the feed line 3.1 mm and the depth of the inset 7mm have been adjusted to match the antenna impedance to 50 Ω . The distance between inset line and the patch (y) is 1.2 mm. In figure 3, simulated return loss result of four steps of antenna is shown. The simulation was done by HFSS software and result was in good agreement with the experimental results.

Conclusion

On the basis of the simulated results it is observed that the proposed antenna parameters has improved significantly by employing proposed metamaterial structure at 1.6 mm layer from the ground plane of the antenna. There are very less variation has been observed between the experimental results and the simulated result of proposed antenna.

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Analysis of Photonic Band-Gap Reflector with a Spiral Antenna

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Abstract

A spiral antenna backed by photonic band-gap (PBG) reflector is analysed. The peripheral length of the antenna is in between 1 and 2 wavelengths, which is optimized for first-mode circularly polarized (CP) radiation in presence of a flat finite PEC reflector. The analysis uses the finite- difference time domain method (FDTD). As the antenna height decreases, the reverse current towards the feed point increases, which results the deterioration of the axial ratio. To overcome this deterioration, the PEC reflector is replaced by PBG reflector composed of mushroom- like elements. Analysis reveals that the spiral radiates CP wave even when spiral is located close to the PBG surface.

Keywords: Photonic band-gap reflector, finite difference time domain method (FDTD), Spiral antenna.

Introduction

During past two decades periodic structures such as photonic band-gap (PBG) materials have attracted extensive research interest in the microwave and millimeter wave domains. The major characteristic of PBG surfaces is the existence of one or more bands whereby propagating surface waves are effectively suppressed. In addition, many of these structures also have the property that they reflect normally incident plane waves with a unit reflection coefficient within certain frequency-bands. There are numerous applications of these periodic materials in the field of antenna design [1] – [3].

The properties of PBG only exist within a specific band and thus careful design is necessary in order to select the resonant frequency and the band-width (BW) of a particular array geometry.

This paper focuses on the effect of radiation characteristics of using a photonic band-gap reflector with spiral antenna. The mushroom- like structure is known to have an effective band-gap for surface-wave propagation, which can be useful to improve antenna radiation patterns. However, the surface wave band-gap cannot guarantee the effective radiations of a low profile wire antenna. This is because complicated interactions occur between the wire antenna and the EBG surface and electromagnetic waves are not restricted to surface waves that propagate in the horizontal plane. In light of this a dipole antenna is horizontally positioned near the EBG surface and its radiation performance is accurately characterized using the FDTD method. By observing both the return losses and radiation patterns of dipoles operating at different frequencies one can find the useful operational

frequency band of the EBG surface. Since the surface-wave frequency band-gap that improves radiation patterns has been thoroughly addressed in [4] and [5].

The new antenna structures have been investigated where the conventional PEC reflector is replaced with other types of reflectors: corrugated reflectors [6] and electromagnetic band-gap (EBG) reflectors [7]. A dielectric substrate drilled with a rectangular lattice of circular holes have been applied to an equiangular spiral antenna [7]. The experimental results show that the reflector is useful in improving radiation characteristics. However, little data based on theoretical analysis of spiral antenna with this reflector is currently available. In this paper we deal with the photonic band-gap reflector using theoretical technique based on the finite-difference time domain (FDTD) method. Since FDTD is a time-domain method, the starting point for its implementation is the time-dependent Maxwell equations. Specifically, the curl equations are discretized using a rectangular grid (which stores the field components and the material properties ϵ and μ) and central differences for the space and time derivatives. The procedure results in a set of finite difference equations, which updates the field components in time. The FDTD equations for various types of materials, together with computational procedure, source incorporation procedure stability criteria, and so on are presented in a very clear and complete way in [8].

Numerical Analysis

In this analysis firstly, a spiral antenna backed by a perfect electric conductor reflector is considered which is referred to

as PEC spiral antenna. The outer most peripheral length of more that one wavelength but less than two wavelengths is taken. This peripheral length is the length required to support first mode CP radiation. The PEC spiral is optimized to radiate a CP wave at an antenna height of one-quarter wavelength. The analysis reveals that as the antenna height decreases from one-quarter wavelength, the cross-polarization component increases. Next, this reflector for the spiral is replaced with a photonic band-gap reflector composed of mushroom-like elements. The analysis of PBG-spiral is performed for a small antenna height. The distance between the spiral and PBG surface is 0.06λ .

In Fig. 1, a spiral antenna backed by very thin flat reflector is shown where h is the distance between spiral and PEC plane. The two arms of width w each composed of N -elements, consists of the spiral. The element lengths are : S_1 (First element length) and $S_i = Z(i-1)S_1$ for $i = 2, 3, \dots, N$. The peripheral lengths is chosen to be $1.44\lambda_6$ with $S_1 = 0.03\lambda_6$ where $N = 7$ and λ_6 is the wavelength at a test frequency 6 GHz. This spiral antenna supports the first mode radiation of 6 GHz [9].

It is seen that the spiral radiates CP-wave at $h = \frac{\lambda_6}{4}$ as

desired. The spiral is optimized to radiates CP-waves. The cross polarization component is small and axid ratio in Z -direction is calculated to be 0.6 dB. The high axial ratio is due to an increased reverse current flowing towards the feed point which generates a cross polarization component.

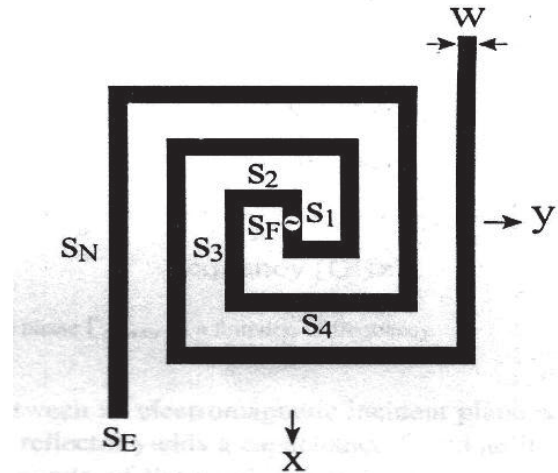
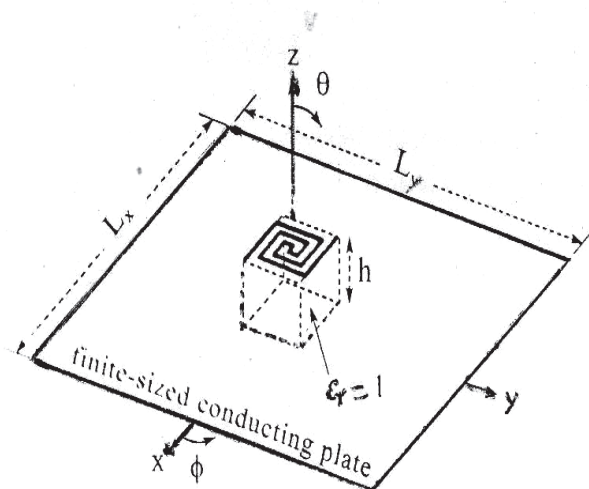


Fig. 1(a) and 1(b):

In order to improve the deteriorated axial ratio, the PEC reflector is replaced with PBG-reflector. This reflector consists of Mushroom like elements having various patches each being connected to ground plane with conducting pin. The patch printed on a dielectric substrate of relative permittivity ϵ_r is square with side length S . The thickness of the reflector is denoted by B . The number of patches in the x and y directions are N_x and N_y , respectively, and the space between the neighboring patches is denoted by δ . The PBG reflector is so designed that it resonates at 6 GHz as used for PEC reflector, using $\epsilon_r = 2.2$, $S = 0.2\lambda_6$, $B = 0.04\lambda_6$ and $\delta = 0.02\lambda_6$. The Fig. 2, shows the reflection phase when an EM-plane wave illuminates the PBG reflector. As seen from this figure, the M-EBG reflector has a reflection phase of $\Gamma \approx 0^\circ$ at 6 GHz, realizing the characteristic of a PEC.

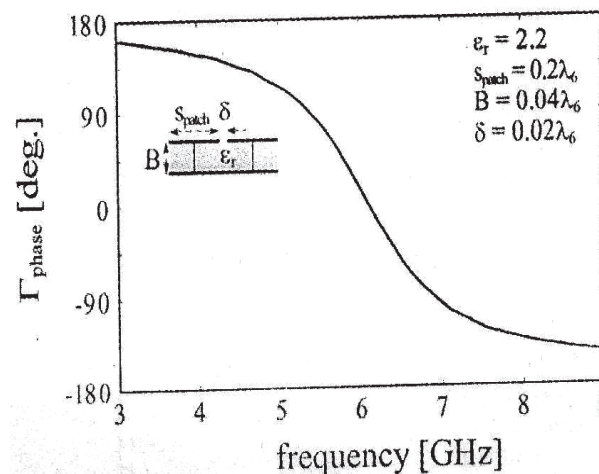


Fig. 2

The theoretical radiation pattern of mushroom like PBG spiral antenna at 6 GHz together with the experimental

results [10] is shown in Fig. 3. It is clear that the CP radiation has been restored for this spiral, while maintaining a small antenna height of $h = \frac{1}{10} \lambda_6$. The theoretical axial ratio is approximately 1.2 dB in the z direction.

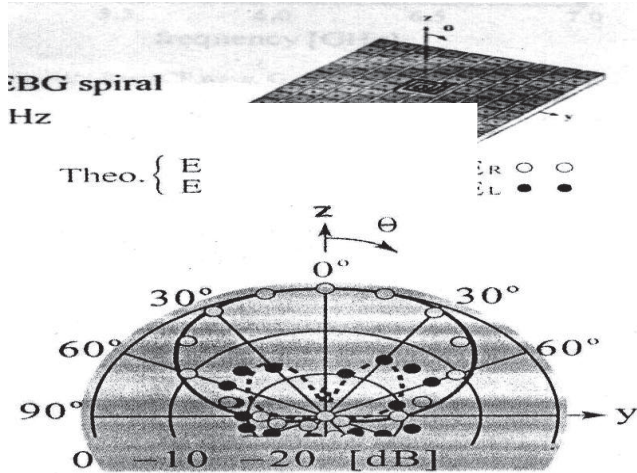


Fig. 3

Conclusion

The analysis of a spiral antenna with mushroom like photonic band gap reflector is made. A spiral antenna, whose peripheral length is more than one wavelength, but less than two wavelengths, is optimized for first-mode CP radiation in presence of a flat finite PEC reflector. The optimization is performed for an antenna height of one-quarter wavelength. As the antenna height decreases, the axial ratio for the spiral deteriorates. Improvement in the axial ratio at a lower antenna height is investigated. For this investigation, a representative antenna height of $h = 0.1 \lambda_6$ and the distance from the PBG reflector surface to the spiral is extremely small as $h' = 0.06 \lambda_6$. The analysis reveals that the spiral backed by PBG reflector radiates a circularly polarized wave as desired.

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Role of Smart Antennas in Mobile Communication

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Abstract

Increase in global demand of data, voice and video has led to some major changes in infrastructure. Coaxial cable, power line and fiber optics all have certain limitations in terms of cost and bandwidth. a real sense of communication is a “last mile” connection which provides a worldwide network, requires both cost and bandwidth. if we view this constraint from a technical point, the information transmission requires resources in the form of power and bandwidth. In wired communication, the transmission is set up independently over each link and in fiber optics, the useful information or energy is confined to a small region in space making the system more reliable and efficient. But same is not true in the wireless communication; reliability is achieved on the cost of wastage of energy. Transmission over short distances in space requires a large amount of space, from which a small portion is received by the intended user and rest behaves as a interference to the other potential users of the system.

Keywords: Adaptive Array Antenna; Delay Spread; Fading; Smart antenna; Switched Beam Antenna.

Introduction

The range is defined by the amount of power that can be transmitted and the capacity is given by the available spectrum. For a given amount of power and a fixed amount of bandwidth, (the resulted capacity i.e. bits/sec/hz/unit area, really per unit volume) is finite that can be made available to user by operators to any given locations. Specifically two problems had to be overcome in such systems: spectral efficiency or ways to increase the capacity at lower cost while maintaining the quality in the area where demand is high and coverage or ways to increase the coverage area where demand is less so that the infrastructure and maintenance costs decreases. [1].

With the emerging new network, the many advancements in past few years came into picture like improved air interface, modulation schemes, deployment of smaller radio cells with combinations of different cell types in hierarchical architectures and advanced signal processing but none of them was able to fully utilize the multiplicity of spatial channel that arises because each mobile occupies a unique place in the space. This approach of utilizing space as a domain to separate the spectrally and temporally overlapping signals from different mobile users was termed as space division multiple access (SDMA). it is a hybrid multiple access technique enhancing the FDMA, TDMA and CDMA schemes. SDMA allows multiple users within the same radio cells to use same frequency and time slot. The full exploitation of this technique in mobile communication was

done through smart antennas. smart antenna can be defined as array of antenna system that adjust itself to the change in environment i.e. it changes its antenna pattern dynamically to adjust with the noise, interference in the channel and mitigate multipath fading effects on the signal of interest. In contrast to simple RF antenna, it uses digital signal processors, general purpose processors and software signal processing algorithms to achieve optimum transmitted or received radiation pattern. In true sense, an antenna is not itself smart.

Types of Smart Antenna

The different types of smart antenna deployed according to the level of intelligence are as follows:

1. Switched Beam Antennas: it is deployed at base Station having multiple directive antennas or predefined beam of an array. It switches between one beam to another on the basis of received signal power. The output is sampled to give the best reception beam. It is an enhanced approach towards the cellular sectorization in which each sectorized cell has three 120° macro sectors. These macro sectors are further divided into micro sectors, these sectors will have fixed beam pattern and it is selected according to the weighted sum of the combined antenna array. These beam have highest sensitivity at centre and lower near the periphery. The selected beam is moved in different direction by changing the phase of the current for reception of maximum desired signal. When the mobile station enters a macro sector, a fixed micro sector beam pattern is chosen according to the best reception and the antenna system monitors regularly the signal strength to

other fixed beams. In comparison to Omni directional, the switched beam increases the range of abse station to about 20-200%. There are some limitations like:

As the user moves toward the periphery, the signal strength decreases rapidly before it comes under the other fixed beam pattern.

It cannot distinguish between interference signal and user signal. If the noise signal is present at the center of the beam and user at the periphery, noise signal would be enhanced similar to user signal thus degrading the strength of desired user signal.

2. Dynamically Phased Array: In contrast to switched beam antenna, this type of smart antenna does not utilize predetermined and fixed beam. It checks the user signal by tracking through a Direction of Arrival (DOA) algorithm. In case of fixed beam, as the signal crosses the mid of beam and reaches periphery, the signal strength decreases and gain reduces while intra cell handover.

3. Adaptive Antenna Array: This type of smart antenna is considered to be smartest of all. An adaptive antenna array is a set of array elements that changes their radiation pattern according to the changes in environment. These variations are updated regularly so that it has maximum gain in the direction of desired user and minimum towards channel noise and interference in order to improve SNR of s desired signal. This procedure is also known as “adaptive beam forming” or “digital beam forming”. They are further classified into forms: (i) Phased Array : in this array, only phase of the current is changed according to the weights. (ii). Adaptive Array antenna: it is a strict sense array in which both the amplitude and phase of the current change to produce a beam in desired direction to maximize SNR. With the use of sophisticated algorithms, the adaptive antenna array distinguish between the interference, multipath and desired signal. This helps to move the beam towards the desired signal and nulls towards the interference according to calculation of direction of arrival by the algorithm. Another advantage of adaptive antenna that makes it more efficient than switched beam is its ability to create spectrum. Through the technique of accurately differentiating between desired user, interference signal and multipath, the adaptive array makes it possible to use the channel within the same cell users. Capacity enhances by decreasing intra cell and inter cell frequency reuse pattern. In switched beam, the beam moves only for desired signal and not in response to the interfering signal. As when the interfering signal approaches near the centre of beam, it get identical processing degrading the overall communication. But in adaptive array, it is able to distinguish between the desired user and interfering signal. It

will hand over any two channel user whether they are intra cell or inter cell before they get too close and interfere with each other [1,2].

Comparison of Smart Antenna

Antenna is a type of transducer that converts the electrical energy into a electromagnetic waves for transmission purposes to the outside world and in reverse collect the RF energy from the outside world for reception. Antennas are most neglected part in mobile communication. But the goal is to study why smart antenna replaced a single omnidirectional antenna at base station.

Omni directional Antenna: - Earlier when no specific knowledge about user was available or needed, the single dipole antenna radiates and receives the RF energy well in all directions. But the amount of energy reaching the desired user is small as compared to the transmitted energy due to unfocused scattering of signals. the only method of improving the overall signal broadcast system is by boosting the power level of signals. if the signal miss the intended user, it will behave as an interference to the other users. in this approach, there was no strategy to reject signals interfering with several users, adversely affecting the spectral efficiency, limiting frequency reuse. but these limitations gained attention in the role of antenna system in mobile communication.

Directional Antenna and Sectorized Systems: A single antenna can be used to have directional transmission and reception by deploying sectorized antenna system. This approach uses same traditional cellular area by dividing them into sectors and each sector is covered using directional antenna from same base station. Each sector can be treated as an independent cell. This system increases the range as compared to Omni directional antenna due to focused power and it is commonly referred to as antenna element gain. Sectorized Antenna system increase the possible reuse of a frequency channel in such cellular system by decreasing the interference across the original cell. Adjacent sectors use different frequency to reduce co-channel interference, by increasing handoff between sectors. This technique stills lag behind in overcoming the major disadvantage of filtering the unwanted interference from adjacent cells [1].

The major problems encountered in wireless communication that limits the performance are : Multipath Fading due to multiple signals reaching the receiver antenna, the overall amplitude and phase vary in antenna location, direction and polarization as well as the with the time of the received signal. Delay Spread- it is the difference in propagation delay among the multiple paths. When this

factor exceeds the ten percent of symbol duration, it results in inter symbol interference thus limiting the maximum data rate. The reduction in average output SNR ratio for a given BER with fading is termed as diversity gain. Co-channel interference- the number of available frequency channels are subdivided into channel set and each channel set is allocated to cell for transmission and reception using frequency reuse technique. Due to improper spacing and increase in the capacity of each cell, co channel interference increases with the decrease in channel set. In TDMA systems, the co channel interference is mainly from one or two user. On the other hand, in CDMA system the co channel interference occurs both within the cell and from adjacent cells. Diversity gain can be decreased in three different ways: Spatial Diversity – the antennas are spaced far enough to provide lower diversity gain. The angular spread or the angle at which the antenna receives the signal decides the spacing between them. In case of mobile handset and indoor base station angular spread is 3600, therefore spacing of quarter wavelength proves to be an efficient whereas for outdoor base station, the angular spread is very low therefore 10-20 wavelength spacing is required, making the size of the antenna array an issue.

Polarization Diversity

Two orthogonal polarizations are used. These orthogonal polarization have low correlation and the horizontal polarization can be 6-10 dB lower than the vertical polarization thereby diversity gain. Angle Diversity – the signal is arrived at the main lobe having same angular spread therefore other adjacent lobes can be used to receive weaker signal thereby decreasing the diversity gain. Diversity can be utilized for increasing the overall strength of the received signal by using: Switched Diversity – Any one of the antenna is in favorable location at a moment to receive the highest possible signal power.

Components of Smart Antenna

Smart antenna receiver is mainly divided into four units – antenna array, radio unit, beam forming unit and a signal processing unit. Antenna array has a set of antenna element. The signals received from these elements are combined and then send to rest of the units for further processing. Generally the number of elements is kept low to avoid any complexity. Different types of array geometries are employed for covering different areas- Uniform linear array and Circular array are mainly used for beam forming in the horizontal plane (azimuth). Uniform linear array has an element spacing of about Δx and perform beam forming in azimuth angle within an angular spread. Circular array has a bird view with

an element spacing of $\Delta\phi = 2\pi/M$ and can perform beam forming in all the azimuth angles without any transcends on angular spread. 2- Dimensional and 3- Dimensional grid array are desirable for indoor or dense urban environments. It can perform beam forming in azimuth angle within an

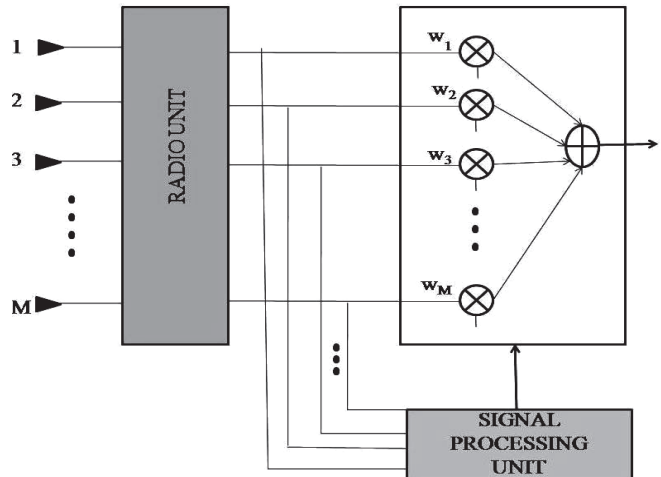


Fig.1. Smart Antenna Receiver

angular spread. Circular array has a bird view with an element spacing of $\Delta\phi = 2\pi/M$ and can perform beam forming in all the azimuth angles without any transcends on angular spread. 2- Dimensional and 3- Dimensional grid array are desirable for indoor or dense urban environments. It can perform beam forming in the directions, azimuth and elevation angles. 2-D grid has an element spacing of Δx in horizontal and Δy in vertical plane. 3-D cubic structure has element spacing of Δx , Δy and Δz .

Radio unit consists of down conversion unit and Analog to digital convertors each for one of the M element in an antenna array.

In signal processing unit, each incoming signal from M element of the array is multiplied by the weights $W_1, \dots, W_M$. The weights are calculated depending upon the employment of switched beam or adaptive array antenna system. Theoretically we can completely nullify the interference sources but practically due to multipath propagation, it is not possible to complete remove but can be decreased. These weights will decide the radiation pattern in the uplink. If the Switched Beam is used, then all the weights vectors are examined and one of them is selected giving the strongest received signal level. If adaptive smart antenna is used, the beam is directed towards the strongest signal to achieve maximum gain. Therefore first Direction of Arrival is calculated and then the weights are calculated accordingly. When the beam forming is done digitally, the beam forming and signal processing parts are integrated in the same unit.

For maximization of SIR (for AA), the optimum weight vector W_{opt} can be calculated using number of algorithms [1].

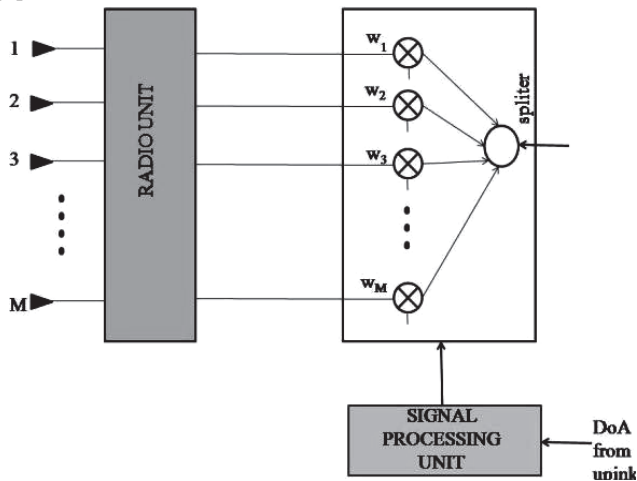


Fig.2. Smart Antenna Transmitter

In transmitter, the incoming signal is first split into M branches which are weighted by the complex weights $W_1, W_2, \dots, W_M$ in the beam forming units. The weights that decide the radiation pattern in downlink is calculated by signal processing unit. The radio unit consists of up convertor chains and digital to analog convertor. Practically antennas and digital signal processing is similar as on receiver. The main difference between uplink and downlink is no availability of spatial channel response on downlink. In TDD system, the same carrier frequency only separated in time is utilized by the base station and mobile. Thus the weights utilized on uplink are deployed on downlink, if the channel is not changed i.e. if user does not moves in high speed. In FDP, the channel response depends on frequency; the optimal weights are not same. As optimum beam forming is typical on downlink, the technique utilized is the geometrical approach of estimating the DOA. This technique

is based on the concept of directional reciprocity i.e. it is assumed that the direction of received signal on uplink is the direction in which signal will be transmitted on downlink. Thus the DOA is estimated from which the main part of the user signal is received. This direction is utilized on downlink for calculation of weight so that the radiation pattern is directed towards the desired user similar to the phased array systems. For reducing the interference, zeros are positioned towards the other users as the base station as base station interference is observed by mobile station and interference for mobile station is observed by base station. For sub-optimum system, we can average the uplink channel over a period of time to choose the downlink direction.

Conclusion

Smart Antenna is the most promising technique for using the space as a domain to increase coverage, capacity and signal strength. It suppresses the interference by operating highly directive beams towards the desired user. Many algorithms are deployed along with the digital processing technique to enhance the system performance.

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Study of Design and Simulation of a Double S - Shaped Metamaterial in the Microwave Range

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Abstract

Metamaterials (MTMs) are artificial EM materials that have unusual properties not available in nature. When their EM structures are effectively homogeneous, these MTMs behave like real materials and exhibit the effective, macroscopic constitutive Parameters: permittivity ϵ and permeability μ . However there are also some new metamaterial structures being explored which exhibit the same behaviour. In this paper, we present the design and simulation of a double S - shaped metamaterial in the microwave range. Simulation has been done by HFSS software. The extraction of effective parameters by the method of reflection- transmission coefficients demonstrates the metamaterial behavior of the said structure.

Keywords: Metamaterials, artificial electromagnetic materials, permittivity, permeability and HFSS software.

Introduction

The metamaterials present an emergent and promising research area that states to bring up important scientific advances in several areas. Some studies delimit the metamaterials to artificially structured periodic media in which the periodicity is much smaller than the electrical wavelength. This concept of metamaterials is directly related to the artificial dielectric work made on microwave frequencies. In 1968, the Russian scientist V. Veselago suggested a new type of material which has simultaneously negative permittivity and negative permeability, and he presented general properties of electromagnetic wave propagation in such material.

The theoretically created a lossless meta - material and showed the extraordinary properties of this material which is not found in nature, in particular a negative refraction, a negative group speed, the reversal of Doppler Effect and Cerenkov radiation. The Veselago's intuition remained silent for 29 years until year 1996 when Prof J.B Pendry proposed his design of Thin - Wire (TW) structure that exhibits the negative value of permittivity ϵ and the Split Ring Resonator (SRR) with a negative value of permeability μ in 1999. Later, Smith and his colleagues demonstrated a new metamaterial that shows simultaneously negative permittivity and permeability and carried out microwave experiments to test its unusual properties in 2000. The first experiment showing negative refraction was performed using a metamaterial consisting of a two dimensional array of repeated unit cells of copper strips and split ring resonators in 2001 by Shelby et al. Most of the MMs are made of periodically arranged

metallic structures much smaller than the working wavelength in size, the electromagnetic responses of MMs can be characterized as homogeneous media. As a new type of artificial materials, it can be applied to many disciplines, such as filter, waveguide, resonator and antenna. Several appointments appeared since the synthesis of such a medium; left hand medium, media with negative refractive index, backward - wave which wants to say medium where the wave moves behind, DNG (double negative materials) and meta - material.

The meta - material is thus an assembly of two structures, one which has a negative permeability (SRR) and the other which has a negative permittivity (TW).

Nevertheless, several work showed the possibility of synthesis of a similar medium by the means of various forms, in particular the omega form, the U form, the V form and the triangular form.

In this paper we present a structure which has the characteristic to have at the same time a negative permeability and a negative permittivity. By the means of the commercial software HFSS which uses the finite elements as calculation method, and in normal incidence, the « S » parameters for a single unit cell is calculated with the mentioned boundaries along the wave propagation, and by inversion technique of the Fresnel coefficients, the effective material parameters are given. From the simulation results, the real part of the refractive index is found to be negative at frequencies where both real parts of the permittivity and permeability are negative. Thus we show that there is a

frequency band where the effective refractive index of the medium is negative.

Geometry of the Proposed Structure and Mathematical Derivation

A theory of homogenization is used to obtain the effective electromagnetic parameters of the structure. The main purpose of this theory is to describe in a simple and macroscopic way the microscopic complexity of the response of objects to an incident electromagnetic radiation.

Indeed, the idea was to model the metamaterial as an isotropic homogeneous slab, and to calculate the effective parameters ϵ , and μ of the homogenous slab from the transmission and reflection coefficients obtained by simulations under HFSS software.

For an isotropic homogeneous slab in a vacuum space, the transmission t , and reflection r have the following relations with the refractive index n and the impedance z , of the slab

$$t^{-1} = \left[\cos(nkd) - \frac{i}{2} \left(z + \frac{1}{z} \right) \sin(nkd) \right] \quad \text{..... (1)}$$

$$\frac{r}{t} = \left[-\frac{i}{2} \left(z - \frac{1}{z} \right) \sin(nkd) \right] \quad \text{..... (2)}$$

Where k and d are the wave vector and the thickness of the slab respectively.

Equation (1) and (2) can be reversed to calculate n and z from t and r , we get

$$z = \pm \sqrt{\frac{(1+r)^2 - t^2}{(1-r)^2 - t^2}} \quad \text{..... (3)}$$

$$\cos nkd = \frac{1}{2t} (1 + t^2 - r^2) \quad \text{..... (4)}$$

By using the S parameter, the effective material can be extracted

$$z = \pm \sqrt{\frac{(1+S_{11})^2 - S_{21}^2}{(1-S_{11})^2 - S_{21}^2}} \quad \text{..... (5)}$$

$$re(n) = \pm re \left[\frac{\cos^{-1} \left(\frac{1}{2S_{21}} (1 - S_{11}^2 + S_{21}^2) \right)}{kd} \right] \quad \text{..... (6)}$$

$$im(n) = \pm im \left[\frac{\cos^{-1} \left(\frac{1}{2S_{21}} (1 - S_{11}^2 + S_{21}^2) \right)}{kd} \right] \quad \text{..... (7)}$$

The ambiguity on the signs of the equations (5), (6) and (7) is prevented if account is held owing to the fact that the real part of the impedance is positive if it is about a passive medium, and the imaginary part of the refractive index is positive to ensure that the incident wave amplitude decreases inside the structure. Then, the effective permittivity and permeability can be computed from the equations: $\epsilon = n / z$ and $\mu = n$.

Proposed Structure Geometry and Simulation

In this paper, structure "S" made up of two microrubans lines laid out on the two opposite faces of a dielectric substrate which has a permittivity $\epsilon = 3.38$, and a thickness of 1.6 mm, the periodicity in space is 3.63 mm.

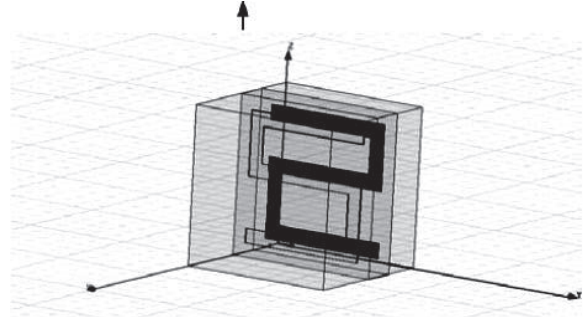


Figure 1: Proposed structure

The unit cell dimensions are presented in figure 1. Simulation is done by HFSS software. Boundary conditions of magnetic and electric wall are applied respectively according to axes y and x . The structure is polarized so that the magnetic field is directed along x axis, the electric field is directed along y axis, the wave vector is according to the z axis. The Simulation is made on a frequency band between 10 GHz and 20 GHz with a 0.01 GHz increment with HFSS software.

The simulated S parameters (coefficients of reflection transmission) for this structure are shown in figure 2.

As shown in figure 2, there is a transmission peak around 15.68 GHz, which indicates the existence of a resonance frequency which is due to the capacitive effect created by the geometry of the structure.

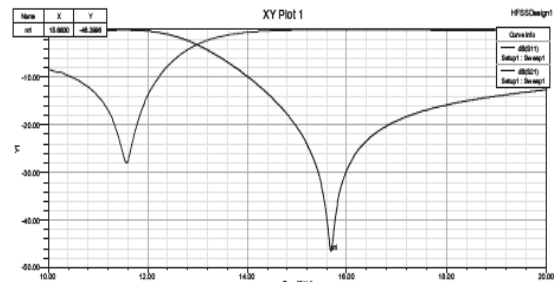


Figure 2: Amplitude of one unit cell

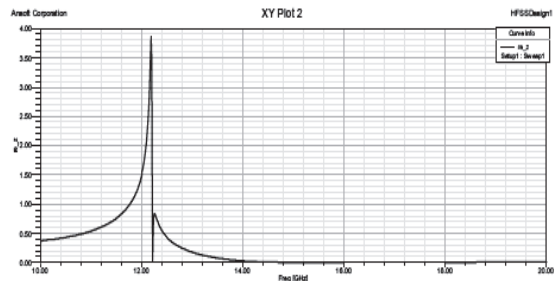


Figure 3: Impedance

So, we have an LC resonator who has a resonance frequency which depends only on the inductance and capacitance of the equivalent structure where $\omega = \frac{1}{\sqrt{LC}}$. The variation of the effective impedance of the wave propagating through the structure is illustrated in figure 3.

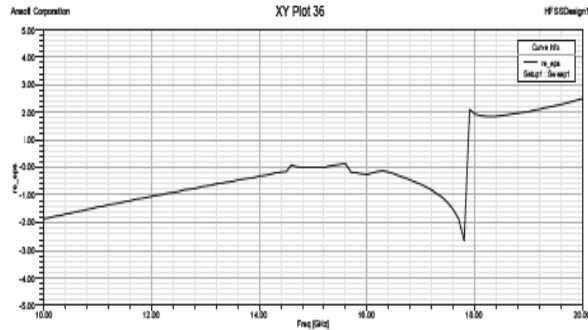


Figure 4: Evolution of the real part of the permeability

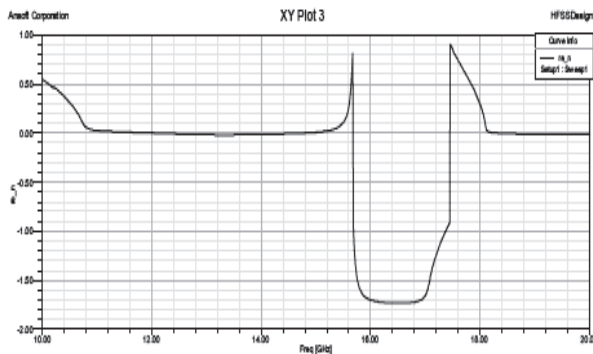


Figure 5: Evolution of the real part of the permittivity

A positive real part, ensuring that the medium is passive is verified. According to the theory of metamaterials, the real parts of the permittivity and permeability must be negative, thus negative real parts for the permittivity and permeability are observed for this structure. As shown in figure 4, the real part of the permeability shows Lorentz response behaviour, it is negative in the frequency range between 15.63 GHz and 17.8 GHz. In addition, the permittivity real part is negative in the frequency range located before 17.8 GHz, which is in good agreement with the Drude model as it appears in the figure 5.

The negative band for the refractive index approximately lies between 15.67 GHz and 17.43 GHz; it is the frequency range where the permittivity and the permeability are simultaneously negative. It is possible to control the transmission by modifying a certain number and opening of the gap, nature substrate and its thickness. Thus it is shown that the synthesis of a medium with negative refractive index is possible with similar geometry without having recourse to the array of metallic wires as it is the case for traditional metamaterials.

Conclusion

In this paper, we presented a particular meta - material structure shaped "double S". Its characteristic comes owing to the fact that it exhibits at the same time electric and magnetic activities because of its character bianisotropic. Such a geometrical configuration enabled us to highlight a frequency band where the structure behaves like a medium with negative refractive index.

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Study of a New Photonic Band Gap Substrate for Multi Frequency Microstrip Slot Antenna

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Abstract

Photonic band gaps are an excellent manifestation of the phenomena that arise from the interaction of the radiation field and matter. Photonic band gap materials (PBG) are artificially fabricated periodic structures, which give rise to allowed and forbidden regions for the propagation of light. These photonic band gap materials are also commonly known as photonic crystals (PCs). Based on the C-shaped microstrip slot antenna, a new photonic band gap substrate for multi frequency microstrip slot antenna is designed. Comparing with the C-shaped microstrip slot antenna, the resonant frequency of the antenna was reduced by 230 MHz, and the low frequency bandwidth was increased from 12.63 % to 18.95 %, both the radiation and multi-frequency characteristics of this proposed antenna are improved. The result demonstrates that the structure is efficient in improving the antenna gain and radiation directivity by suppressing surface wave of the microstrip antenna.

Keywords: Photonic band gap structure, C-shaped slot microstrip antenna, Surface wave and frequency gain.

Introduction

While active research into PBG structures developed only in the last decade, the theory predicting the existence of PBGs will be familiar to researchers in fields as diverse as crystallography, optics, and radome design. This is because the PBG problem is a matter of solving the macroscopic Maxwell's equations for wave propagation in periodic media. Indeed, the first hypothesis describing PBG behavior referred to an "electromagnetic band gap" rather than a photonic bad gap. In 1987, a laser researcher at Bell Communications Research (Bellcore) suggested that an artificial analog of a natural crystal structure might function as an omnidirectional spatial filter for EM radiation (Yablonovitch, 1987). Since then, researchers from such disparate backgrounds as quantum physics, computational electromagnetics, materials science, and electrical engineering have come to collaborate in the development of a growing subject known as photonic bad gap, or PBG, engineering. Simply defined, a PBG crystal is a man - made object with a periodic structure such that it prevents the propagation of certain EM modes in all possible directions within its bulk. It can be dielectric, metallic, or a mix of both. Properly designed, it can have the same degree and type of controllability for EM wave propagation that a doped semiconductor has for electronic propagation. Such a structure could clearly have a host of uses. Any application requiring a broad, uniform, omnidirectional frequency stop band is a candidate application for a PBG structure (PBGs). In order to control surface waves and improve the antenna

gain and radiation efficiency, lots of PBG microstrip antenna structure is used.

These studies remains a lots of issue, such as how to significantly improve the antenna performance indicators, reduce the antenna volume, weight and reduce the costs of processing the antenna.

Aiming at the above problems, also in order to improve the radiation and multi-frequency characteristics of the C - shaped slot microstrip antenna, this paper presents a PBG structure based on small - scale multi - band antennas, proposed new type of PBG microstrip antenna structure is introduction by different highly on substrate with the periodic cylindrical air holes in the triangle. The antenna is not only maintained a low profile microstrip antenna, simple structure, easy to feed the advantages, but also work with multi - band characteristics, in particular to ensure its excellent radiation performance of broadband RF communication will have a good prospect.

Simulation Model - Frequency Microstrip Antenna Design

This paper designed a working band is 2.0 GHz ~ 5.0 GHz's new multi-frequency slot antenna, the specific indicators are: centre frequency of the work are 2.40 GHz, 3.70 GHz, 4.50 GHz, when the return wave loss is less than -10 dB, the antenna relative bandwidth greater than 3%. Which 2.4 GHz is the common country's ISM band, the current WLAN, Zig Bee, Bluetooth and other wireless networks are operating in this band. To achieve multi-band antenna design, this single

point through the coaxial probe feed, use an ordinary rectangular patch in the open C-slot patch structure and design a new type of multi - frequency microstrip antenna. The idea is: add - C radiation plate gap, the gap divides the rectangular patch, the current, thereby increasing the current path, the path of the double bending method of the current multi-band effect.

Because the centre frequency of the traditional microstrip antenna can be achieved by changing the antenna length the width of slot change can improve the reflection coefficient and port standing wave ratio; width W the size of the antenna directivity function, input impedance, radiation impedance related. In general, the width of the higher efficiency, wider bandwidth, but the radiant panel is easy to produce high-order mode, caused by field distortion.

After the above analysis, the final design of microstrip antenna size: radiant panel size; ground plate size; C - type gap width is 1 mm. Designed planar microstrip antenna shown in Fig. 1.

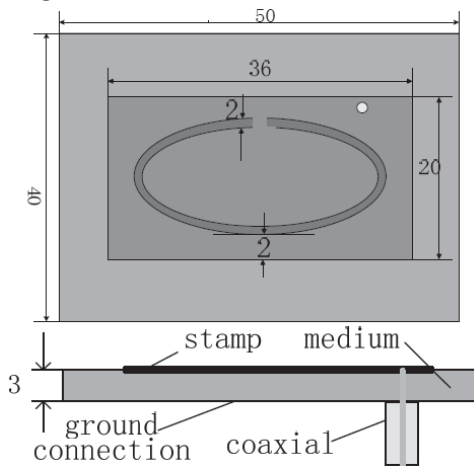


Figure 1: View of the Antenna

For simulation, we use electromagnetic simulation software HFSS on the C - slot microstrip antenna. Antenna return loss graph and the radiation pattern graph are shown in Fig. 2 and Fig. 3.

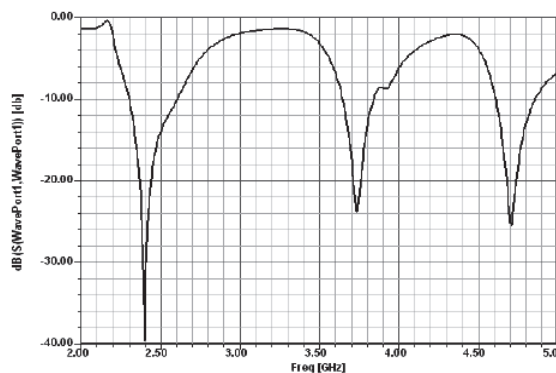


Figure 2: C-type slot microstrip antenna return loss

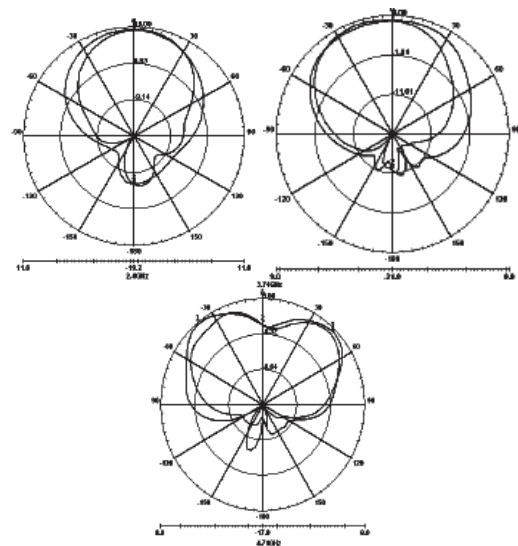


Figure 3(a), 3(b) and 3 (c): C-type slot antenna radiation patterns microstrip

It can be seen from Fig. 2, the general C-slot microstrip antenna with multi-frequency performance, the resonance frequencies are 2.40 GHz, 3.73 GHz and 4.71 GHz. When the return wave loss less than -10dB, the antenna bandwidth are 310 MHz (2300 MHz ~ 2610 MHz), 210 MHz (3640 MHz ~ 3850 MHz) and 250 MHz (4610 MHz ~ 4860 MHz), the relative bandwidth respectively are 12.63 %, 5.61 % and 5.28 %. At the same time, can be seen from Fig. 3, when the frequency is 2.4 GHz, the antenna's main radiation direction of the antenna in the -45~40, 00 when the maximum forward gain is 10.101 dB, the horizontal direction, a gain of -5.777 dB, antenna at this frequency before and after the ratio is 16.134 dB; Antenna resonant frequency is 3.74 GHz at the time of the main radiation direction - 45~45, when the antenna has maximum gain decrease of about 7.682 dB, the front ratio about 19.574 dB; and when the frequency attains to 4.71 GHz, the antenna main lobe split occurred, the antenna's main radiation direction ± 45 . And the antenna to the direction of radiation and the radiation level is large, the antenna radiation performance is poor, it takes the antenna structural improvements.

Structure Design of PBG

Microstrip patch antenna is not only a radiator, but also a resonator. To avoid the area increasing, we can use the PBG of the band - stop characteristics make the antenna operating frequency of the harmonic generation, thereby reducing the surface waves and inhibit basal energy loss and improve the antenna radiation efficiency. PBG structure currently used mainly in drilling the base type, ground corrosion, high

impedance surface structure and so on, we use the base type of PBG structure drilling program. As the PBG structure will be directly applied to the dielectric substrate of the microstrip antennas, this will not increase the size of the antenna.

The traditional type of base drilling, that is, the media surrounding the antenna radiation plate substrate to introduce a two - dimensional PBG structure, usually in the substrate can be used in a number of periodic holes drilled to implement the structure. The properties of photonic band gap crystals with grid constant a , filling factor r/a and dielectric constant and other factors. TM photonic crystals can map to proper selection of the band gap r/a value, so that the antenna operating frequency in the band gap of photonic crystal structures within the substrate to suppress the surface wave antenna to improve the overall performance of the antenna. Gonzalo base through regular periodic air holes drilled in the structure, making the hole - type photonic crystal substrate patch antenna. Gonzalo, who proved that the design of surface wave suppression is very obvious, and weakened the antenna side lobe back flap, increasing the gain of the forward radiation, but this design likely to cause the base size is too large, the antenna weight increase, so this antenna applications in miniaturization difficult to promote.

Compared with the traditional base drilling PBG structure, this paper proposed a PBG structure, which mainly is drill the dielectric substrate that under the PBG antenna radiation board, in order to form the triangular periodic distribution of cylindrical air holes, and the radiant panel grooves dug beneath the dielectric layer, so that including the introduction of photonic crystal defect, and the antenna patch on the bottom of the groove, through the rational design cycle grid size and pore size to obtain the wide band gap, and make the design of the antenna operating frequency falls we calculate photonic crystal within the band, this design in addition to surface wave suppression, can also be seen from Fig. 4, the electromagnetic wave radiation to space the band gap of the photonic crystals will also be the presence of inhibition, so this structure can better focus the radiation, allowing the radiation of space than the patch antenna on the photonic crystal surface on the microstrip antenna to be small. Based on the above analysis and calculations, we get the result, which is when the triangle side length of cylindrical air gap is 4.33 mm, horizontal distance between the two air band's distances is 10.75 mm, the vertical spacing of 8.25 mm, the antenna's overall performance is good. Again simulate the antenna with HFSS, PBG structures can be the antenna return loss of microstrip antenna radiation pattern diagram and are shown in Fig. 5 and Fig. 6. It can be seen from Fig. 5, the introduction of the C-gap PBG structure the resonant

frequency of the antenna are 2.17 GHz, 2.40 GHz, 3.63 GHz and 4.62 GHz, respectively of its -10 dB bandwidth of 450 MHz (2150 MHz ~ 2600 MHz), 140 MHz (3580 MHz ~ 3720 MHz) and 110 MHz (4580 MHz ~ 4690 MHz). It is thus clear, as the PBG structure was added to the C-type slot antenna, the patch antenna return loss has significantly improved.

Besides, the low frequency bandwidth was increased to 18.95 %, comparing to 12.63 % of the common C type slot antenna and the antenna bandwidth increase by 6.32 %, all of these results achieve the broadband performance of the antenna; at the same time, the resonant frequency of the antenna was reduced by 230 MHz in order to achieve miniaturization of the antenna miniaturization performance.

Fig. 6 shows the radiation pattern of PBG structure antenna at the resonant frequency. It can be seen from Fig. 6, that after the introduction of PBG structure, the antenna radiation in the work on the consistency of a good band, with and without the introduction of PBG structure compared to the antenna radiation characteristics, antenna gain of the first to have improved. When the operating frequency of 2.17 GHz, the antenna forward gain achieves to 10.382 dB, then the horizontal gain of -20.360 dB, the antenna can be seen in the low frequency side lobe has been greatly suppressed, and the backward radiation of the antenna is quite small, the front - to back ratio is around 18.774 dB; antenna at the resonant frequency of 2.40 GHz, it has the coincident radiation pattern with the ordinary slot antenna at the same resonance frequency, in the mean time, the forward gain antenna was increased to 10.533 dB, the front-to-back ratio is about 14.650 dB. Seen after the introduction of PBG structure, the gain of microstrip antennas have a more significant improvement, because the media had a surface wave PBG structure on the inhibition, so that power is no longer part of the strengthening of the back - propagation to the front, thereby reducing the side lobe and back lobe radiation, and enhance the main direction of radiation intensity. However, in the high frequency, the antenna's resonant frequency has decreased, when the frequency is 3.63 GHz, the antenna at -300 when the maximum antenna gain prior to 9.011 dB, and the front-to-back ratio of 22.697, but when the frequency of high frequency 4.62 GHz, the main beam splitting of microstrip antenna which not loaded C-shaped slot can be improved, then the antenna's main radiation direction of -600 to 600, -350 when the maximum antenna gain is 7.810 dB, the ratio is around 16.511 dB. PBG can be seen in the high frequency antenna gain and directivity of radiation have been effective in improving the same time, and after the radiation has also been reduced. Although the whole of the PBG

antenna proved effective in the inhibition of surface waves, increase the gain directional antenna radiation and improve the effectiveness and other properties, but the high frequency antenna in the direction of the radiation level has not been inhibited. This is mainly around the PBG structure of radiant panel caused by periodic small. Therefore, to inhibit the substrate in all directions in the surface wave is necessary to introduce the ideal three dimensional PBG structure.

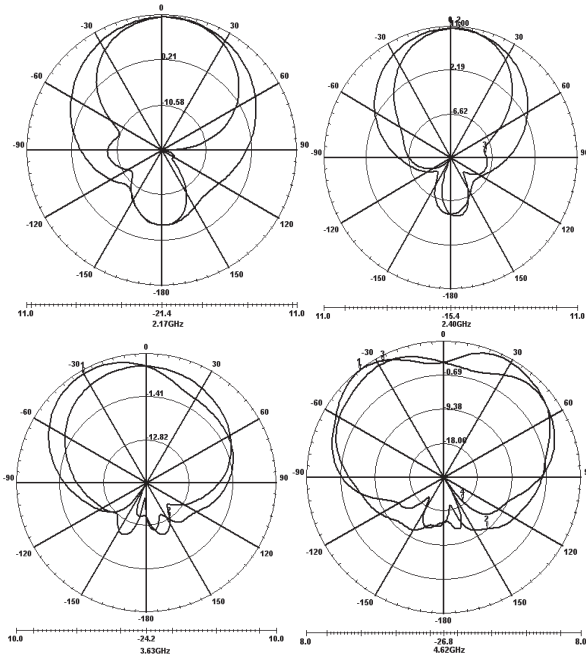


Figure 4. Figure 5. Figure 6. and Figure 7.

Conclusion

This paper presents a new base drilling photonic crystal substrate patch antenna. Simulation results show that the performance of the PBG structure which we designed is not only to suppress surface wave propagation, to a certain extent, an increasing of the antenna gain, but also increasing the antenna bandwidth in the low frequency by 6.32 %. In addition, the designed PBG structure size relative to the

traditional drill-based PBG structure the base, such as reference is smaller, more use of the antenna to achieve miniaturization of photonic crystals. The simple design of the antenna structure, good performance, in the wireless mobile communication system has a good prospect. However, we also note that the high frequency antenna work in a narrow band, and the poor performance of high frequency, so the next step we should focus on produce antenna model, and on how to improve the high frequency of the antenna radiation performance's issue get in depth study.

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Applications and Characteristics of Various III -V Compound Semiconductor Alloys

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Abstract

The III –V semiconductor alloys are promising candidates for many device applications such as high - speed electronic and long wavelength photonic devices because their band gaps cover a wide spectral range. III -V compound semiconductors have played a crucial role in the development of optoelectronic devices for a broad range of applications. Major applications of InP or GaAs based III -V compound SC are devices for optical fiber communications, infrared and visible LEDs / LDs and high efficiency solar cells. GaN based compounds are extremely important for short wavelength light emitters used in solid state lighting systems. In this paper we discussed the important device applications and characteristics of various III -V compound SC materials.

Keywords: III – V semiconductor, optical fibre, light emitting diode (LEDs), solar cell and laser.

Introduction

Semiconductor alloys provide a natural means of tuning the magnitude of the forbidden gap and other material parameters so as to optimize and widen the applications of semiconductor devices. With the advent of small-structure systems, such as quantum wells and super lattices, the effects of alloy compositions, size, device geometry, doping and controlled lattice strain can be combined to achieve maximum tenability.

The III–V semiconductor alloys are promising candidates for many device applications such as high-speed electronic and long wavelength photonic devices because their band gaps cover a wide spectral range. Among these materials, the (Al,Ga)(As,Sb) ternary alloys are covering wavelengths between the visible and infrared region. $\text{Al}_x\text{Ga}_{1-x}\text{As}$ is the most important and the most studied III–V semiconductor alloy, its key role in a variety of transistor and optoelectronic devices has necessitated a precise knowledge of the fundamental energy gap as well as the alignment of the three main conduction-band valleys. Semiconductor devices have become an integral part of our lives. Si is the prime elemental semiconductor that has revolutionized the microelectronics industry. However, Si is an indirect bandgap semiconductor and thus an inefficient light emitter/absorber. Optoelectronic devices involve interaction between photons and electrons. Direct bandgap semiconductors that show strong absorption/emission characteristics are best suited for such applications Figure 1. InGaAsP alloys lattice matched to InP

are widely used for fabricating devices (lasers, modulators, semiconductor optical amplifiers) for long haul fiber communications as they have bandgap energies in the low dispersion, low loss window for the optical fibers. However, poor temperature characteristics of InP based lasers have prompted research in the direction of fabrication of long wavelength ($\sim 1.3 \mu\text{m}$) lasers based on GaAs and its alloys. Another important application of GaAs based compound semiconductors is for infrared emission and detection using intersubband transitions.

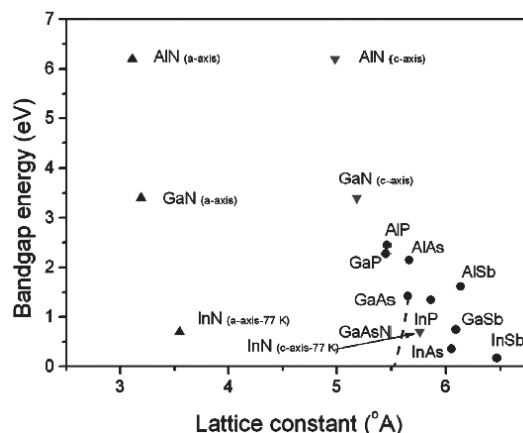


Figure 1: Bandgap energy (300 K) vs. lattice constant for III-V compound semiconductors

The bandgap of GaAs is near optimal for achieving maximum possible theoretical efficiency for conversion of sunlight into electricity using a single p-n junction. Another important class of semiconductors is GaN based compounds

that have been used for short wavelength emitters for solid state lighting applications. In this paper we present an overview of various InP based semiconductor materials studied for applications in lasers, modulators and semiconductor optical amplifiers, GaAs based material systems studied for applications in long wavelength emitters, infrared detectors and high efficiency solar cells and GaN based materials studied for emitters in the blue region of the visible spectrum.

Light Emitting Diodes and Lasers

One of the most important applications of laser diodes is as sources for telecommunications. The optical fibers have minima in attenuation at wavelengths of 850 nm, 1.31 μm and 1.55 μm . The 1.55 μm windows has a maximum transmission length of ~ 100 km and is used for long haul communications, whereas the 850 nm and 1.31 μm windows are used for short distance data communications. InGaAsP / InP material system is currently used for fabricating 1.3 μm lasers for telecommunications. However, the major disadvantages of lasers fabricated from this material system arise due to small bandgap offsets between the InGaAsP quantum wells (QW) and InP barriers (0.39 eV), lack of lattice matched layers with InP with high mismatch in the refractive index that can be used for effective optical confinement and poor thermal conductivity of the InGaAsP layers.

Very small bandgap offsets between the QWs and the barrier results in poor temperature characteristics of these lasers (To ~ 60 K). Using InAlAs barriers increases the bandgap offset between the QWs and the barriers and results in improved temperature characteristics. In spite of the improvement in the T_0 , the InP based diode lasers require thermoelectric cooling. The above factors also result in very high fabrication and operation costs of such lasers. These disadvantages can be alleviated by using GaAs based emitters emitting at 1.31 μm . The GaAs based lasers use lattice matched AlGaAs layers for better carrier (higher bandgap offsets) and optical (higher refractive index mismatch) confinements. Materials that offer the possibility of long wavelength emission on GaAs substrates are:

- In(Ga)As
- GaAsSb and
- InGaAsN

InGaAs Quantum Wells (QWs)

The lattice mismatch between InGaAs and GaAs limits the In content and/or thickness of the QWs so that intense photoluminescence (PL) signal can only be obtained at or below ~ 1.1 μm . However, lasing at ~ 1.2 μm was

demonstrated by employing lower growth temperatures and/or strain compensating buffer or capping layers for the QW active region. Use of a thin In_{0.24}Ga_{0.76}As strain compensating layer just below the active region in the GaAs buffer layer or GaAs_{0.9}P_{0.1} capping layers instead of highly strained GaAs layers⁶ allowed increasing the emission wavelength of the QWs that is usually limited by critical thickness. Even though the long wavelength GaAs based lasers have better characteristic temperatures compared to InP based lasers, they have higher transparency currents and lower quantum efficiencies compared to 0.98 μm InGaAs QW lasers, possibly due to current leakage and scattering through crystal imperfections associated with low temperature growth of highly strained structures. There are no reports to date, to the best of our knowledge, on strained InGaAs QW lasers with lasing wavelength beyond ~ 1.26 μm . The advantages of using QDs in the laser active region arise from three dimensional confinements of charge carriers that lead to delta function like density of states. QD lasers are expected to exhibit lower threshold current densities and better temperature stability than QW lasers.

Lasing in QD structures was first demonstrated in 1947. These lasers used a single InGaAs QD layer as the active region. At low temperature ($T < 100$ K), these devices exhibited lasing from QD states, whereas at room temperature, the lasing shifted to the wetting layer (WL) states, accompanied with an increase in threshold current density.

These problems were associated with

- Gain saturation in the active region and
- Thermal escape of carriers from QD to the WL and barrier states.

Gain saturation is a consequence of lower surface coverage and broad size distribution, inherent to self assembled QDs. Gain saturation issues in QD lasers can be addressed by increasing the gain volume in the active region that is achieved by stacking several QD layers together. Lasers with multiple stacks of InGaAs QDs were fabricated to increase the net maximum gain obtainable from the active region.

The lasers with stacked QD active region demonstrated QD ground state lasing even at room temperature and improved temperature characteristics.

(In)GaAsN

Incorporation of N and In into the GaAs lattice results in lowering the bandgap energy of GaAs, while minimizing the strain as In tends to increase while N tends to reduce the lattice constant of GaAs. This provides the possibility of lattice matching InGaAsN layers to GaAs.

Emission wavelengths of InGaAsN layers can be tailored between 1.2 μm and 1.6 μm by varying the In and N content. For 1.3 μm emission, In composition of $\sim 35\%$ and N composition of $\sim 1\%$ are routinely used. In compositions in excess of 35% lead to very high compressive strain and N composition beyond 1% deteriorates the crystal quality due to large miscibility gap between GaAs and GaN and increases the non - radiative recombination in the QWs. (In)GaAsN lasers were first demonstrated in 1996. However, InGaAsN QW lasers exhibit higher threshold currents and lower characteristic temperatures, T_0 , compared to InGaAs QW lasers.

Quantum Cascade Lasers

Unlike the interband transitions used for lasing in conventional devices, inter-sub -band transitions (similar to the transitions used for IR detection using GaAs based materials) in relatively wide bandgap semiconductors can be used to achieve lasing at lower energies (in the mid and far IR region of the electromagnetic spectrum using GaAs and at 1.3 μm and 1.5 μm using GaN based materials). GaAs based unipolar devices (the quantum cascade lasers) find applications in IR spectroscopy, point to point atmospheric communication and optical radars and GaN based unipolar devices may be suitable for optical telecommunications. Intersubband transitions result in weak temperature dependence compared to interband transition lasers and the laser performance is not limited by Auger recombination typical of IR lasers based on interband transitions. First GaAs based quantum cascade lasers were demonstrated in 1998. GaAs - AlGaAs material system is widely used for quantum cascade lasers based on GaAs. Large conduction band offsets are desirable in the active region of the quantum cascade lasers that suppress thermally activated processes which limit device performance at high temperatures and also allow increasing the energy separation among the subbands.

GaN based Light Emitting Devices

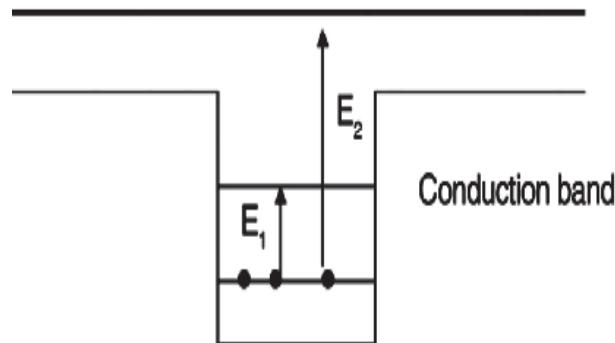
Short wavelength light emitting devices are necessary for applications like full color displays, laser printers, read- write laser sources for high density information storage and light emitting diodes for solid state lighting. GaN related materials ((AlInGa) N) have bandgap energy that can be tuned in the blue - green region of the visible spectrum and UV - region. Bulk InGaN doped with Si and Zn was used as the active region for demonstration of blue LED. The luminescence in InGaN / AlGaN double heterostructure LEDs was due to impurity assisted recombination. The longest peak

electroluminescence wavelength demonstrated in InGaN / AlGaN double heterostructure LEDs is ~ 500 nm.

InGaN quantum confined active regions (QW and QD) are used for more efficient emission in the blue - green region of the visible spectrum. GaN or AlIn(Ga)N are usually used as barriers for InGaN QW or QD light emitting devices in accordance with the emitting wavelength of interest. Typical InGaN QW devices exhibit reduced efficiencies at higher injection currents due to carrier escape from the active region. This is a serious problem with high power lasers that typically operate at high injection densities. Use of AlGaN/GaN multi quantum well barriers as carrier blocking layers has resulted in improved slope efficiency and reduced threshold currents in InGaN devices.

Infrared Photodetectors

HgCdTe (MCT) is a small bandgap material commonly used in conventional photodetectors based on interband absorption for detection of infrared radiation. The difficulties associated with growth and processing of MCT motivated the study of intersubband photodetectors based on larger bandgap III-V compound semiconductors. QWs or QDs embedded in larger bandgap materials are used to create bound energy states within the valence or conduction bands of the device (Fig. 2). The device active region is doped to provide free charge carriers and the infrared detection is based on excitation of carriers from occupied energy levels in the conduction/valence band to unoccupied levels in the same band. AlGaAs-GaAs is the standard III-V semiconductor material for quantum well infrared photodetectors (QWIPs). The spectral response for detectors based on this material system can be tuned between 6-20 μm . Even though p-type QWIPs are sensitive to normal incidence radiation, they exhibit very weak absorption. To overcome this disadvantage, QWIPs fabricated from indirect bandgap n-type semiconductors have been investigated.



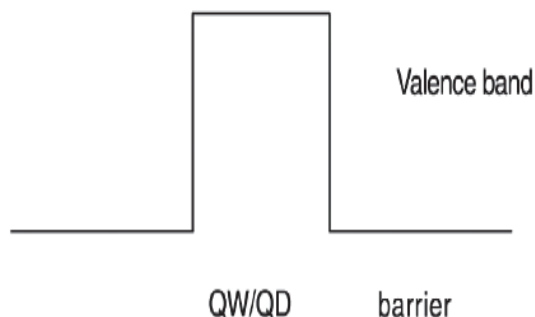


Figure 2: Schematic showing the typical transitions used for infrared detection in QWIPs/QDIPs

InGaAs / GaAs

Compared to the GaAs / AlGaAs system, the binary barriers in the InGaAs / GaAs QW system have superior carrier transport properties. In addition, n - type doping of the InGaAs / GaAs long wavelength QWIPs enhances the absorption of TE modes. Grating free n - type normal incidence QWIPs for 8-10 μm applications have been fabricated that exhibit similar responsivity compared to QWIPs with 1D gratings. These QWIPs were operated at 77 K with detectivity of $2 \times 10^{10} \text{ cm Hz}^{1/2}/\text{W}$ and responsivity of 0.23 A / W. Broadband detection in the 7-14 μm range was demonstrated using p - doped InGaAs / GaAs QWIPs.

Quantum Dot Infrared Photodetectors (QDIPs)

Inter subband photodetectors based on QDs offer the following advantages over QWIPs:

- In QWIPs, selection rules prohibit the absorption of normal-incidence radiation. QDIPs have intrinsic sensitivity to normal incidence light.
- QDIPs present a potential for increase in signal to noise ratio compared to QWIPs.
- QDIPs enable higher temperature operation.
- QDIPs have higher responsivity due to longer lifetime of excited electrons due to the reduced recombination rate associated with a multi photon relaxation step - 'phonon bottleneck'.
- QDIPs have a broader IR response range than the QWIPs because the QDs have several discrete states and an inhomogeneous size distribution.

Most of the QDIPs fabricated utilize vertical transport through stacked self-assembled QDs. Because of inhomogeneity in size and spatial alignment of the self - assembled QDs, the tunneling transport through stacks of self - assembled QDs is strongly affected by the disorder. So it is difficult to obtain high photoconductive gains in the vertical transport structures. Increased photoconductive gain has been demonstrated in QDIPs by utilizing high mobility doped

AlGaAs / GaAs layers for the transport of photo excited carriers.

Modulators

Modulation is the process of superimposing a low frequency signal onto a high frequency carrier wave. Modulating signals at several GHz provides capacity to use the enormous bandwidth of optical fibers. Its other advantages include noise reduction and multiplexing.

In external modulators, the signal from the source (LED/laser) is used to modify the bias conditions of the modulators, resulting in a high frequency output encoded with the input signal. External modulation avoids the problems associated with chirp in high speed ($> 1 \text{ G bit / s}$) direct modulation of lasers. Only external modulators are discussed in this paper. The external modulators can be broadly classified as electro - absorption type or electro - optic type.

Electro-absorption Modulators

Electro-absorption is the process of absorption of photons with energy lower than the bandgap energy of the semiconductor in presence of an external electric field. It can broadly be explained in terms of Frank-Keldysh effect, which is a photon assisted electron tunneling process and the Stark effect, which is the splitting of atomic energy levels in the presence of an external field for bulk semiconductors. The strong confinement of charge carriers in quantum wells results in much stronger electro-absorption effect than in bulk materials.

Strong localization of electrons and holes in the quantum well increases the overlap of their wavefunctions and hence the transition probabilities between these bound states. As a result, the absorption spectra of quantum wells show strong resonances due to excitonic transitions near the bandedge, unlike bulk material. In the presence of an external electric field, the electron and hole wavefunctions are pushed in opposite directions as shown in Figure 3.

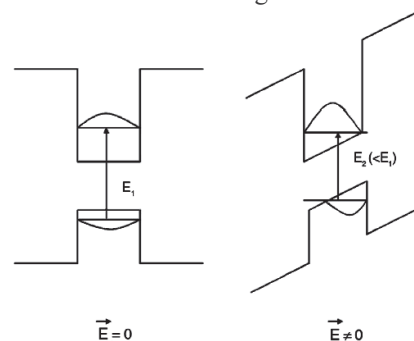


Figure 3: Schematic illustrating the quantum confined stark effect in quantum wells

The reduced overlap in the electron and hole wavefunctions results in reduced absorption due to the excitonic transitions. The probability of carrier tunneling from the quantum wells increases, resulting in reduced carrier life times and thus a broadened absorption spectrum.

The ground state transition energy is reduced due to the modification of envelope functions and the absorption spectrum shifts to lower energies. This energy shift is much larger than the shift produced due to Franz-Keldysh effect or Stark effect and occurs at relatively weaker external electric fields.

Electro-optic Modulators

Electro-optic effect is defined as change in the refractive index of a material due to an external electric field. If the change in refractive index depends linearly on the applied electric field (as in bulk GaAs / InP), the effect is called Pockel's effect, whereas if the refractive index exhibits quadratic dependence on the applied field, the effect is called Kerr effect (as in quantum confined structures).

Two material systems of interest for long wavelength applications are InGaAs / InAlAs and InGaAsn / InP. The conduction band offset in the InGaAs / InAlAs system (0.5 eV) is larger than that in the InGaAs / InP system (0.2 eV). Thus the quantization effects should be more pronounced in the former system.

Semiconductor Optical Amplifiers

The principal application of semiconductor optical amplifiers (SOAs) for optical fiber communications include:

- Booster amplifier to increase the transmitter power
- In-line amplification to compensate for transmission losses
- Pre-amplifier to improve the receiver sensitivity

In addition, semiconductor optical amplifiers can be simultaneously used for other functionalities such as modulation, wavelength conversion, switching, photodetection and can easily be integrated with other active components. It is necessary to achieve polarization insensitive amplification at 1310 nm and 1500 nm for above applications.

InGaAsP/InP and InAlGaAs/InP are quaternary material systems that can be bandgap engineered for 1.3 and 1.5 μm applications and several functionalities have been demonstrated using SOAs based on these active regions. The optical bandwidth and gain dynamics of the active region of the semiconductor optical amplifiers determine the spectral region over which the semiconductor optical amplifier operates and its switching rates, respectively. QD active regions with ultrafast gain dynamics can be used to fabricate low noise SOAs with large bandwidths for high speed

applications. The optical bandwidth and gain dynamics of the active region of the semiconductor optical amplifiers determine the spectral region over which the semiconductor optical amplifier operates and its switching rates, respectively. QD active regions with ultrafast gain dynamics can be used to fabricate low noise SOAs with large bandwidths for high speed applications.

Solar Cells

Solar cells operate on the principle of photovoltaic effect. Sunlight incident on the semiconductor material creates electron-hole pairs that are collected in an external circuit. Two important parameters that give an indication of the efficiency of a solar cell are its short circuit current density and the open circuit voltage. The maximum short circuit current density and open circuit voltage obtainable from any semiconductor material are both functions of the bandgap energy of the material. As the bandgap decreases, the maximum possible short circuit current density increases, as more photons in the solar spectrum are now capable of creating electron hole pairs in the material. However, the maximum obtainable open circuit voltage decreases with decreasing bandgap energy of the semiconductor.

Thus there exists optimum bandgap energy for the semiconductor material in the range 1.4 to 1.6 eV to maximize the solar cell efficiency.

In a single junction device, the photons of energy lower than the bandgap energy of the semiconductor are not absorbed, while the excess energy in photons with energy greater than the bandgap energy of the semiconductor is lost as heat. A more efficient conversion of sunlight into electricity can be achieved by employing multi-junction solar cells, where different bandgap semiconductors are used for absorption in different energies of the solar spectrum. The bandgap and the thickness of each junction should be optimized for highest efficiencies. Until the late 90s, the use of III -V compound semiconductor photovoltaic devices was confined only to space applications (due to light weight, greater efficiency, better radiation hardness, higher electron mobility important for high frequency operation and higher temperature stability) since their price is much higher than the conventional Si solar panels. But the demonstration of high efficiency multi-junction solar cells has generated lots of interest in III -V based solar cells for high concentration terrestrial applications.

Conclusion

Material systems of interest for GaAs based emitters at $\sim 1.3 \mu\text{m}$ include In(Ga)As, GaAsSb and InGaAsN. Lasing close to $1.3 \mu\text{m}$ has been demonstrated in both In (Ga)As QW and

QD systems. Even though 1.3 μm lasing has been demonstrated in GaAsSb / GaAs QW lasers, the lasers have high threshold current densities and low internal quantum efficiencies as a result of type - II bandgap alignment between the GaAsSb layers and the GaAs barriers. For InGaAsN/GaAs QW laser structures emitting at 1.3 μm , hole leakage (due to a small valence band offset) results in high threshold current densities. Use of larger bandgap GaAsP or GaAsN barriers reduces hole leakage and improves the temperature characteristics of the lasers. (In)GaAs - (Al)GaAs based alloys are used for mid and far IR emitters based on inter subband transitions. AlInGaAs is the material of choice for fabricating high luminescence blue-green laser diodes and LEDs. AlGaAs - GaAs is the standard material. InP based materials used for long wavelength light emission, modulation and amplification applications include InGaAs / InAlAs, InGaAs / InP and InGaAsP / InP, InAlGaAs / InP, respectively for infrared photodetectors based on inter - subband transitions, with response in the range of 6 - 20 μm . GaAs based multi - junction / intermediate band structures offer the possibility of high efficiency III - V solar cells for concentrator systems.

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Simulation of High Frequency Wave Scattering from Anisotropic Dielectric Media

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Abstract

A mathematical technique for modeling the behavior of high frequency wave scattered from a random anisotropic media is presented. The random media is characterized by a thin dielectric layer. A pair of second order boundary conditions is used to simulate an anisotropic thin dielectric layer as an infinitesimally thin sheet. The basis of the formulation is the Fourier Integral transform technique. The field terms are evaluated with the help of Wiener-Hoff solutions obtained by standard method. The uniqueness of the solution is satisfied by imposing an edge constraint.

Keywords: Anisotropic media, dielectric layers, Fourier integral transformation and edge diffraction.

Introduction

The increasing demand on high data rates in modern communication systems has pushed the operating high frequencies wave regime. Due to absorbing characteristics of dielectrics, scattering of such structures has of great importance in electromagnetic communities. Also the scattering by dielectric strip yields the GTD analysis.

Edge diffraction is important because the main contributions to coherent high frequency scattering are local phenomena associated with specular points at surfaces and edges or object to object interactions within a scene. A pair of boundary conditions was introduced by Idemen to simulate a special type of anisotropic dielectric layer that is homogeneous and isotropic with respect to the directions parallel and perpendicular to the plane, while the half-space lying beneath and above the layer are homogeneous and isotropic media as shown in Fig.-1.

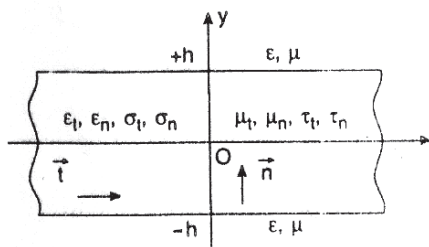


Fig. 1.

The boundary conditions on an infinitesimally thin anisotropic dielectric sheet placed in free-space are as follows :

$$\vec{n} \times (\vec{E}^+ - \vec{E}^-) = A (\vec{E}_1^+ + \vec{E}_1^-) + B \vec{n} \times \text{grad} (H_n^+ + H_n^-) \quad \dots\dots (1)$$

and

$$\vec{n} \times (\vec{H}^+ - \vec{H}^-) = C (\vec{H}_1^+ + \vec{H}_1^-) + D \vec{n} \times \text{grad} (E_n^+ + E_n^-) \quad \dots (2)$$

For a layer of thickness $2h$ with $\vec{n}$ denoting the unit vector normal to the layer; and also it is assumed that $2h < \lambda$, where λ is the wavelength. The constant coefficients in Eqs. (1) and (2) are

$$\begin{aligned} A &= -i\omega(\epsilon'_t - \epsilon_0) \\ B &= \frac{-h(\mu'_n - \mu_0)}{\mu'_n} \\ C &= i\omega(\mu'_t - \mu_0) \\ D &= \frac{-h(\epsilon'_n - \epsilon_0)}{\epsilon'_n} \end{aligned} \quad \dots(3)$$

Here the scattering of high frequency waves by an anisotropic dielectric plane is investigated. Simulation is done with the help of second order boundary condition as in equations (1) and (2). These have the same mathematical form with higher order boundary conditions involving second order derivatives [6].

Theoretical Basis

An H- polarized plane wave incident in the plane perpendicular to the edge is considered as in Fig.-2.

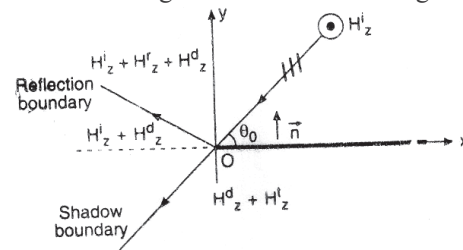


Figure 2

The boundary value problem involving an infinitesimally thin half-plane with anisotropic dielectric layer is imposed as

$$H_z^i(x, y) = \exp \left\{ -ik \left(x \cos \theta_0 + y \sin \theta_0 \right) \right\} \quad \dots(4)$$

Where $0 \leq \theta_0 \leq \pi$. The time dependence is assumed as $\exp(-i\omega t)$ and suppressed throughout the analysis. For the sake of analytical convenience, it is assumed that the surrounding medium is slightly lossy and consequently the wave number has a small positive imaginary part.

The boundary conditions given by Equ. (1) and (2) are written as follows for the H-polarized case:

$$\left[H_z^+ - H_z^- \right] + \frac{1}{ik\gamma} \frac{\partial}{\partial y} \left[H_z^+ - H_z^- \right] = 0, \quad x > 0 \quad \dots(5)$$

$$\frac{\partial}{\partial \gamma} \left[H_z^+ - H_z^- \right] + \frac{1}{ik\gamma(\gamma_1 + \gamma_2)} \left[\frac{\partial^2}{\partial \gamma^2} - k^2(\gamma_1 \gamma_2) \right] \left[H_z^+ - H_z^- \right] = 0, \quad x > 0 \quad \dots(6)$$

The equations (5) and (6) in the forms of first and second order generalized transition conditions.

The total field is written as the sum of the incident and scattered fields for all y

$$H_z(x, y) = H_z^i(x, y) + H_z^s(x, y) \quad \dots(7)$$

and the following integral representation is introduced for the scattered field:

$$H_z^s(x, y) = \begin{cases} \int_L M(\alpha) e^{i\Gamma(\alpha)y - i\alpha x} d\alpha, & y > 0 \\ \int_L N(\alpha) e^{i\Gamma(\alpha)y - i\alpha x} d\alpha, & y < 0 \end{cases} \quad \dots(8)$$

Which satisfies the radiation condition for $|y| \rightarrow \infty$.

Using equations (4) and (8) the boundary conditions for $x > 0$ and for $x < 0$ and inverting the resulting integral equations, the following set of equations are found

$$\frac{\Gamma(\alpha)}{k} \left[\frac{k}{\Gamma(\alpha)} + \frac{1}{\Gamma} \right] [M(\alpha) - N(\alpha)] = \Phi_1^-(\alpha) - \frac{\sin \theta_0}{i\pi\gamma(\alpha - k \cos \theta_0)} \quad \dots(9)$$

$$\left[\Gamma(\alpha) \frac{\Gamma^2(\alpha)}{k(\gamma_1 + \gamma_2)} + \frac{k(\gamma_1 \gamma_2)}{(\gamma_1 + \gamma_2)} \right] [M(\alpha) - N(\alpha)] = \Phi_1^-(\alpha) - \frac{k(\gamma_1 \gamma_2 + \sin^2 \theta_0)}{i\pi(\gamma_1 + \gamma_2)(\alpha - k \cos \theta_0)} \quad \dots(10)$$

$$M(\alpha) - N(\alpha) = \Phi_1^+(\alpha) \quad \dots(11)$$

$$\Gamma(\alpha) [M(\alpha) + N(\alpha)] = \Phi_2^+(\alpha) \quad \dots(12)$$

The functions $\Phi_1^\pm(\alpha)$ and $\Phi_2^\pm(\alpha)$ are unknown functions where $\Phi_{1,2}^+(\alpha)$ are regular in the upper complex, α -plane for $\text{Im}(\alpha) > \text{Im}(k \cos \theta_0)$ and $\Phi_{1,2}^-(\alpha)$ are regular in the lower complex α -plane for $\text{Im}(\alpha) < \text{Im}(k)$.

Scattered Field Estimation

The scattering field analysis based on the integrals in Eq. (8), require to determine the spectral coefficients $M(\alpha)$ and $N(\alpha)$, which can easily be obtained from Eqs. (11) and (12).

$$M(\alpha) = \frac{1}{2} \left[\Phi_1^+(\alpha) + \frac{\Phi_2^+(\alpha)}{\Gamma(\alpha)} \right] \quad \text{and}$$

$$N(\alpha) = -\frac{1}{2} \left[\Phi_1^+(\alpha) - \frac{\Phi_2^+(\alpha)}{\Gamma(\alpha)} \right] \quad \dots(13)$$

Substituting $M(\alpha)$ and $N(\alpha)$ expressions into Eq. (8), the scattering field integrals can be found as

$$H_z^s(r, \theta) = + \frac{k}{2\pi i} \int_L \left\{ \frac{K_1^+(\alpha) K_1^-(k \cos \theta_0)}{\Gamma^+(\alpha) \Gamma^-(k \cos \theta_0)} \frac{\sin \theta_0}{\gamma(\alpha - k \cos \theta_0)} + \frac{K_2^+(\alpha) K_2^-(k \cos \theta_0)}{\Gamma^+(\alpha) \Gamma^-(k \cos \theta_0)} \right. \\ \left. - \frac{1}{\gamma_1 \gamma_2} \left\{ - \frac{k(\gamma_1 \gamma_2 + \sin^2 \theta_0)}{(\alpha - k \cos \theta_0)} + C(\gamma_1 + \gamma_2) \right\} \right\} e^{\pm i\Gamma(\alpha)y - i\alpha x} d\alpha \quad \dots(14)$$

Where the upper and the lower signs in Eq. (14) correspond to the integrals for regions $y > 0$ and $y < 0$, respectively. By applying the standard saddle-point method, the diffracted field expressions are obtained as

$$H_z^d(r, \theta, \theta_0) = \frac{D^d(\theta, \theta_0) e^{ikr}}{\sqrt{kr}} \quad \dots(15)$$

With

$$D^d(\theta, \theta_0) = \frac{e^{ikr}}{\sqrt{2\pi}} \left\{ \frac{1}{\gamma_1 \gamma_2} \frac{K_2^-(k \cos \theta_0) K_2^-(k \cos \theta)}{\sqrt{1 - \cos \theta} \sqrt{1 - \cos \theta_0}} \left\{ - \frac{\gamma_1 \gamma_2 + \sin^2 \theta_0}{(\cos \theta + \cos \theta_0)} + (\gamma_1 + \gamma_2) C \right\} \right. \\ \left. + \frac{K_1^-(k \cos \theta_0) K_1^-(k \cos \theta)}{\sqrt{1 - \cos \theta} \sqrt{1 - \cos \theta_0}} \frac{AZ_0 \sin \theta_0 \sin \theta}{(\cos \theta + \cos \theta_0)} \right\} \quad \dots(16)$$

The reflection and transmission coefficients are

$$T^{(r)} = - \left\{ \frac{AZ_0 \sin \theta_0}{1 + AZ_0 \sin \theta_0} - \frac{ikD \cos^2 \theta_0}{\sin \theta_0 + ikD \cos^2 \theta_0} \right\} \quad .. (17)$$

and

$$T^{(t)} = \frac{1}{1 + AZ_0 \sin \theta_0} - \frac{ikD \cos^2 \theta_0}{\sin \theta_0 + ikD \cos^2 \theta_0} \quad .. (18)$$

Numerical Results

Some numerical examples for the nonmagnetic case under H polarization is obtained for reflection, transmission and diffraction coefficients given as $20 \log_{10} [T^{(r)}]$, $20 \log_{10} [T^{(t)}]$ and $20 \log_{10} [D^{(d)}]$, respectively. The diffraction

coefficients involve in the split functions $K_1^{\pm}(\alpha)$ and $K_2^{\pm}(\alpha)$ which are written in terms of the Maliuzhinetz's function. By using the approximate formulas given by Volakis and Senior [7], Maliuzhinetz's function is expressed in [8].

The diffracted field coefficient is computed for different values of the constitutive parameters and for different thickness of the anisotropic layer.

First investigation is accomplished with respect to the thickness of the anisotropic dielectric layer in order to reveal the effectiveness of the simulation. From Fig.-3 and 4 it can

be deduced that the optimum values of $\frac{2h}{\lambda}$ should be chosen between 0.05 and 0.15 to simulate the layer effectively with anisotropic dielectric layer boundary conditions.

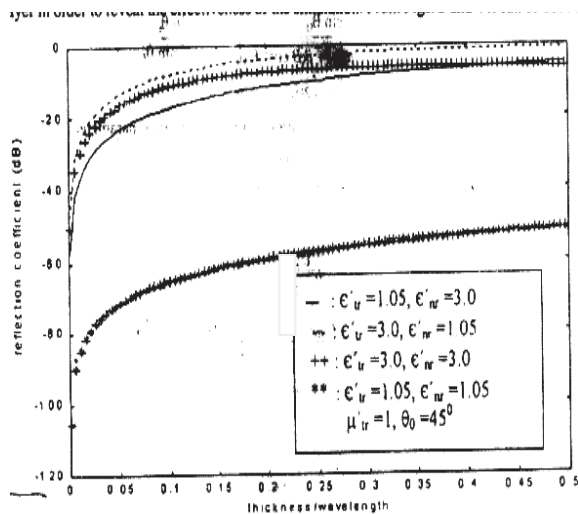


Figure 3

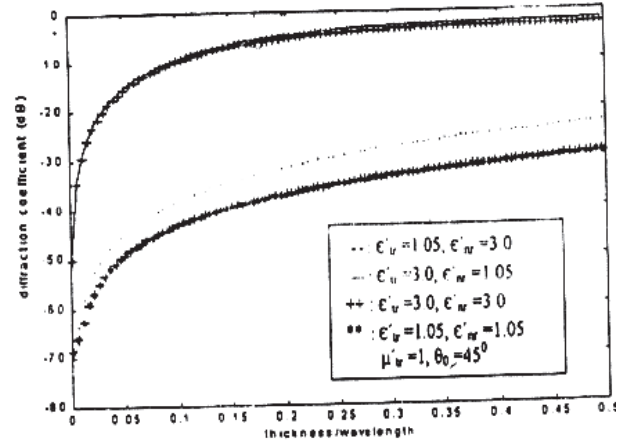


Figure 4

Also, it is seen that reflection and diffraction mechanism are enhanced for $\text{Re}(\epsilon'_{ir}) > \text{Re}(\epsilon'_{nr})$ cases, which means that the effect of $\text{Re}(\epsilon'_{ir})$ is dominant.

Conclusion

In this paper, the problem of an anisotropic thin dielectric layer illuminated by an H - polarised plane wave has been solved and a uniform expression for the edge diffraction co-efficient has been derived. The solution was accomplished by using a pair of boundary conditions to model an anisotropic thin dielectric layer as an infinitesimally thin sheet. This leads to a pair of uncoupled Wiener - Hoff equations whose solution can be expressed in terms of Maliuzhinetz's function. The uniqueness of the solution is achieved by using an edge constraint. The results are compared with the previously obtained ones for $\epsilon'_{ir} = \epsilon'_{nr}$ and $\mu'_{ir} = \mu'_{nr}$ and it has been shown that they are in good agreement.

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Study of Some Metamaterial - based Sensors and its Characteristics

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Abstract

Metamaterials, artificial composite structures with exotic material properties, have emerged as a new frontier of science involving physics, material science, engineering and chemistry. Metamaterials are artificial media structured on a size scale smaller than wavelength of external stimuli, and they can exhibit a strong localization and enhancement of fields, which may provide novel tools to significantly enhance the sensitivity and resolution of sensors, and open new degrees of freedom in sensing design aspect. In this paper mainly we presents the recent progress in metamaterials - based sensing,

Keywords: Metamaterials, sensitivity, resolution of sensor, biosensor and metamaterial sensor.

Introduction

Metamaterials are artificially made electromagnetic materials consisting of periodically arranged metallic elements which are less than wavelength of incident electromagnetic (EM) wave in size.

Moreover, the materials can manipulate electromagnetic wave beams in surprising ways and exhibit some exotic electromagnetic properties which are not readily available in Nature, such as backward propagation, reverse Doppler effect, reverse Vavilov - Cerenkov effect, negative refraction, diffraction - limit breaking imaging, cloaking, etc. These exotic properties strongly depend on the geometry of metamaterial molecules rather than their composition. Since they were theoretically proposed by Pendry and experimentally demonstrated by Smith metamaterials have attracted intensive research interest from microwave engineers and physicists in recent years because of their wide applications in super-lenses, slow light, data storage, optical switching. So far, researches into fabrication, design, and application of metamaterials have been extended to a fairly wide range of the EM spectrum including far-, mid-, and near-infrared regimes and even optical frequencies. metamaterials can exhibit a strong localization and enhancement of fields so that they can be used to actually improve the sensor selectivity of detecting nonlinear substances and to enable detection of extremely small amounts of analytes. Based on this property, many new or improved applications of metamaterials have been proposed recently. Currently, the metamaterial science has reached a high degree of sophistication. As a result of continuous progress in design and fabrication of metamaterials on the submicron and nanometer scales, novel

enhanced properties of metamaterials can be created. Thus, the interdisciplinary boundary between metamaterials science and sensing technology has become a fertile ground for new scientific and technological development. Moreover, the considerable advances have been made on research into sensors based on metamaterials, and they have led to the development of various metamaterial-based components to detect the information of substance and circumstance.

It is well known that sensing devices can detect a small change, depending on the following four criteria: first, the sensors must have an operating frequency low enough to avoid the background and substrate absorption. This poses a significant challenge as conventional sensing devices have a limited area, and such reduced space tends to increase the operating frequency of sensors. Therefore, it is required to maintain a small layout of the sensor while decreasing its operating frequency as much as possible. Second, the sensors must produce a strong and measurable readout signal with a resonant behavior sharp enough to accurately track the shift in transmission spectra. The third criterion pertains to the linearity of sensing which is related to the quality factor of sensors. The fourth criterion is the sensor sensitivity. If there are a limited number of data points in one frequency scan of the network analyzer, it is easier to resolve smaller shifts in the transmission spectra in response to the externally applied load when the sensitivity is higher. The purpose of this paper is to illustrate the performance of metamaterial - based sensors.

Biosensor

Biosensors are essential in many areas, such as disease diagnostics, environmental monitoring, and food safety, and they are also vital tools in the investigation of biological

phenomena. Fluorescence-based methods have proven useful in analyzing both genomic and proteomic microarrays and in imaging, including single molecule detection inside living cells. However, labeling molecules with fluorophores can be expensive and time-consuming. It may even be infeasible for certain applications.

In addition, biological reactions generally rely on the three-dimensional structure of the biomolecules which may be affected by the addition of a fluorescent marker. Therefore, there has been a drive for bioanalytical sensing techniques that can directly detect the target molecules without labeling. Recently, biosensing technologies based on metamaterials have attracted significant attentions from the microwave to optical frequency because of their cost-efficient and label free biomolecule detection. According to operating frequency of sensing biomolecule and component, the metamaterial-based sensors are classified into three types: microwave biosensor, terahertz biosensor and plasmonic biosensor.

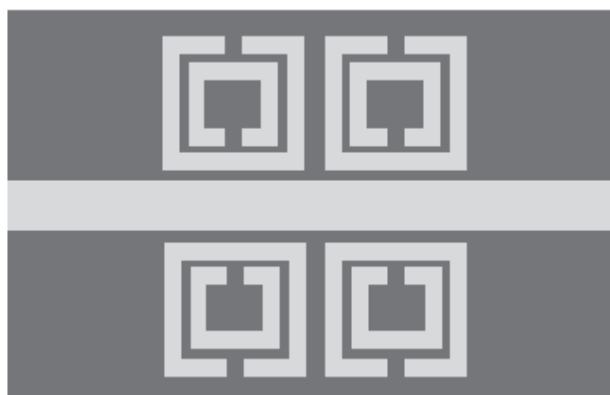


Figure 1(a)

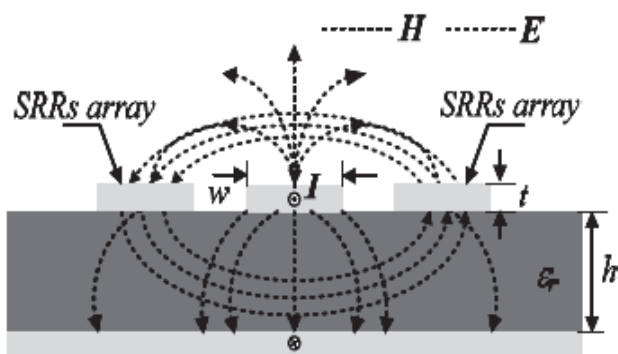


Figure 1(b)

Generally, a biosensor can be defined as a device incorporating a biological sensing element related to a transducer. It consisted of the sensitive biological element, transducer or the detector element, and associated electronics

or signal processors. To study the sensitivity and selectivity characteristics of SRR-based biosensors, the biosensor surface was coated with gold (Au) and the single-stranded deoxyribonucleic acid (ss-DNA)-linked biotin was used for immobilization on the Au surface because of good chemical attraction between ss-DNA and Au. After immobilization, binding of biotin and streptavidin was achieved by a bioprocess because the biotin-streptavidin binding is a well-known affinity, as shown in Figure 2.

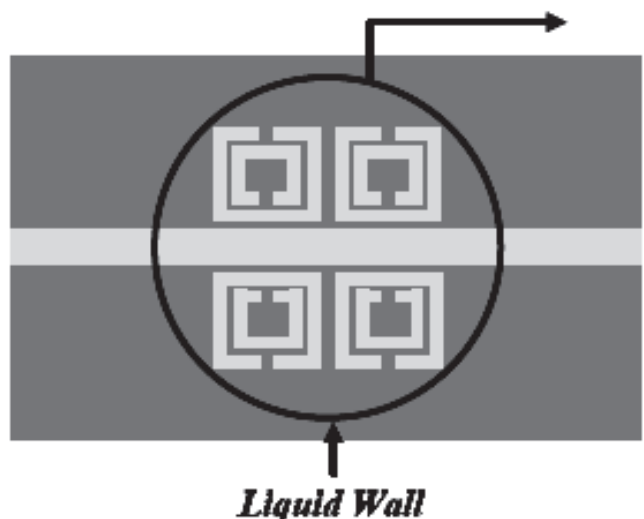


Figure 2(a)

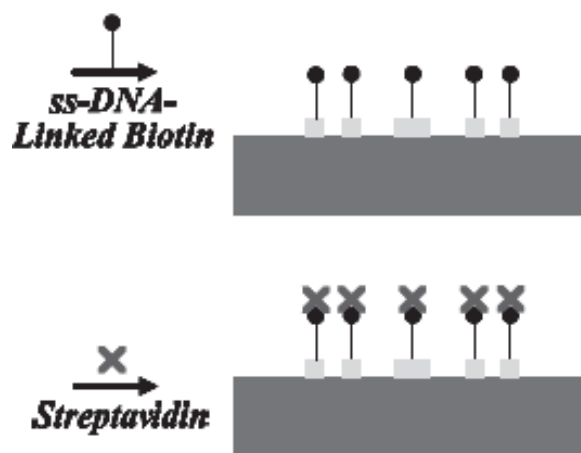


Figure 2

In the absence of biomaterials, the resonant frequency of the SRR-based biosensor was 10.82 GHz. When biotin was introduced, the resonant frequency shifted to 10.70 GHz. The change of frequency in this case was $\Delta f_B = 120$ MHz. After the binding of biotin and streptavidin, the frequency shifted to 10.66 GHz. The change of frequency in this case was $\Delta f_{B-S} = 40$ MHz. These shifts in the resonant frequency stem

from the change in the capacitance due to the binding of biotin and streptavidin. Thus, the SRR-based biosensor can be directly used for label-free biomolecular detection. Even though biosensors based on the resonator array have exhibited a remarkable resonant frequency shift, the occupied areas of the devices were relatively large and their biomolecular concentration was relatively high ($\sim 1 \mu\text{g/mL}$).

Another key advantage of the metamaterial-based sensor consisted of the essentially discontinuous porous nanotexture of the nanorod matrix. This enabled the implementation of new sensing geometries and strategies, not feasible with conventional film-based SPR. Indeed, by functionalizing the nanorods and immobilizing a receptor on their surface, one can follow the binding of a selective analyte with the receptor inside the nanorod matrix. The considerably increased surface area given by the nanoporous texture of the metamaterial significantly increased the amount of biomaterial that can be incorporated into the matrix within the probe depth available, maximizing the biological sensitivity of the system.

Furthermore, the distance between the nanorods can be selected to match the size of biological species of interest, giving access to a further size selectivity option that was important for many tasks in the detection of immunoassays, virus and protein.

Challenges and Prospects

In recent years, applications of metamaterials in sensing provide novel opportunities for developing a new generation of sensing technologies. Metamaterials can improve the mechanical, optical and electromagnetic properties of sensors, therefore, the metamaterial-based sensors are developing towards single molecular biosensor and high throughput sensor arrays. However, like any emerging field, they also face many challenges. For example, the performance of sensors based on metamaterials is limited by fluctuation phenomena which may be caused by various external and internal mechanisms and result in noise appearing in the sensor readout. External noise sources include the quantum noise/shot noise, photodetector noise and noise of processing circuitry, while intrinsic noise is mainly connected with the adsorption and desorption (AD) of analyte particles on the metamaterial. Moreover, the AD noise remains the main limiting factor of the ultimate performance of all such devices, because each metal-dielectric structure must be exposed to some kind of environment and thus AD noise will influence to a certain degree the performance of any metamaterial structure, as well as that of conventional SPR sensors and nanoplasmonic

structures. To decrease the influence of AD noise, Djuric analyzed the effect of AD noise on the performance of microelectromechanical (MEMS) structure, and found that this noise was generated by instantaneous differences in the rates of adsorption and desorption of contaminant molecules from the resonator surface, because the analyte particles were adsorbed onto the MEMS resonator surface and/or desorbed from it, modifying its mass and causing the fluctuations of its resonant frequency. Jaksi analyzed the performance of metamaterial-based sensors which were limited by refractive index fluctuations caused by adsorption and desorption processes. To further avoid losses and scaling issues of metal-based structures, the purely dielectric configuration metamaterials instead of metal-structured metamaterials were proposed.

Conclusion

Metamaterial-based biosensors still need improvements in sensitivity and accuracy, mainly because of the difficulty in obtaining high-resolution, small feature sizes on substrates where conventional photolithography techniques are difficult to apply. This is mainly due to the inability to use conventional photolithography-based microfabrication techniques where chemical solutions are generally used. Thus, the advanced fabrication technology must be developed.

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A Study on IRON (III) Complexes with Schiff bases derived from Vanillin Derivatives

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Abstract

Iron (III) metal complexation with Schiff bases such as Vanillidineaminopyridine Schiff base, vanillidine o-toluidine Schiff (vots) base, vanillidine p-toluidine Schiff (vpts) base, Vanillidine m-nitro aniline Schiff (v-m-N-as) base and Vanillidine 2-aminophenol Schiff base (Vaps) have been prepared. These have been synthesized and characterized by elemental analysis IR, and electronic spectral studies complexes. Their physical properties like solubilities, melting points, magnetic properties and conductances have been measured.

Keywords: Vanillidine Schiff bases, IR and TLC

Introduction

Study on 3d metal Schiff base complexes is classical but yet most intensive and exhaustive area of research.⁶ Normally 3d metal (II) complexes are being taken for study and 3d metal (III) ion complexation studies are not much.⁷⁻⁸ Biological activity of metal complexes with 4-amino antipyrine Schiff bases of 5-chloro salicylaldehyde and 3-ethoxysalicylaldehyde have been studied by B. Anupurna & Ch. Venkata Reddy et al¹⁰ which have been opened a new area of research. Redox properties and DFT computations of [MH(CO)(PPh₃) N-hetro cycle Schiff base] and their applications to the synthesis of Ag-nano particles have also opened a new area for research regarding Schiff base complex.¹¹

Experimental

All chemicals such as Iron (III) chloride, ethanol, acetone, DMSO, DMF, 2-amino pyridine, 2-amino phenol, o-toluidine, p-toluidine, o-m-nitro aniline etc were of Anal-R grade and were purchased from Merck / Qualikems / CDH.

The purity of all the synthesized complexes was checked by performing TLC for single spot on silica gel-G plates. Their melting points were determined in open capillaries and are thus uncorrected carbon, hydrogen and nitrogen were carried out on Thomas CH analyzer 35 Carlo Erba-1106 and Coleman N-analyzer. Chloride were estimated gravimetrically. Metal content was graviometrically determined. IR spectra were recorded on Perkin-Elmer spectrophotometer modal RX - 1 using KBr - pellets. Electronic spectra were recorded on model ultra-scan UV 2600 double beam spectrometer in DMSO solution at room temperature. Dipole moments were determined by using

Gouy-balance. Conductances of the complexes were measured using digital conductivity meter model no 304 (systronics).

Synthesis of Vanillidine Schiff bases

The ethanolic solution of (a) 2-amino pyridine (b) o-toluidine (c) p-toluidine (d) m-nitro aniline and (e) p-aminophenol was gradually added to the ethanolic solution of Vanillin with vigorous stirring each separately. After refluxing over a water bath for about 3-4 hours, these were allowed to stand overnight. The pale yellow crystals separated out. The precipitate was filtered, washed and dried finally in a dessicator over anhydrous calcium chloride. The purity of the ligand was checked by thin layer chromatography. The r_f values was found 0.18 ± 0.03 .

Synthesis of Iron (III) complex

The ligand solution was added to the metal salt solution and refluxed for about three hours. It was cooled and sodium acetate solution was slowly added to maintain the pH of the solution till the precipitation was complete. The precipitated complex was filtered off. It was repeatedly washed with distilled water and ether and dried in a dessicator over anhydrous CaCl₂ where by coloured complex was obtained. It was found insoluble in water and alcohol but soluble in DMSO and DMF. Same process was repeated with other aromatic amines too which were component for other Vanillidine Schiff bases.

Table: 1

S. N.	Ligand	Mol. Formula	Colour	Calc / found		
				C	H	N
1	Vaps, L _a	C ₁₃ H ₁₂ N ₂ O ₂	Pale yellow	68.4 (68.05)	5.26 (4.98)	12.28 (11.97)
2	Vots, L _b	C ₁₅ H ₁₅ O ₂ N	Cream yellow	74.68 (74.05)	6.022 (5.96)	5.80 (5.15)
3	Vpts, L _c	C ₁₅ H ₁₅ O ₂ N	Cream yellow	74.68 (74.05)	6.22 (5.97)	5.80 (5.15)
4	Voaps, L _d	C ₁₃ H ₁₃ O ₂ N	Cream yellow	69.13 (68.05)	5.34 (5.12)	5.76 (5.15)
5	VmNas, L _d	C ₁₄ H ₁₂ O ₄ N ₂	Deep yellow	61.76 (60.87)	4.41 (4.03)	10.29 (9.85)

Table: 2

S N	Iron complex	Colour	Calc / found				
			C	H	N	Cl	Fe
1	[Fe(Vaps) ₂ (H ₂ O) ₂] Cl	Brick Red	54.01 (53.65)	4.01 (3.78)	9.89 (9.63)	6.21 (6.10)	9.01 (8.94)
2	[Fe(Vots) ₂ (H ₂ O) ₂] Cl	Deep Red	59.89 (59.26)	4.89 (4.61)	4.89 (4.61)	5.98 (5.84)	8.70 (8.50)
3	[Fe(Vpts) ₂ (H ₂ O) ₂] Cl	Brick Red	64.87 (64.82)	4.03 (3.98)	4.03 (3.98)	5.12 (5.05)	7.67 (7.39)
4	[Fe(Voaps) ₂ (H ₂ O) ₂] Cl	Deep Red	55.05 (54.95)	4.01 (3.92)	4.78 (4.58)	5.99 (5.81)	8.88 (8.50)
5	[Fe(VmNas) ₂ (H ₂ O) ₂] Cl	Reddish	50.31 (50.19)	3.38 (3.29)	8.43 (8.36)	5.41 (5.30)	7.80 (7.77)

Table: 3

S N	Iron complex	M. Pt.	Solubility in	μ_{eff} (in BM)	Conductance in $\Omega^{-1}\text{cm}^2\text{mol}^{-1}$
1	[Fe(Vaps) ₂ (H ₂ O) ₂] Cl	205	Ethanol, DMSO & DMF	5.76	76.6
2	[Fe(Vots) ₂ (H ₂ O) ₂] Cl	227		“	78.6
3	[Fe(Vpts) ₂ (H ₂ O) ₂] Cl	204		“	82.2
4	[Fe(Voaps) ₂ (H ₂ O) ₂] Cl	207		“	79.2
5	[Fe(VmNas) ₂ (H ₂ O) ₂] Cl	205		“	80.8

Table: 4
IR Spectral values

Compound	ν (OH)	ν (C=N)	ν (OCH ₃)	ν (ArH)	ν (M-N)	ν (M-O)	ν (H ₂ O)
A	3520	1560	1120	3075	580	430	840
B	3520	1560	“	3100	580	430	850
C	3555	1570	“	3100	580	440	835
D	3500	1530	“	3100	580	430	840
E	3535	1555	“	3100	580	440	810

Results and Discussion

The analytical results suggest the Fe (III) & ligand ratio is in 1:2. Complexes are coloured and stable in air. Mostly complexes are insoluble in water, benzene, acetone but soluble in ethanol, DMSO & DMF. The magnetic moment values of Fe (III) complexes are in good agreement with the presence of five unpaired d-electrons.¹² Electronic conductance of all the Iron complexes show electrolytic nature of complexes. Electronic spectral values of Iron (III) complexes and values of Racah parameter (β), nephelauxetic ratio β in complex / β in free ion calculated suggest appreciable amount of covalent character in metal ligand bond.

From IR values, ligands have been proved to be bidentate co-ordinating through phenoxide (O⁻) ion¹³ and N of C=N group of Schiff base.^{14, 15} Two water molecules are co-ordinated supporting octahedral structure of complexes.¹⁶ One Cl⁻ is present in outer sphere of the complex to neutralize remaining one positive charge on Fe (III) ion. On the basis of above results & discussion, following tentative structures of complexes are assigned:

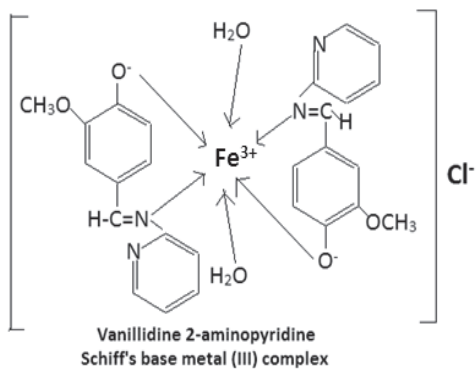


Figure: 1

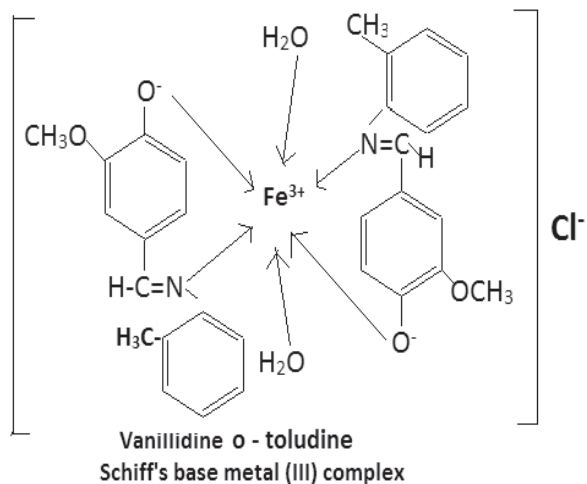


Figure: 2

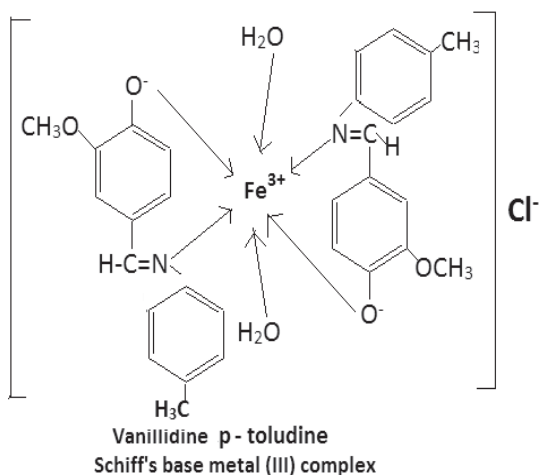


Figure - III

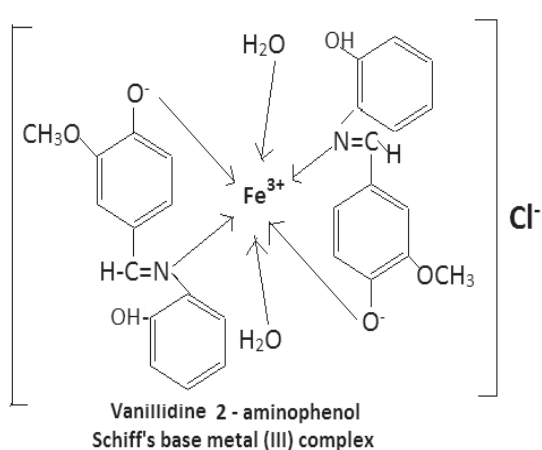


Figure - IV

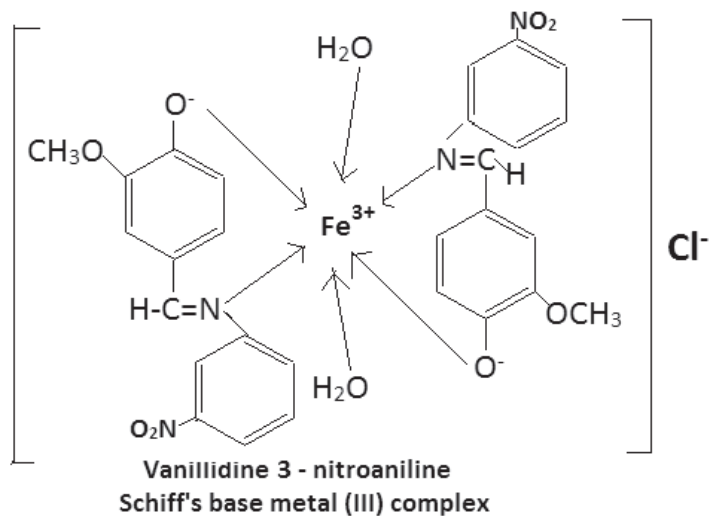


Figure - V

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Study of Recent Developments in Dye - Sensitized Solar Cell Applications

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Abstract

The worldwide demand for energy is expected to double by the year 2050 and triple by the end of the century. An abundant supply of energy is necessary for global political, economic and environmental stability. There is growing concern, however, that the production of oil will soon not be able to keep up with the growing demand leading to dire economic consequences. In addition, the combustion of fossil fuels has been implicated in anthropogenic global warming, which is predicted to produce widespread environmental damage. The development of carbon-free sources of energy that are scalable to meet increasing societal demands is therefore one of the major scientific challenges of this century, Light from the sun is arguably the ideal source of energy. While technologies have been developed to harness solar energy efficiently, they are not yet an economically viable alternative to fossil fuels. Dye-sensitized solar cells (DSSCs) have attracted much attention as they offer the possibility of extremely inexpensive and efficient solar energy conversion. we explore the potential of future generation DSSCs for advancing energy-conversion performance.

Keywords: Dye - sensitized solar cells, solar energy, photovoltaic device, cyanines and hemicyanines.

Introduction

The energy requirements and the concerns for environment have led to an increasing demand for development of energy sources that are cleaner and also abundant in nature. In this context, the most important one could be considered as the solar energy. The vast energy given by the sun could be harnessed in a suitable way so as to compensate for the world's energy needs. Therefore, the discovery of dye sensitized solar cells could be considered as a crucial step in this direction.

Dye-Sensitized Solar Cells (DSSCs), also known as Gratzel solar cells, were invented by Brian O'Regan and Michael Gratzel in 1991 (O'Regan and Gratzel 1991). These cells are the only photovoltaic devices that separate two varied functions, solar light harvesting and charge carrier transport, whereas conventional cells perform both of these operations simultaneously, consequentially requiring high purity materials like silicon with higher production costs. Solar cells of mesoscopic type are based on an interpenetrating network or bulk junctions and offer the benefits of providing flexible devices, of vital importance for particular applications. These cells are usually fabricated by low cost standard techniques like spin coating, dip coating and screen printing which in turn avoids the high-vacuum energy-intensive steps required for fabrication of other film based solar cells. A recent life cycle analysis has shown that, in

case of dye based cells, the carbon footprint is lower than that for other PV cells owing to which the energy payback time for the DSSC will be less than one year. In contrast to silicon cells, the materials required by DSSC are eco-friendly and abundantly available so that the technology can be scaled up to terawatt scale without running into feedstock supply problems. All these factors gives it an advantage over the two other major competing thin film photovoltaic (PV) cells made out of CdTe and CuIn(As)Se that use toxic materials of low abundance. However, the major challenge in commercial application of DSSC is reaching the high efficiency values as those of conventional cells (Nazeeruddin et al. 2005). The scientific progress in nanocrystalline or mesoscopic TiO₂-based DSSCs has come a long way since the advanced findings first reported in 1991. Dye-sensitized solar cells have become one of the most promising low-cost alternatives for the photovoltaic conversion of solar energy compared with the conventional solid p-n junction photovoltaic devices. This cell was invented by Michael Gratzel et. al. in 1991 and is also known as Gratzel cell.¹ Although commercially available solar cells are mainly based on silicon based inorganic cells, however dye based solar cells are highly promising and cost-effective alternative for the photo-voltaic energy sector. Currently, these flexible and tunable solar cells based on ruthenium(II)-polypyridyl complexes as key sensitizers have reached an overall power

conversion efficiency (η) greater than 11% under standard (Global Air Mass 1.5) illumination (Nazeeruddin et al. 2001; Gao et al. 2008; Nazeeruddin et al. 2005; Nazeeruddin et al. 2007). Besides transition-metal complexes, a variety of organic molecules have been explored, including coumarin (Hara et al. 2005) squaraine, indoline, hemicyanine (Chen et al. 2005) and other conjugated donor acceptor organic dyes (Hara et al. 2005, Thomas et al. 2005, Kitamura et al. 2004, Hara et al. 2003) and the best efficiency reported was 8%. In addition, porphyrin dyes (Campbell et al. 2004, Wang et al. 2005) and phthalocyanine dyes have also been explored. Although the most efficient sensitizers to date are ruthenium complexes, metal free organic dyes have been attracting attention because of their ease of synthesis, high molar extinction coefficient, tunable absorption spectral response from the visible to the near infrared (NIR) region, environmental friendliness and inexpensive production techniques. The organic sensitizers can be divided in three major parts; donor, linker and acceptor components.

The HOMO energy level has π -characteristics being delocalized over the entire molecule and the LUMO energy level also has π -character where the electron density is shifted towards the acceptor group of the chromophores. Thus, when subjected to sunlight, an intramolecular charge separation occurs between the donor and acceptor components of the chromophores also known as pronounced push-pull effect supported by the HOMO-LUMO concept. The general characters of the orbitals are independent of the linker length. Ionic dyes could be regarded as potential sensitizers for DSSCs owing to their tunable absorption in the red to near-infrared region and also high absorptivities (ca. $10^5 \text{ Lmol}^{-1}\text{cm}^{-1}$). The ionic dyes tend to form strong self-associated aggregates in solution or at the solid-liquid interface. Amongst ionic dyes, much research has been done in development of dyes from class of cyanine dyes (Sayama et al. 2001) merocyanine dyes (Sayama et al. 2000; Sayama et al. 2003) and hemicyanine dyes (Wang et al. 2000) have been investigated in such applications due to their high absorptivity mainly in the red to the near-IR region. Dye-sensitized solar cells (DSSCs) are extensively studied as promising systems for photoelectric conversion, especially ruthenium complex-sensitized TiO_2 solar cells. However, the high price of RuCl_3 and difficulties in the synthesis of Ru(II) - polypyridine complexes limit the practical application of DSSCs at the present time. Compared with Ru(II) complexes, pure organic dyes not only are easier to prepare but also cost less. Therefore, some pure organic dyes with rich photophysical properties are expected to be promising sensitizers for nanocrystalline solar cells.

Recently, pure organic dye-sensitized solar cells have made great progress, indicating that pure organic dyes would be a promising type of sensitizer for DSSCs.

Of pure organic dyes, cyanine dyes have intense and broad absorption bands in the visible and near-infrared regions, and excellent sensitizing properties in photography. Recently, they have been used as sensitizers in DSSCs and some progress has been made in this area. Guo studied the co-sensitization of two new cyanine dyes. It was found that the aggregates of the cyanine dyes were efficient in light harvesting and that a mixture of two cyanine dyes could be employed to sensitize the solar cell over the entire visible spectrum. They generated a photoelectric conversion yield of 3.4%.

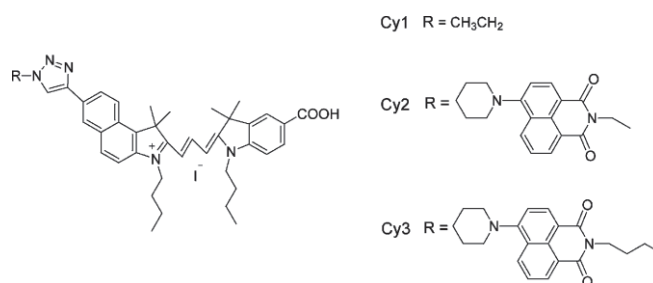


Figure 1: Molecular structure of some cyanine dyes

The main functioning of DSSC lies in the mesoporous oxide layer involving nanometre-sized particles sintered together for better electronic conduction. The metal oxide usually chosen is titanium dioxide (anatase) but many reports of other wide-gap oxides such as ZnO , SnO_2 and Nb_2O_5 have also been done. The key component comprises of dye that is responsible for sensitization of DSSC. The monolayer of the dye is attached to the nanocrystalline film of TiO_2 . The solar cell works on the principle of continuous cycle of oxidation and reduction of dye assisted by the redox electrolytic system. The cycle is initiated by the photo-excitation of dye followed by injection of a dye electron into the conduction band of the oxide film. The original state of the dye is subsequently restored by electron donation from the electrolyte via regeneration reaction. The redox electrolyte usually comprises an iodide/triiodide redox system that is dissolved in an organic solvent. Several other alternatives for this organic solvent has been thought upon with more emphasis on lesser volatility of the solvent. In this context, attention is increasingly being focused on ionic liquids, gelled electrolytes, polymer electrolytes and deep eutectic solvents (Wang et al. 2003). The regeneration of the oxidized dye to normal state is obtained by gain of electron by iodide

ion of redox system. The iodide ion is in turn regenerated by the reduction of triiodide at the counter electrode, with the electrical circuit being completed via electron migration through the external load. The dye component being the key material has been explored extensively via change in structural features, chromophores, conjugation bridges, ligands, metal ion, and nature of complexes. This systematic study has resulted in the development of mononuclear and polynuclear dyes based on metals such as RuII (Nazeeruddin et al. 1993; Nazeeruddin et al. 1997), OsII, PtII, ReI, CuI and FeII. Besides transition-metal complexes, a variety of organic molecules have been explored, including coumarin (Hara et al. 2005) squaraine, indoline, hemicyanine (Chen et al. 2005) and other conjugated donor-acceptor organic dyes (Hara et al. 2005, Kitamura et al. 2004, Hara et al. 2003) and the best efficiency reported was 8%. In addition, porphyrin dyes (Wang et al. 2005) and phthalocyanine dyes have also been explored. Development in the use of organic sensitizers began when Hara et al. (Hara et al. 2003a; Hara et al. 2005a) started using oligoenes containing dialkylaminophenyl groups as the donors and cyanoacrylic acid as the acceptor. The use of these dyes as sensitizers in liquid electrolyte systems yielded overall power conversion efficiencies of up to 6.8%.

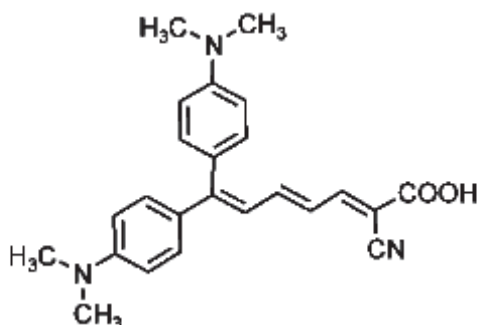


Figure 2: Molecular structure of Hemicyanine.

Among the metal-free organic dyes studied in DSSCs, coumarin-based dyes are promising sensitizers owing to their good photo-response in the visible region, good long-term stability under one sun soaking (Wang et al. 2007) and appropriate lowest unoccupied molecular orbital (LUMO) levels matching the conduction band of TiO₂. Some coumarin sensitizers (Hara et al. 2001; Hara et al. 2003b; Hara et al. 2003c; Hara et al. 2003d; Hara et al. 2005b; Hara et al. 2005c) as donor groups has reached efficiencies of upto 8.2 % values comparable to the standard N719 sensitizer. The importance of coumarin is understood from the fact that

on coumarin moiety was replaced N,N-dimethylaminophenyl (DMA) donor group (Kitamura et al. 2004), a significant hypsochromic shift of the maximum absorption peak is observed indicating that the coumarin is a stronger donor than the DMA group. In order to further obtain a red shift, it would be desirable to increase π - conjugation by extending the methine unit ($-\text{CH}=\text{CH}-$). However, such an extension by more double bonds would increase the instability of the dye molecule, owing to the possibility of isomer formation. It has been observed that introduction of thiophene moieties improves the solar cell performance mainly by broadening the absorption spectra thereby resulting in a large photocurrent and also relatively lowers the positions of the LUMO levels of the dyes. Also when the number of thiophene units are increased, the LUMO are much lowered due to increased π - π stacking leading to insufficient driving force for electron injection.

One of the factors limiting the efficiency of DSSCs can be non-optimized photocurrent arising from the poor optical response of the dyes in the NIR region. Hence, it is crucial to increase the light harvesting capability of the dyes towards the near-IR region. This is mainly because of the widened scope of improving efficiency of dye-sensitized solar cells. The conventional DSSCs are very effective in the region between 400–800 nm after which though they reduce to almost zero value of IPCE. This means that they usually do not absorb all the photons in the visible and near-infrared region so that a big fraction of photons extending till wavelength of 1200 nm are not used. Since traditional silicon devices makes use of these photons, it is crucial to increase the absorption range to obtain efficiency values closer to these semiconductor cells. 1200 nm are not used. Since traditional silicon devices makes use of these photons, it is crucial to increase the absorption range to obtain efficiency values closer to these semiconductor cells. Apart from the usual type of cells, there have been reports of a tandem cell incorporating a standard ‘dye-sensitised’ solar cell (DSSC) connected in series with a second cell sensitised by an organic dye, capable of absorbing photons in the near IR region without loss in Voc. The application of near-IR dyes also gives rise to the possibility of using them in transparent solar cells.

Conclusion

The importance of the low cost/high efficiency sensitized solar cells suggests that its need to be developed. Current DSSC conversion efficiencies need to catch them up with other more mature technologies (rigid modules) which have solar energy conversion efficiencies beyond 13-15%. While

the flexible DSSCs can have efficiencies in the range 5-10%, definitely comparable with more matured technologies, it is very important to note that these cells have an added advantage over other PV technologies in that they are superior solar energy harvesting devices under diffuse light conditions. The net effect of this will be to bring the operational efficiency of DSSCs into line with current technologies.

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Equations Showing Roll of Solvents Thermodynamic and Kinetic Control

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Abstract

In chemical reaction, choice of an appropriate solvent is required. Hence solvent with more than fifteen relative static permittivity can be divided in polar protic and aprotic solvents. It has been observed that the reaction in which the products are more polar than the reactants, then a polar solvent accelerates the rate of reactions. The rate of reaction is considerably influenced if either the reactant or the product or the activated complex interacts with the solvent. When the reactants interact with the solvents and are solvated leading to lowering the potential energy of the reactants then the activation energy increase lowering the reaction rate. In present study, we see that the polarity of the solvent has significant influence on the rate as well as mechanism of a reaction apart from its dielectric constant, external pressure, internal cohesion (viscosity) hydrogen bonding etc. For verification of the above parameter, we have discussed some important equation.

Keywords: Chemical reaction, static permittivity, polar protic, aprotic and solvents.

Introduction

The choice of appropriate solvents allows thermodynamic and kinetic control over a chemical reaction. Solvents with relative static permittivity more than 15 can be divided into protic and aprotic. The polar protic solvents solvate ions via hydrogen bonding. It can also stabilize cations by unshared electron pair. Aprotic solvents lack acidic hydrogen. It solvates charged species through opposite end of the dipole. It has been observed that the reaction in which the products are more polar than the reactants, then a polar solvent accelerates the reactions. Also, the rate of reaction is considerably influenced if either the reactant or the product or the activated complex interacts with the solvent. When the reactants interact with the solvents and are solvated leading to lowering the potential energy of the reactants then the activation energy increase lowering the reaction rate. On the other hand if activated complex interacts with solvent and is solvated lowering its potential energy, the activation energy decrease and rate of reaction increases. Therefore, it is concluded that the polarity of the solvent has significant influence on the rate as well as mechanism of a reaction apart from its dielectric constant, external pressure, internal cohesion (viscosity) hydrogen bonding etc. certain equations correlates the above parameters which are very useful to explain the different phenomena.

The Important equations are:-

1. Amis and Jaffe Equations.
2. The Laidler and Landskroener Equation
3. Laidler and Eyring Equation
4. Correlation Equation
5. Hammett's Equation
6. Grunwald and Winstein Equation
7. Swain's Equation
8. Taft Equation

Amis and Jaffe Equations

Amis and Jaffe have derived an equation for the rates of reactions between ions and dipolar molecules. In deriving the equation Amis and Jaffe first obtained that at Zero ionic strength the log of the rate constant changes should vary as $1/D^2$. They applied their equation to the inversion of sugar by hydrochloric acid and found that the data reproduced well. The difficulty here lies in the fact that in k versus $1/D$ plots should always have positive slopes which is contrary to much experimental data unless r^* is less than r_A when the slope would be negative. It seems that r^* would have little chance of being less than r_A even including the influence of solvation, since r^* is the radius of only the molecule plus any attached solvent.

The Amis modification of the Amis-Jaffe Equation, Since there was some uncertainty about the extrapolation to a reference state of dielectric constant in the Amis-Jaffe approach to the theory of ion-dipolar molecule reactions, Amis derived the dielectric constant dependence of k' using coulombic energy considerations.

This approach is limited to dielectric constant effects but it is simple and can be applied to various electrically charged and electrically unsymmetrical combinations of reactants. The complex formation activity coefficient approach is perhaps still best for elucidating the salt effect using the equation.

$$K_O' = Z_e (-E_D / RT) \quad K_O' = Z_e (-E_D / RT)$$

Amis obtained an equation expressing the rate change of dielectric constant in terms of mutual potential energy change between an Ion and a dipole, the equation may be written as -

$$\ln K'D_2 = \ln K'D_1 + Ze\mu_0 / K Tr^2 (1 / D_2 - D_2)$$

If a reference dielectric constant D_2 is taken as infinite in magnitude so that all electrostatic effect between reactants vanish and D_1 is taken dielectric constant D_1 the above equation reduces to

$$\ln K'D_2 = \ln K'_{\infty} + Ze\mu_0 / K Tr^2 D.$$

The Laidler and Landskroener Equation

$$\ln f_1 = Z^2 I e^2 / 2K T b_1 C (1/D-1) + 3 G i e^{2\gamma} / 8K T b^2 i (2/D-1)$$

Now for a reaction between a molecule A and ion B the rate equation may be written as –

$$\ln K' = \ln K_O + \ln f_A f_B / f^*$$

Where f^* is the activity coefficient of the complex formed by an ion and the molecule.

$$\ln K^1 = \ln K_O + e^2 / 2KT (1/D - 1) Z^2 A / b_A + Z_B^2 / b_B - (Z_A + Z_B)^2 / b^* + 3/2 (G_A / b_A^3 + G_B / b_B^3 - G^* / b^{*3})$$

Here G_A , G_B and G^* are complex function of the charges and structures of the respective species and are proportional to dipole moment.

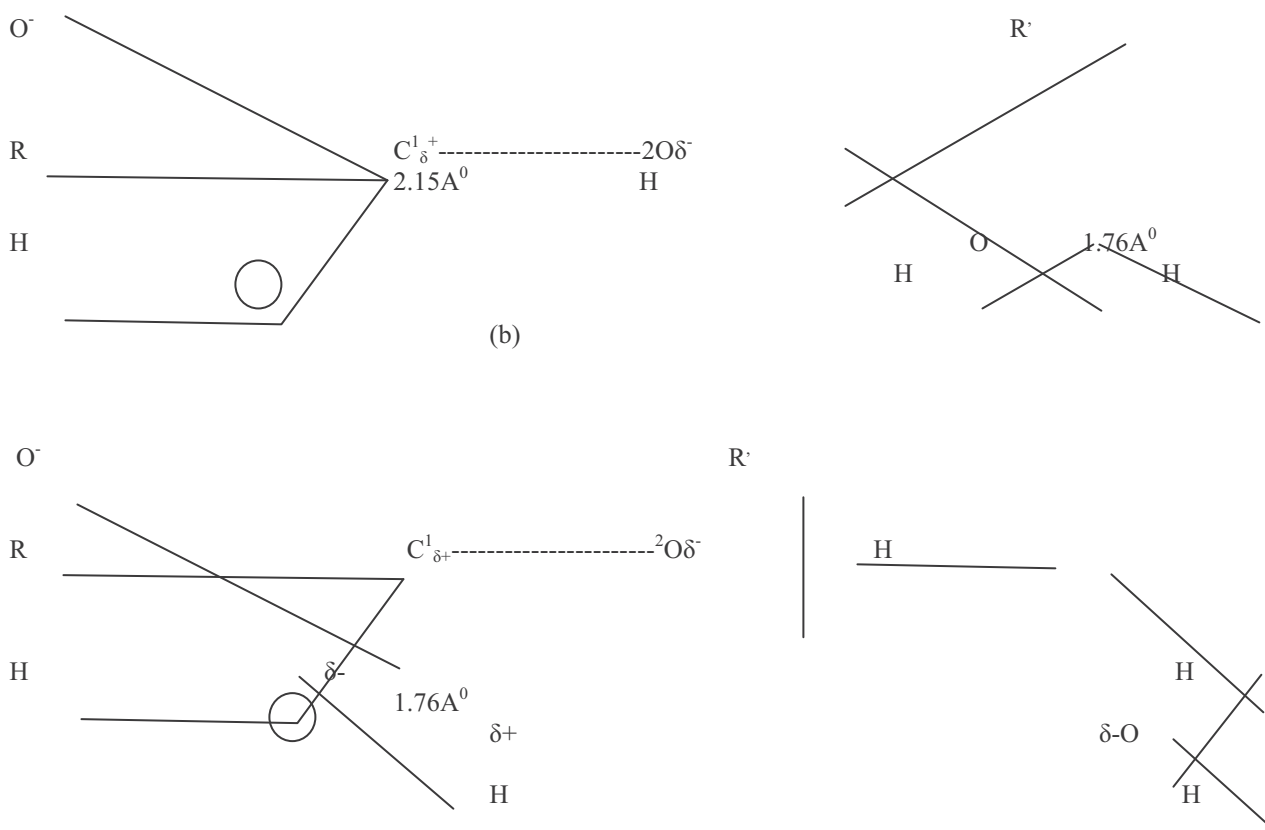
According to equation a plot of $\log K'$ versus $1/D$ should yield a straight line of slope S , given by the expression.

$$S = e^2 / 2.303 (2KT) Z^2 A / b_A + Z_B^2 / b_B - (Z_A + Z_B)^2 / b^* + 3/2 (G_A / b_A^3 + G_B / b_B^3 - G^* / b^{*3})$$

For a univalent ion reacting with a dipolar molecule, $Z_A = +1$, $Z^2 A = 1$, $Z_B = 0$, $G_B = 0$ and G_B is negligible. Under these circumstances equation becomes,

$$S = e^2 / 2.303 (2KT) 1/b_A - 1/b^* - 3/2 G^* / b^{*3} = e^2 / 2.303 K T b_A 2b^* - 2b_A b^{*2} - 3b_A G^* / b^{*3}$$

Landskroener and Laidler pointed out that in order to evaluate b^* and G^* a model for the activated complex has to be constructed. Their models for the acid and base hydrolysis of esters are as follows



Where diagram (a) is a model for the acid hydrolysis and diagram (b) the base hydrolysis.

The model for the acid hydrolysis Yield $G^* = 9.1 \times 10^{-14}$ sq.cm. and if b_A is taken as 1.7×10^{-8} cm. equation becomes $S = 1.07 \times 10^4 / b^{*3} T (2b^{*3} - 3.4b^{*3} - 47.31)$

The model for the base hydrolysis gives $G^* = 5.4 \times 10^{-16}$ sq.cm. and if b_A is taken as 1.4×10^{-8} cm equation becomes $S = 1.30 \times 10^4 / b^{*3} T (2b^{*3} - 2.8b^{*3} - 22.68)$

To test their theory, Landskroener and Laidler plotted $\log k^1$ versus $1/D$, for both the acid and base hydrolysis of esters. The slopes of the straight lines obtained were used to calculate b^* from equation in the case of acid hydrolysis. These values of b^* should be of the order of magnitude of a molecular dimension in the case of conformity of theory with experiment.

Laidler and Eyring Equation

by Vander Waals forces cannot, except as a very rough approximation, be neglected. They used Kirkwood's expression for the free energy of transfer of a strong dipole of moment μ from a vacuum to a medium of dielectric constant, D for a symmetrical charge distribution within a molecule, this expression for a molecule of radius 'r' is given as

$$\Delta G = 2 / r^3 D - 1 / 2 D + 1 + \phi$$

Where ϕ is a non-electrostatic term using the bimolecular reaction



And applying the above equation the expression for the specific velocity constant can be obtained as follows:

$$\ln K' = \ln CX^{KT/h} K^* 8 - 1 / K T. D - 1 / 2 D + 1 \mu^2 A / r^3 A + \mu^2 B / r^3 B / \mu^{*2} / r^{*3} + \phi_A + \phi_B - \phi^* / KT$$

In this equation, D is the dielectric constant of pure solvent provided the non-electrostatic terms are negligibly small, a plot of $\log K'$ versus should give a straight line. Laidler and Eyring plotted the above function using the dielectric constant of the solvent for quaternary ammonium salt formation at 29°C in alcohol - benzene mixtures and for the acid hydrolysis of ethyl orthoformate in ethanol - water mixtures. These authors argue that from the stand point of the influence of the solvent, ester hydrolysis should be regarded as a reaction between the two dipoles, water and ester. For some reactions in benzene-nitrobenzene solvents, however, similar plots show considerable curvature. In solutions rich in nitrobenzene the reactions are much faster than theory predicts. Specific solvent effects are common in quaternary ammonium salt formation reactions. For example, the triethylamine - ethyl iodide reaction, studied by Grimm, Ruff, and Wolff in a variety of solvents, gave rates and dielectric constants that are not even in same order. Laidler and Eyring explained the deviations in terms of the nonelectrostatic contributions to which the most important contribution is probably the solvating power of the solvent.

Correlation Equation

A number of workers have explained solvent effect theoretically and formulated mathematically. Therefore, the mathematically formulated equations are more helpful in the prediction of the effect of solvent on the rate of reactions. Following equations require careful discussion.

Hammett's Equation

Considering the effect of a substituent upon the rate or equilibrium constant of a general reaction for a number of a class of aromatic compounds. Hammett proposed an equation taking a general reaction in which a compound and it will commonly undergo. This equation is as follows $\log k' = \log k_o + \rho \sigma$

Where K_o is the specific rate or equilibrium constant for a general reaction, σ is the substituent constant that depends upon the nature of substituent and ρ is the reaction constant

Laidler and Eyring believed that dipolar forces somewhat strong than non-electrostatic ones depending upon the temperature, the medium and the nature of the reaction. Hammett has given a list of values. The values for a number of reactions were also tabulated. The works of Hartmann *et al.* Supported the theory. They determined the said catalyzed esterification of twenty two substituent benzoic acid with cyclohexanol and found that experimental values of esterification rate constant for meta and para substituent's were in a close agreement with those using Hammett's equation. The average difference between $-\log K'$ calculated and $-\log K'$ observed by them was 0.082 or about 2%.

Branch and Calvin emphasize that Hammett's equation is an approximate empirical one i.e. applicable only within definite limits.

Grunwald and Winstein Equation

Grunwald and Winstein found that for reactions taking place by S_N^1 mechanism where value of $\log k$ for solvolysis of any one compound in the same solvents, a straight line was obtained. They illustrated a function Y for a quantitative measurement of ionising power of a given solvent and for the definition of Y the following equation was illustrated $\log K^{BuCl} = Y + \log K_o^{BuCl}$

Where $\log K^{BuCl}$ and $\log K_o^{BuCl}$ are the specific rate constants for the solvolysis of tertiary butyl chloride at 25°C in the given solvent and in 80% ethanol - water solvent.

The specific velocity constants for solvolysis in solvents 1 and 2 are K_1 and K_2 . Therefore, the general equation relating the rate constants for the "ionising power" Y_1 and Y_2 of the respective solvents can be represented as: $\log K_1 = m (Y_1 - Y_2) + \log K_2$

Where m is a characteristic constant at a given temperature of the compound being studied Or, $\log (K_1/K_2) = m (Y_1 - Y_2)$ Denoting K_1 as K and K_2 as K_o , the general equation become as $\log k / k_o = mY$

Where k is the specific velocity constants for the solvolysis of a given compound in a given solvent at a given temperature and k_o is the specific velocity constant for the solvolysis of a compound 80% (V/V) ethanol at the same temperature.

Swain's Equation

The Swain's equation is linear free energy relationship. Swain Co-workers reported the polar displacement reaction of uncharged organic molecules with a nucleophilic reagent N and an electrophilic reagent E by the equation given below:



They relate the overall rate constant k of this reaction to the corresponding rate constant K_o where water acts as both N and E in the same medium at the same temperature :



The linear free energy equation correlating these rates would be $\log k/k_0 = S_n + S_e$

Where, n = nucleophilic constant and is a quantitative measure of the nucleophilic reactivity of N ($n=0$ for water), e the electrophilic constant i.e., a measure of the electrophilic reactivity of E ($e = 0$ for water), and S and S' , substrate constants among different reagents N and E .

From equation other corollaries are derivable, which are well known and useful quantitative relationship between structure and reactivity.

Therefore, all these correlation equation are empirical and do not throw in the microscopic behavior of the reagents inserting the rate.

Recently, Singh et al. have reported that the tendencies of the solvent to break the water cluster in case of DMSO-water solvent system is also responsible for enhancing the rate of hydrolysis of esters. They have also further reported that the tendencies of solvent to retard the polarity of water is responsible for depressing the rate of hydrolysis as such in case of DMF – H₂O and also in case of DMSO – H₂O (beyond 9 mole % of DMSO) solvent systems.

The equation mentioned above says that the solvating power cations (carbonium ions, at least) is proportional to nucleophilicity and that there is a smooth continuous transition from S_N^1 and S_N^2 mechanism. According to Hine, n and e represent some correlations of n with basicity and e with acidity, and the values of S and S' show compounds reacting S_N^1 mechanism to be relatively more susceptible to nucleophilic attack.

Taft Equation

As the failure of Hammett equation is partly steric in nature, Taft developed a procedure to separate the polar and steric effects of substituent, based on suggestions of Ingold, that in the acid and the base catalyzed hydrolysis of esters (RCOOR'), the ratio of the base catalysed rate constant K_B to that of acid catalyzed rate constant k_A is a function of the polar effect of substituent only. Taft assumed that the steric and resonance effects for acid and basic hydrolysis of ester are the same, so according to him, the rate differences for these two types of reactions, would therefore be caused only by the inductive effects of R and R' IN RCCOR' , and the inductive effects of x could be determined by measuring the rates of acid and base catalyzed hydrolysis of series of $\text{XCH}_2\text{COOR}'$ where R' would be kept constant.

Taft originally wrote,

$$I = e^* \sigma^*$$

With the value of σ^* for a group of R defined by

$$\sigma^* = \frac{1}{2} \cdot 48 [\log (k/k_0) - \log (k/k_0)A]$$

Where, K = the rate constant for the hydrolysis of XCH_2COOR K_0 = the rate constant for the hydrolysis of the reference ester, $\text{CH}_3\text{COOR}'$.

The subscript B and A denote the base and acid catalyzed hydrolysis at the same temperature and

σ^* = the substituent constant and is characteristic of the substituent X .

2.48 is an arbitrary constant.

Finding that σ^* thus defined for $R = \text{XCH}_2$ correlates well with the acidity constants of the acids XCH_2COOH , Taft then defined the value of σ' for the group X as $\sigma' = 0.045 \sigma^*$

the choice of the co-efficient 0.45 was intended to make σ' equal to $\log k/k_0$ for the acidity constants of the Roberts and Morel and acids in water.

Hence Taft modified his original equation with $I = \sigma'^1 e^1$

Where e^1 depends only on the reaction and σ' depends only on the composition and structure of the non reacting part of the molecules. The value of σ' for a group X is defined by the relation, $\sigma' = 0.262 \log k/k_0$

Where k_0 is the acidity constant of acetic acid in aqueous solution at 25°C and k is the corresponding constant for the acid XCH_2COOH .

Since a large proportion of the available σ'^1 values derive from the data on the acidity constant of XCH_2COOH , the modified equation of Taft is preferable to the original one. The substitution of x for hydrogen in acetic acid would not be expected to involve either steric or resonance effects, and Taft found variety of other reactions in which inductive effects alone are to be expected and to which the above equation applies with considerable precision. The equation has found limited utility in heterocyclic systems.

Conclusion

Almost recent works on linear solvation energy relationship are based on a powerful predictive model, which has provided a framework for numerous studies into specific molecular thermodynamic solute system. Model is based on an equation having three conceptually explicit terms.

The conclusions about the influence of heterocyclic ring substituent on the tautomeric equilibria are summarized. It is difficult to correlate them within any well-defined scheme. The energy differences between pairs of tautomers are generally quite small and attempts to analyze the using.

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Study of Antioxidant Properties of Some Phenolic Compounds

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Abstract

Reactive oxygen species and free radicals play a major role in food deterioration. In the recent past, there has been an increasing significance in the therapeutic efficacy and preservative properties of natural plant antioxidants to prevent oxidative reactions both in food stuffs and in biological systems. The plant - based essential oils have provoked interest as the sources of natural products because of their safety, wide consumer's acceptability and the likelihood of their exploitation for potential versatile functional uses. The addition of synthetic antioxidants to oils and/or foods is one of the most efficient ways to prevent lipid oxidation. However, the safety of synthetic additives has been questioned stimulating the evaluation of naturally occurring compounds with antioxidative properties. Although there is no assurance of the safety of natural antioxidants, there is some comfort knowing that such antioxidants are purified from natural products that have been consumed for generations. Phenolic compounds in plants are recognized as important compounds in conferring stability against oxidation. Natural antioxidant phenolics can be classified into a lipophilic group, tocopherols, and a hydrophilic group, including simple phenolics, phenolic acids, anthocyanins, flavonoids and tannins. In this work we studied the most important groups of natural antioxidants, with some peculiarities related to chemical composition.

Keywords: Phenolic compound, antioxidant, essential oils, phenolic compound and lipid oxidation.

Introduction

In living organisms there is equilibrium between natural antioxidant defence mechanism and reactive oxygen species (ROS), which are produced in the body by various exogenous and endogenous redox cycling processes as by-products. When the equilibrium is disturbed, the ROS can induce oxidative damage to various biomolecules including carbohydrates, lipids, proteins, DNA and RNA in the human body associated with cellular damage, tissue injury and genetic mutation. This oxidative damage may accelerate aging, cancer, cardiovascular disorders, inflammation and neurodegenerative disorders. The harmful action of the free radicals can, however, be blocked by antioxidant substances which scavenge the free radicals and detoxify the organism. Current research into free radicals has confirmed that foods rich in antioxidants play an essential role in the prevention of cardiovascular diseases and cancers and neurodegenerative diseases. Therefore, plant derived antioxidants are now receiving a special attention. A large number of phenolic compounds present in vegetable foods, such as fruits and nuts, have been reported to possess good antioxidant properties. Moreover, the essential oils and various extracts of aromatic plants have been of great interest for their potential antioxidative effects for the preservation of the foods from the toxic effects of the oxidants.

Antioxidant Properties of Some Phenolic Compound

The term "phenolic" is used to define substances that possess one or more hydroxyl groups (OH) substituents bonded onto an aromatic ring. Compounds that have several or many phenolic hydroxyl substituents are often referred to as polyphenols. Due to its chemical structure phenolic compounds have the ability of phenoxide ion delocalize. The phenoxide ion can lose a further electron to form the corresponding radical which can also delocalize. In reference to this property, phenolic compounds have radical scavenging and antioxidant activity.

Phenolics are large and heterogeneous groups of secondary plant metabolites that are distributed throughout the plant kingdom. They have been implicated in a number of varied roles including UV protection, pigmentation, disease resistance, and nodule production. Phenolics have a wide variety of structures. Flavonoids, tannins, and phenolic acids are the main phenolic compounds.

Due to toxicological concerns of synthetic antioxidants, there have been increasing interests in identifying phenolic compounds in plants to minimize or retard lipid oxidation in lipid-based food products. Most of these natural antioxidants come from fruits, vegetables, spices, grains, and herbs.

Hundreds of natural phenolic compounds have been reported to possess high antioxidant properties. Their use in foods,

however, is limited by certain requirements not the least of which is adequate proof of safety. Only a few of them can be commercially applied in foods. The main lipid-soluble antioxidants currently used in food are monohydric or polyhydric phenols with various ring substitutions. For maximum efficiency, primary antioxidants are often used in combination with other phenolic antioxidants or with various metal sequestering agents, e.g. tocopherols with citric acid and isopropyl citrate. Most important commercially available natural antioxidants are tocopherols (vitamin E), ascorbic acid (vitamin C) and rosemary extract. Compounds such as β -carotene, ascorbic acids have demonstrated to have antioxidant and synergistic activity in despite of their non-phenolic structure.

Natural phenolic compounds with antioxidant properties can be classified into a lipophilic group (tocopherols mainly) or a hydrophilic group (phenolic acids and flavonoids). Tocopherols and tocotrienols are the most important lipid-soluble antioxidants present naturally in vegetable oils. Both tocopherols and tocotrienols share the same ring structure, but tocotrienols have unsaturated carbon chains.

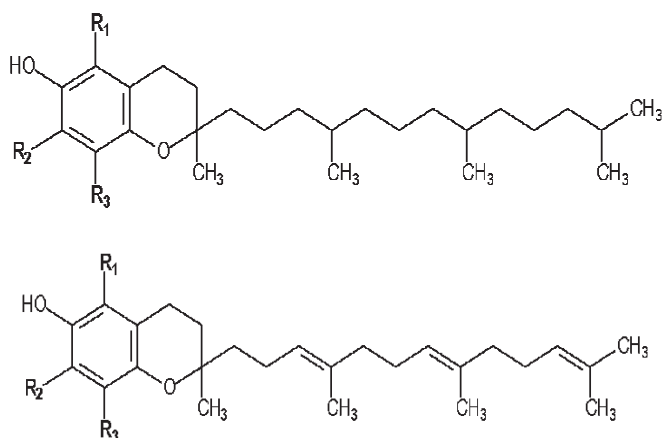


Figure 1& 2

Position of Methyl Group on Aromatic Ring	Tocopherol Structure	Tocotrienol Structure
R_1, R_2, R_3	α – Tocopherol	α – Tocotrienol
R_1, R_3	β – Tocopherol	β – Tocotrienol
R_2, R_3	γ – Tocopherol	γ – Tocotrienol
R_3	δ – Tocopherol	δ – Tocotrienol

Table 1

A relatively high proportion of the tocopherols present in crude vegetable oils survives the oil processing steps and

remains in sufficient quantities to provide oxidative stability in the finished products. As antioxidants, tocopherols exert their maximum effectiveness at relatively low levels, approximately equal to their concentration in vegetable oils. Tocopherols interrupt free radical chain reaction by capturing free radicals generated during oil autooxidation.

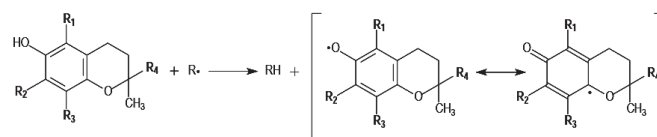


Figure 3 and 4

Although there is some controversy about the antioxidant capacity of the different tocopherol homologs, most studies show a relative antioxidant activity in the order $\delta > \gamma > \beta > \alpha$. This order, however, may be altered by effects of temperature, substrate composition and tocopherol concentration. Hence, the optimum tocopherol concentration to inhibit peroxide formation should be established for each specific and well defined system.

Another group of phenolic substances with antioxidant properties are the phenolic acids. Among them, gallic acid, a natural product arising from tannin hydrolysis, has attracted considerable interest. Novel antioxidative food additives are being developed using gallic acid as a starting compound. Strlic et al. demonstrated that in Fenton - type systems containing Fe (III) H_2O_2 , gallic acid at a molar ratio of GA:Fe $> 2:1$ has an antioxidative role due to its pronounced $HO\cdot$ scavenging properties.

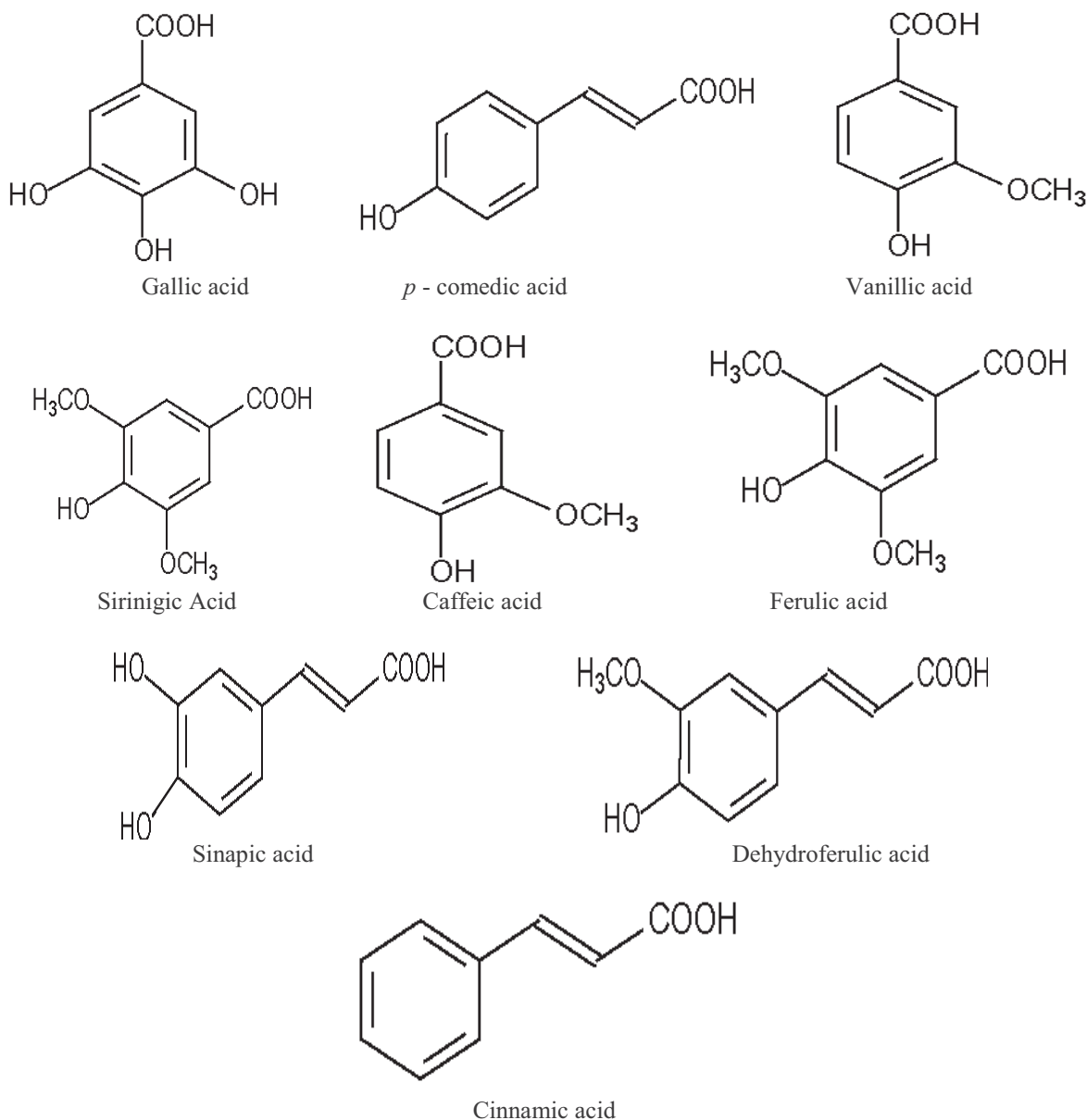
Oil Source	α – Toc	β – Toc	γ – Toc	δ – Toc
Canola	103 - 170	134	174 - 403	10, 3 - 41
Sunflower	622 - 723	19 - 41	5 - 65	
Soybean	81 - 151	27 - 33	545 - 624	207 - 300
Walnut	15 - 29	1 - 8	207 - 355	30 - 62
Peanut	49 - 304	1 - 41	99 - 389	
Olive	104 - 425	1 - 3	5 - 30	
Jojoba	11 - 23		24 - 41	
Wheat germ	984 - 1845	386 - 739	231 - 681	105
Safflower	386 - 477	8 - 9	2 - 44	10
Palm	3 - 185		4 - 36	
Sesame	12 - 113	3 - 6	244 - 400	5 - 32
Corn	134 - 344	14 - 31	412 - 729	36 - 39
Cottonseed	402 - 573	1,5 - 40	317 - 572	7,5 - 10

Table 2

Study of Antioxidant Properties of Some Phenolic Compounds

Hydrolysis has attracted considerable interest. Novel antioxidative food additives are being developed using gallic acid as a starting compound. Strlic et al. demonstrated that in Fenton-type systems containing Fe (III) / H₂O₂, gallic acid at a molar ratio of GA:Fe > 2:1 has an antioxidative role due to its pronounced HO• scavenging properties. However, at a molar ratio of GA:Fe < 2:1 the overall role is prooxidative due to the promotion of HO• production. However, at a molar ratio of GA:Fe < 2:1 the overall role is prooxidative due to the promotion of HO• production. Monohydroxylated phenolic acids, such as vanillic, caffeic,

sinapic, siringic, ferulic and coumaric acids are relatively ubiquitous. Their antioxidant capacities have been demonstrated in several “*in vitro*” and “*in vivo*” studies. Hydroxycinnamic acids such as *p*-coumaric acid are abundant in higher plants and are found with various hydroxyl and methoxyl substitutions and may exist in esterified forms or bound to proteins. As an antioxidant, *p*-coumaric acid is an effective radical inhibitor *in vitro* but contains only moderate inhibitory properties against lipid peroxidation.



The antioxidant activity of phenolic acids is mainly due to their redox properties, which allow them to act as reducing agents, hydrogen donors and singlet oxygen quenchers. In addition, some phenolic acids have a metal chelation potential.

Conclusion

Aromatic herbs have a long history of culinary use. Not only are many of the flavours and aromas distinctively pleasant, but they can be used to conceal off flavours and odours. The high susceptibility to oxidation of the fat and oil polyunsaturated fatty acids used in human foods requires the application of antioxidants. Total antioxidant activities of the plant extracts cannot be evaluated by any single method, due to the complex nature of phytochemicals. Thus, two or more methods should always be employed in order to evaluate the total antioxidative effects of vegetables or food stuff.

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Anti Diabetic Medicinal Plants of West Bengal - A Review

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Abstract

We are at risk. Modern sedentary life style has made us susceptible to extremely serious endocrinal disorder of hyperglycemia i.e., Diabetes mellitus. This disease has deeply affected our national health care systems. Nature our mighty healer offers a holistic multi-modality approach in treating this disorder in the form of various anti-diabetic herbs. Traditional plants have been used throughout the world for the treatment of diabetes mellitus. Among many pharmacological therapeutics and alternatives ; several medicinal herbs have been known to cure and control diabetes; additionally they have no side effects and are lower in cost. The present paper is an attempt to enlist the anti-diabetic herbs and their potential mechanism of action found in West Bengal. Some of the plants are described in this review which emphasize the use of herbal plants for hassel-free curing of diabetes over a period of time. Though the effect of these plants may delay the development of secondary complications due to diabetes but they fairly provide good amount of antioxidants that are known to prevent different diseased states.

Keywords: Diabetes Mellitus, Medicinal plants, antioxidants, hyperglycemia

Introduction

Diabetes is the condition when a person does not produce enough of insulin or is unable to use it effectively. Diabetes is of two types. In type 1- Infantile or juvenile onset diabetes or Insulin Dependent Diabetes Mellitus [IDDM] and type 2 – non- insulin dependent diabetes mellitus which sets in adults generally over 35 years of age and accounts for over 90 % of diabetics. The normal levels are 80 [fasting] and 120 [after food intake]. The important thing is to bring down the blood sugar levels to normal as soon as possible. It is thought that Type- 1 is the result of auto-immune disease occurring after a viral infection, like chicken pox, that sometimes damages the pancreas. The extent of damage varies from person to person. Since adequate insulin is not produced, it has to be given from outside to such patients. This is the probable reason how children with no such family history, develop diabetes. Genetic reasons can also cause it. Type 2 diabetes is the most common form of diabetes. It is a metabolic disorder due to deficiency of insulin, a hormone produced by the pancreas. Either sufficient insulin is not produced or the body does not use it properly, resulting in high blood sugar levels. When this condition persists, diabetes sets in. this can lead to problems affecting the heart, nerves, kidneys and eyes besides leading to other complications.

The symptoms include –

- Thirst
- frequent urination

- increase in blood sugar levels
- Frequent hunger
- Compulsive desire to eat

Obesity, lack of exercise and improper diet contribute greatly to the development of diabetes. It is mostly found out by accident, like a blood test is being done for a regular health check up. People of Asian origin develop this condition even without obesity, as the genetic susceptibility is more. Obese teenagers can also develop type 2 diabetes. Improper diet consisting of fast foods, soft drinks and untimely food intake along with lack of exercise are the main causes that result in **Insulin Resistance** – a precursor of type 2 diabetes. Undetected and untreated, this condition can cause damage to the heart and full-blown diabetes can set in. Here the pancreas produces insulin but the body develops resistance to it.

Plants that Show Hypoglycemic Activity

Since decades, traditional medicines all over the world have advocated its utility in diabetes treatment. The ethno-botanical information reported about 800 plants as anti-diabetics (Alarcon et al., 1998). Most of the drugs from plant sources are secondary metabolites which have insignificant role in plant metabolisms but play an important role in the plant defense mechanism. Plant derived active principles representing numerous chemical compounds has demonstrated activity consistent with their possible use in the treatment of Diabetes Mellitus (Bailey and Day, 1989;

Marles and Farnsworth, 1995). Approximately 200 pure compounds from plant sources reported to show blood glucose lowering effect. The active principals are steroids, carbohydrates, glycosides, flavonoids, peptides amino acids, lipids, alkaloids, phenolics, glycopeptides terpenoids and iridoids. Several herbs have shown antidiabetic activity when assessed using presently available experimental techniques (Saifi et al., 1999). Though, oral hypoglycemic agents/insulin is the main therapeutic agent for treatment of hyperglycemia, they have prominent side effects and fail to significantly alter the course of diabetic complications (Rang and Dale, 1991). More than 1200 species of plants have been screened for activity on the basis of ethno pharmacology or on random basis. The products obtained range from marine algae and fungi to phylogenetically advanced classes of compounds. West Bengal, which lies in the North-eastern part of India, is rich of its flora and fauna; and is one of the hotspots of biodiversity. The flora of this region includes aromatic and medicinal plants with a number of bioactive compounds. A report of some commonly used anti-diabetic plants which are found in West Bengal is given below :

***Acacia arabica* or *nilotica* (gum arabic tree)**

Feeding of 94% seed diet to normal rats showed significant hypoglycemic effect versus controls. However, the same diet failed to show any hypoglycemic effect in alloxanized rats (175 mg/kg SC) indicating that plant acts through release of insulin (Singh et al., 1975). Powdered seeds of *A. arabica* administered in doses of 2, 3 and 4 gm/kg body weight exerted a significant hypoglycemic effect in normal rabbits by initiating the release of insulin from pancreatic beta cells. No acute toxicity and behavioral changes were observed at these doses (Wadood et al., 1989).

***Aegle marmelos* (Bael / Bilva)**

Oral administration of aqueous decoction of *Aegle marmelos* root bark (1 ml/100 gm) showed hypoglycemic effect which was maximum (44%) at 3 h in normal fasted rats. In addition, the same extract completely prevented peak rise of blood sugar at 1 h in OGTT. The hypoglycemic activity was reduced upon storage of extract (Karunanayake et al., 1984). Aqueous extract of the leaves (1 gm/kg for 30 days) significantly controlled blood glucose, urea, body weight, liver glycogen. Serum cholesterol of alloxanized (60 mg/kg IV) rats as compared to controls and this effect was similar to insulin treatment (Ponnachan et al., 1993). When fed as aqueous leaf extract (1 gm/kg/day) to STZ (45 mg/kg IV) diabetic rats for 2 weeks, it decreased malate dehydrogenase levels (an enzyme known to increase diabetes) in comparison to diabetic controls. The extract was equi-

effective in comparison to insulin in restoring blood glucose and body weight to normal levels (Seema et al., 1996). Aqueous leaf extract administered orally for 28 days also normalized STZ (45 mg/kg body weight) induced histopathological alterations in the pancreatic and kidney tissues of rats (Das et al., 1996).

***Allium cepa* (Onion)**

Various ether soluble fractions of onion as a single oral dose (0.25 mg/kg) showed significant hypoglycemic effect in normal fasted rabbits. Ether extract showed most potent hypoglycemic action (Augusti, 1973). Petroleum ether insoluble fraction of the ether extract of dried onion powder (100 mg/kg) given orally for 7 days to alloxanized (180 mg/kg) diabetic rabbits caused a significant antihypoglycemic effect (Mathew and Augusti, 1975). In a preliminary study of seven different fractions obtained from onion bulb, only petroleum ether and chloroform extracts significantly lowered blood sugar in OGTT (2 mg/kg) in rabbits (Gupta, 1977). Administration of a sulphur containing amino acid isolated from *A. cepa*, called S-methyl cysteine sulphoxide (SMCS) to alloxanized rats significantly controlled blood glucose and lipids in serum and tissues, and normalized the activities of liver hexokinase, glucose 6-phosphatase and HMG CoA reductase. The effect was comparable to that of glibenclamide and insulin (Kumari et al., 1995).

***Allium sativum* (Garlic)**

Oral administration of 0.25 mg/kg of ethanol, petroleum ether, ethyl ether extract of *Allium sativum* causes 18.9, 17.9, 26.2% reduction in blood sugar in alloxan- diabetic rabbits (150 g/kg IV) (Jain and Vyas, 1975). Oral administration of 0.25 gm/kg allicin (isolated from *A. sativum*) produced hypoglycemia comparable to tolbutamide in mildly diabetic animals (Mathew and Augusti, 1973). Pretreatment with aged garlic extract (AGE) in stress induced hypoglycemia model of mice significantly prevented adrenal hypertrophy, hyperglycemia and elevation of cortisone without altering serum insulin levels. The efficacy of AGE was the same as that of diazepam. Thus, AGE may prevent stress-induced risk of DM and its progression (Kasuga et al., 1999). Administration of S-allyl cysteine sulphoxide (SACS), a sulphur containing amino acid, significantly decreased the concentration of serum lipids, blood glucose and activities of serum enzymes like alkaline phosphatase, acid phosphatase and lactate dehydrogenase and liver glucose 6-phosphatase. It also significantly increased liver and intestinal HMG CoA reductase activity and liver hexokinase activity (Sheela and Augusti, 1992). In another study, oral administration of

SACS to alloxan diabetic rats for one month ameliorated glucose intolerance, weight loss, depletion of liver glycogen in diabetic rats in comparison to glibenclamide and insulin (Sheela et al., 1995).

***Aloe barbadensis* (Aloe vera)**

Extracts of aloe gum effectively increased glucose tolerance in both normal and diabetic rats (Al-Awadi and Gumma, 1987). Chronic but not single administration of the exudates of the leaves of *A. barbadensis* showed significant hypoglycemic effect in alloxan - diabetic mice. However, single as well as chronic administration of the bitter principle showed significant hypoglycemic effect in the same model. The hypoglycemic effect of single dose of the bitter principle was extended over a period of 24 h with maximum hypoglycemia observed at 8 h while chronic administration (exudates twice daily and the bitter principle once a day for 4 days) showed maximum reduction in plasma through glucose level at the 5th day. Hypoglycemic effect of *A. vera* and its bitter principle is mediated through stimulation of synthesis and/or release of insulin from the beta-cells of Langerhans (Aiabnoor, 1990).

***Areca catechu* (Areca nut)**

An epidemiological study has shown that nitrosamines released during betel chewing may contribute to the risk of developing NIDDM (Mannan et al., 2000). Subcutaneous administration of alkaloid fraction of *A. catechu* (0.05-0.5 mg/kg) in alloxanized rabbits (140 mg/kg of body weight) had shown significant hypoglycemic effect lasting for 4-6 h (Chempakam, 1993).

***Azadirachta indica* (Neem)**

Alcoholic extract of *Azadirachta indica* showed hypoglycemic and anti-hypoglycemic effect in normal, glucose fed and STZ diabetic rats (Chattopadhyay et al., 1987). The plant exerts its pharmacological activity independent of its time of administration, that is, either prior or after alloxan administration (Khosla et al., 2000). The plant blocks the action of epinephrine on glucose metabolism, thus increasing peripheral glucose utilization (Chattopadhyay, 1996). Bitter principles – nimbin, nimbinin and nimbidin – were isolated from this plant.

***Beta vulgaris* (Beet)**

Various glycosides isolated from the root extract of *B. vulgaris* have been shown to increase glucose tolerance in OGTT conducted in rats (Yoshikawa et al., 1996). In addition, the extract also inhibited non-enzymatic glycosylation of skin proteins in STZ diabetic rats (Tunali et al., 1998).

***Biophytum sensitivum* (Little Tree plant)**

The leaf extract of the plant has been shown to exert significant anti-hyperglycemic effect in alloxanized rabbits possibly by pancreatic cells stimulating action as the plant was effective in only mild to moderate and not in severe diabetes (Puri and Baral, 1998).

***Bombax ceiba* (Cotton Tree)**

A C-flavonol glucoside isolated from *B. ceiba* leaves called as Shamimin has been shown to exert significant hypoglycemic activity at the dose of 500 mg/kg in rats (Saleem et al., 1999).

***Brassica juncea* (Indian mustard)**

Oral feeding of *B. juncea* diet (10% w/w) for 60 days to normal rats led to significant hypoglycemic effect. This effect was attributed to stimulation of glycogen synthetase (leading to increase in hepatic glycogen content) and suppression of glycogen phosphorylase and other gluconeogenic enzymes (Khan et al., 1995).

***Cajanus cajan* (Pigeon pea)**

A single dose of un-roasted seeds of *C. cajan* administration as a 60 and 80% diet to normal and alloxanized mice caused a significant reduction in the serum glucose levels after 1-2 h and a significant rise at 3 h. On the other hand, roasted seeds administration caused a significant increase in the serum glucose levels during a 3 h experimental period. Roasting of seeds at high temperature for 30 min resulted in the total loss of hypoglycemic principle but not the hyperglycemic principle present in the seeds (Amalraj and Ignachimuthu, 1998). Cooked diet of *C. cajan* has also shown significant hypoglycemic effect in healthy human volunteers (Panlasigui et al., 1995).

***Capparis deciduas* (kair)**

Oral feeding of diet containing (30%) *C. decidua* fruit powder for 3 weeks to alloxanized (80 mg/kg IP) diabetic rats (blood glucose, 450 mg %) showed significant hypoglycemia (blood glucose, 120-130 mg %) (Yadav et al., 1997).

***Citrullus colocynthis* (Bitter-apple)**

Aqueous extract of its fruit showed dose-dependent increase in insulin release from isolated islets (Abdel-Hassan et al., 2000). Oral administration of aqueous extract (300 mg/kg) in normal rabbits significantly reduced plasma glucose after 1 h and highly significant reduction after 2, 3, and 6 h. Glycosidic extract (50 mg/kg) was more effective in lowering fasting glucose as compared to alkaloid extract. Graded doses (10, 15 and 20 mg/kg) of saponin also reduced

plasma glucose concentration in alloxanized rabbits. Thus, saponins and glycosidic components of *C. colocynthis* are responsible for its hypoglycemic effect (Abdel-Hassan et al., 2000).

***Coccinia indica* (Scarlet vine)**

It is used in Ayurveda and Unani system of medicine for treatment of diabetes, skin eruptions, tongues sore, earache, etc (Chopra et al., 1956). Blood glucose lowering effects of leaves of *C. indica* have been shown in a double-blind control trial controlling 16 patients with uncontrolled maturity onset diabetes and 16 controls. Treatment was given for 6 weeks and 10 patients showed marked improvement in their glucose tolerance (Azad et al., 1979). In a clinical study (n=30), oral administration of dried extract of *C. indica* (500 mg/kg for 6 weeks) significantly restored the raised activity of lipoprotein lipase and the levels of G-6 phosphatase and LDH, which are otherwise increased in the severe diabetics. This action of the plant extract was akin to that of insulin (Kamble et al., 1998). It was suggested that the hypoglycemic effect of *C. indica* are partly mediated through suppression enzyme glucose-6-phosphatase (Hossain et al., 1992).

***Curcuma caesia* (Black turmeric)**

The rhizome of *C. caesia* is used traditionally by the people of West Bengal (Warjeet et al., 2006).

***Eugenia jambolana* (Jamun)**

It belongs to the family Myrtaceae. The fruits, the seeds and the pulp are used as anti diabetics in Indian medicine since ages. The aqueous and alcoholic extract of seeds exhibited significant decrease in plasma glucose levels, upon chronic administration in diabetic rats (Grover et al., 2000; Grover et al., 2001). The possible mechanism suggested is the extract induced enhancement of serum insulin levels by the pulp extract. The active principles isolated were a peptidoglycan, composed of 7-amino acids; and an oligosaccharide.

***Ficus bengalensis* (Indian banyan)**

Oral administration of bark extract showed anti hyperglycemic effect in STZ diabetic rats by raising serum insulin levels or inhibiting insulin activity in liver and kidney. Oral administration of leucopelargonidin derivative isolated from bark of *F. bengalensis* exerts significant hypoglycemic activity in normal and moderately alloxanized diabetic dogs (Augusti et al., 1994). A leucocyanidin derivative isolated from the bark of *F. bengalensis* was hypoglycemic in normal rats.

***Gymnema sylvestre* (Gurmar)**

The leaves of *Gymnema sylvestre* find use as anti diabetic remedy in ancient Indian medicine since 2000 years. The extract of leaf inhibited glucose absorption in small intestine of rats. Early research suggests when a specific gymnema extract (GS4) is taken orally along with insulin or diabetes medications, blood sugar reduction in people with type 1 or type 2 diabetes is enhanced.

***Murraya koenigii* (Curry tree)**

It is one of the most widely acclaimed remedies for the treatment of diabetes.

***Momordica charantia* (Bitter gourd)**

Fruits of *Momordica charantia* have been successfully used by diabetic patients and their crude extract has been shown to possess hypoglycaemic activity. Khanna and Jain isolated a hypoglycaemic peptide (polypeptide-P) from seeds and other tissues of *Momordica charantia*. They reported that polypeptide-p is a very effective hypoglycaemic agent when administered subcutaneously to langurs and humans. Singh et al. have reported hypoglycaemic effect of acetone extract of whole fruit powder of *Momordica charantia*.

Mechanisms of action of herbal anti-diabetics

Certain herbs may lower blood glucose, however, their test results are subject to several factors. Firstly, each herb contains thousands of components, only a few of which may be therapeutically effective. Secondly, different parts of an herb have different ingredient profiles. Moreover, different extraction methods may yield different active ingredients. Thirdly, herbal formulae containing multiple herbs may have synergistic effects. The attributed anti hyperglycemic effects of these plants are due to their ability to restore the function of pancreatic tissues by causing an increase in insulin output or a decrease in the intestinal absorption of glucose. Hence, treatment with herbal drugs has an effect on protecting β -cells and smoothing out fluctuation in glucose levels. In general, there is very little biological knowledge on the specific modes of action in the treatment of diabetes, but most of the plants have been found to contain substances like glycosides, alkaloids, terpenoids, flavonoids etc. that are frequently implicated as having anti diabetic effects. Herbal anti-diabetics are found to have properties of releasing of insulin from islet of langerhans and as an insulin sensitizer.

Conclusion

Diabetes is a metabolic disease that has serious problem of modern society due to the severe long-term health complications associated with it. In spite of known anti

diabetic medicine in the pharmaceutical market, remedies from medicinal plants are used with success to treat this disease. Many traditional plant treatments for diabetes are used throughout the world. Plant drugs and herbal formulations are frequently considered to be less toxic and free from side effects than synthetic ones. The research for alternate remedies (from the plant kingdom) for diabetes mellitus will continue all over the world as the disease poses many challenges not only to the physician but also to the researcher.

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Study of Aquatic Toxicity Risks of Agricultural Pesticides on Non – Target Organism (*Labeo Rohita*)

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Abstract

Aquatic toxicity risks of agricultural pesticides to non-target organisms specially fishes are pivotal. The use of pesticides for an effective control of plant diseases has become crucial in the last decades in the agriculture system since it is estimated that pest infections cause yield reductions of almost 20% of crops worldwide. In the last decades, the use of pesticides in agriculture for disease control has become crucial. The present study was carried out to investigate the LC₅₀ of two different pesticides such as monocrotophos and lambda cyhalothrin on the freshwater fish *Labeo rohita*. Monocrotophos caused 100% mortality of *L. rohita* at 0.0044 ppm and 50% mortality. The LC₅₀ values obtained at 24, 48, 72 and 96 hours exposures and the 95% confidence limits for the two pesticides revealed that lambda cyhalothrin showed higher toxicity than monocrotophos.

Keywords: Monocrotophos, lambda cyhalothrin, mortality, *Labeo rohita* and non – target organism.

Introduction

The birth of the modern pesticide era in the late 1940's was hailed as a major breakthrough for mankind. The philosophy that these new chemicals would stop the innumerable pests in their track there by eradicating disease and eliminating crop reduction led to a progressive increase in their use. A belated second thought on the environmental persistence of organochlorines (new era pesticides) led to the invention of less persistent pesticides like organophosphate, carbamate and synthetic pyrethroids. The problem of pesticidal impact on the ecosystem has assumed considerable proportions owing to the modernisation of agricultural operations and the consequent widespread and indiscriminate permeation of the ecosystem with these pesticides. The effects of pesticides on aquatic fauna, particularly fishes, may be exhibited in a variety of ways, since the majority of them are non—selective and produce detrimental and sometimes fatal side effects on non-target species. Knowledge on pesticide toxicity levels, either by acute toxicity, residual or physiological studies, is essential to develop effective protective measures for time conservation. of our already depleted freshwater fauna.

It is apparent that human chemical additions have introduced or increased environmental stress for aquatic organisms and fishes in particular. Many of the effects of pesticides to fishes are subtle and insidious. Unlike direct eradication of populations (e.g. fish kills), the more serious long - term decline of stocks of fish are caused by indirect factors such

as predation, disease and reproductive failure. Fish which are subjected to unnatural stresses in any part of their life history may be rendered less capable of performing those functions necessary to fulfil their life cycle and if fishes ability to defeat its natal stream is impaired by the presence of pollutants, then it may go unspawned and leave no natural means of perpetuating the species.

Studies on the sublethal effects of pesticides have gained a great deal of impetus in the last decade, partly because of their practical importance and partly owing to academic interest. Quantitative: assessment of the effects of pesticides on fishes has got cardinal importance in fishery management both from the biological and ecological points of view. Moreover, sublethal effects of pollutants are now being recognized by regulatory agencies in establishing pollution controls. Rather than applying an arbitrary "application factor", as a safety factor, to the LC₅₀ data obtained in acute toxicity bioassays, pollution control is now being developed by using the sublethal threshold level, derived by chronic toxicity bioassays, as the limiting concentration. Even in administering the international Convention for the Prevention of Marine Pollution by Dumping of Wastes and Other Matter (U.K. 1972), the term "harmlessness" of a particular substance is being defined by application of data from sublethal toxicity studies, among others. Dermal acute toxicity tests represent an important method for establishing criteria to evaluate water quality and therein to protect the aquatic environments (NAS/NAE, 1972). Acute toxicity

studies are generally employed to compare the sensitivities of different species to different potencies of the chemicals and to derive, by using LC_{50} values, environmental concentration of chemicals which could be considered 'safe'.

Toxicity data for a variety of pesticides such as organophosphate, organochlorine, carbamide and pyrethroid pesticides have been reported for number of fish species by various authors (Anees, 1975; Baskaran *et al.*, 1989; Roy and Nishar Shailkh and Yeragi, 2004). Variations in LC_{50} if any, under different energy balance should reflect on the nutritional status of the animal. This would help to assess the productivity of the aquatic medium concerned and to take appropriate corrective measures. The present study has been carried out to find the lethal concentration 50% (LC_{50}) of pesticides monocrotophos and lambda cyhalothrin on the freshwater fish *Labeo rohita*.

Method and Materials

The freshwater healthy fish, *L. rohita* of the weight nearly 1kg were selected for the experiment and were collected from local Darbhanga market. Fish were screened for any pathogenic infections. Glass aquaria were washed with 1% $KMnO_4$ for two minutes to avoid fungal contamination and then sun dried. Healthy fishes were then transferred to glass aquaria containing dechlorinated tap water. Fish were acclimated to laboratory conditions for 10 to 15 days prior to experimentation. They were regularly fed with commercial food and the medium (tap water) was changed daily to remove faeces and food remnants.

Toxicity tests were conducted in accordance with standard methods (APHA, 1992). Stock solution of monocrotophos with a concentration of 1 ml per litre (equivalent to 1 ppt) was prepared in distilled water and different dilutions were prepared by adding required amount of distilled water. The stock solution of lambda cyhalothrin with a concentration of 1 ml per litre was also prepared in distilled water and the desired degree of concentrations was prepared. Based on the progressive bisection of intervals on a logarithmic scale, log concentrations were fixed after conducting the range finding test. The fish were starving for 24 hours prior to their use in the experiments as recommended by storage to avoid any interference in the toxicity of pesticides. After the addition of the toxicant into the test tank with 10 litres of water having twenty fish, mortality was recorded after 24, 48, 72 and 96 hours. Five replicates were maintained simultaneously. Percent mortality was calculated and the values were transferred into probit scale. Probit analysis was carried out as suggested by Finney (1971). Regression lines of probit against logarithmic transformations of concentrations were

made. Confidential limits (upper and lower) of the regression line with chi-square test were calculated by a computerized programme for Finney's (1971) probit analysis.

Result and Discussion

Fishes exposed to lethal concentration of agricultural effluent for a short-term exposure were studied in terms of general behavior, rate of survival and mortality. The effects of pollutants are generally characterized on survival, reproduction or growth due to physiological alteration in the animal. The physical, chemical and biological components of the environment play an important role in manifestation of biological response to pollutants.

Monocrotophos caused 100 % mortality of *L. rohita* at 0.0044 ppm and 50 % mortality (96 hours) at 0.0036 ppm, and for lambda cyhalothrin, the lethal effect was at 0.0029 ppm and LC_{50} at 0.0021 ppm. The LC_{50} values obtained at 24, 48, 72 and 96 hours exposures and the 95% confidence limits for the two pesticides revealed that lambda cyhalothrin showed higher toxicity than monocrotophos. The LC_{50} values of monocrotophos for 24, 48, 72 and 96 hours were 0.0041, 0.0039, 0.0037 and 0.0036 ppm respectively (Table 1), whereas the LC_{50} value of lambda cyhalothrin for 24, 48, 72 and 96 hours were 0.0026, 0.0024, 0.0022 and 0.0021 ppm, respectively (Table 2).

Exposure Hour	LC_{50}	U. C. L	χ^2
24	0.004084	0.004167	8.2512
48	0.003870	0.003946	9.0221
72	0.003726	0.003791	3.5032
96	0.003556	0.003611	6.7556

Table 1

Exposure Hour	LC_{50}	U. C. L	χ^2
24	0.002575	0.002679	5.6685
48	0.002370	0.002457	8.8858
72	0.002156	0.002298	17.6868
96	0.002030	0.002095	8.9140

Table 2

When the fish was exposed to pesticides monocrotophos and lambda cyhalothrin, erratic swimming, abnormal posture, misbalance, sluggishness, imbalance in posture, increase in surface activity, opercular movement, gradual loss of equilibrium and spreading of excess of mucus all over the surface of the body were observed. Fish exposed to sub lethal concentrations of pesticides irregular, erratic and darting movements with imbalanced swimming activity and attempt

to jump out of the toxic medium were observed. Similar behaviour patterns were observed in fish, trout and *L. rohita* exposed to fenvalerate (Murthy, 1987). Increased opercular movements, loss of equilibrium, erratic swimming and jerky movement and mucous secretion all over the body were observed. Increased opercular movements were seen in the fish, *L. rohita* exposed to pesticides, which was in accordance to the report put forth by Amita kiran and Jha (2009) in *Clarias batrachus* exposed to herbicide, herboclin. The change of body colour, behavioural changes such as irregular swimming movements, loss of equilibrium, restlessness and excess secretion of mucous suggest that *L. rohita* has undergone chemical stress when exposed to pesticide and the present study could be taken as an indicator of aquatic pollution.

Conclusion

Aquatic toxicity risks of agricultural pesticides to non-target organisms specially fishes are pivotal. The use of pesticides for an effective control of plant diseases has become crucial in the last decades in the agriculture system since it is estimated that pest infections cause yield reductions of almost 20% of crops worldwide. In the last decades, the use of pesticides in agriculture for disease control has become crucial. As known, fungicides, pesticides and other chemicals can enter to water bodies through rain water, food, irrigation water or rivers in many cases and may be hazardous for living systems. These chemicals influence metabolism at very low concentrations by altering enzyme activities and disrupting physiological balances.

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Ethnobotanical Medicinal Utility of *Leucas aspera* (Willd) Linn.

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Abstract

Plants of species *Leucas aspera* (Labiatae) are known for its varied medicinal uses. Various plant parts and there ethno-medicinal utility is presented in the article.

Keywords: *Leucas*, Labiatae, Ethno - medicinal information.

Introduction

Leucas aspera (willd) Linn. belongs to the family Labiatae is a small herbaceous, erect, plant with a free blooming nature and flowering in the month of March to October. It is pungently aromatic herb and grows abundantly in the high land crop field roadsides. *Leucas aspera* is an annual, branched, herb erecting to a height of 15-60 cm with stout and acutely quadrangular stem and branches. Leaves are sub sessile or shortly petiolate, linear, lanceolate, 8 cm long to 1.5cm broad, with entire or crenate margin. The decoction of the whole plant is traditionally used for analgesic, antipyretic, anti-rheumatic, anti-inflammatory, antibacterial and antifungal treatment and its paste is applied typically to inflamed areas. It is also used as medicine for coughs, colds, painful swellings and chronic skin eruptions, cobra venom poisoning, tooth decay infections and mosquito repellent by the rural people and possesses wound healing properties.

Material and Methods

During the intensed field work my prime focus was that to gain the information regarding the medicinal utility of the plant being used by the people since old time.

In Madhubani district I visited following villages:

- 1) Brahmottra, Pandaul, Madhubani
- 2) Baadagaon, Pandaul, Madhubanai
- 3) Mangrauni, Madhubani
- 4) Navhad, Rampatti, Madhubani

There was some common information regarding the medicinal utility of the plant *Leucas aspera* given by the peoples are as follows:

Traditional Use of Selected Species & Ethnomedical Information

- 1) Used orally for the treatment of headache, asthma and bronchitis.
- 2) Hot water extract of entire plant is used to treat inflammation, dysleptia and jaundice.
- 3) Entire plant extract is used orally to treat scabies, psoriasis and snake bite.
- 4) The plant *Leucas aspera* is externally used as an insect repellent.
- 5) *Leucas aspera* and *Ocimum canum* are externally used to fumigate swelling/dwelling.
- 6) Flowers roasted in ghee are given orally (5-10g once a day) for treatment of cough and colds.
- 7) The flowers are crushed and aroma is inhaled in the opposite nostrils for the relief of migraine.
- 8) The juice of used aurally for ear pain for pus discharge from ear.
- 9) The paste of leaves ground with chalk is applied to tooth cavity to prevent decay.
- 10) The decoction of leaves is used nasally as an antivenin.
- 11) Infusion of leaves is used externally to treat scabies.
- 12) Leaf paste mixed with turmeric is used to heal wounds and boils.
- 13) Decoction of roots, stems, and inflorescence of *Leucas aspera*, stem of *Phylla nodiflora* and roots of *Ocimum gratissimum* are used orally for high fever for influenza and malarial fever.

Table - 1

Plant Species	Part	Mode	Indication	Reference
<i>L. aspera</i>	Leaf	Overnight in hollow chilly fruit case	Madras eye condition	Sandhya <i>et. al</i> 2006
	Whole Plant, Root	Paste	Antidote to snake bite, in fever, headache & insect & Lights, asthma.	Singh <i>et. al</i> 2002
	Leaf	Warmed leaf juice	Erache and arthritic pain	Yuouf <i>et. al</i> 2006
	Leaf	Boiled leaves vapors	Inhaled to relieve cough and colds	Ganeshan <i>et. at</i> 2007
	Leaf	Paste	Skin diseases, dhobis itch & ringworm	Mahishi <i>et. at</i> 2005
	Leaf	Crushed leaf	Snake bite, one side, headache	Revathi and Primelazhahan 2010

Conclusion

On the basis of the feedback and general overview of this species *Leucas aspera*, It shows that genus *Leucas* being used traditionally due to its immense therapeutic potential to treat/cure various diseases. Every part of the *Leucas* is being used for treating various diseases.

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Glycogen Effects of Herbal Drugs on the Liver and Muscle of Male Albino Rat

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Abstract

This study attempts to evaluate the health status of the albino rat. On the basis of change in tissue biochemistry under the effects of herbal drugs (Urhul and ghrit kumara). Urhul and ghrit kumari orally fed for 20 days. Significant changes were observed in glycogen of liver and muscle.

Keywords: Glycogen effects, herbal drugs, liver, muscles and albino rat.

Introduction

A number of plants are known for their medicinal values. People use herbal drugs to maintain or improve their health. Herbal medicine is also called botanical medicine or phyto – medicine. It is a medical system based on the use of plant or plant extract that may be eaten or applied to skin. Some herbs have potent ingredients and should be treated with same care and respect as pharmaceutical drugs.

In India today the over population has been serious concern. Fertility control programme has gained tremendous importance in medical science at present. There is great need to evaluate the medicinal properties of herbal drugs and their use mainly as contraceptive herbal drugs. In old days people were using herbal drugs and persons were expert enough to know and to identify those plants. Rosa sinensis which is commonly called urhul, jasmine officinate (chambeli), Aloe Vera, ghrit kumari may be contraceptive. In any living tissue, herbal drugs exert their effects first at molecular and biochemical level hence alternation in normal biochemical parameters serve as the earliest indicator of herbal drugs effect on the tissue. Herbal drugs changing the biochemical composition and by altering physiological phenomenon in rats (Mishra 1995, Singh 1997, Alam & Kumari 2009, Prabhakar and Kumari 2010 and Anwar and Choudhary 2011).

Material and Method

Mature male rat of equal weight and age were selected for experiments. After proper acclimatization to laboratory condition rats were divided in following two groups.

Group 1. 10 Albino rat kept as control were fed with normal pellet diet.

Group 2. The second group contain 10 Albino rat were housed into plastic cage with standard pellet diet mixed with 1.38 kg weight of herbal drugs (urhul and ghrit kumari) for

20 days. At the end of exposure period the rats were anaesthetized and processed for estimation of glycogen.

Table 1: Showing the effect of urhul and ghrit kumari (20 days) on liver and muscles glycogen level of male rat

Parameters	Control	Effect of Urhul	Effect of Ghrit Kumari
Liver	45.41 ± 1.2	28.1 ± 1.2	54.68 ± 1.2
Muscle	30.84 ± 1.5	18.62 ± 1.2	40.3 ± 1.2

Result

The glycogen profile of liver in response to urhul and ghrit kumari exposure showed significant decrease and increase while in muscle, it showed incline in ghrit kumara. Liver showed 30 % decline in urhul while 20 % incline in glycogen ghrit kumari while muscle showed 39 % decline in urhul while 30 % incline in glycogen content in ghrit kumari. The glycogen content in control liver and muscle were estimated as 45.41 ± 1.2 mg/g and 30.84 ± 1.5 mg/g as against urhul extract were 28.1 ± 1.2 mg/g and 18.62 ± 1.2 mg/g while under ghrit kumari extract values were 54.68 ± 1.2 and 40.3 ± 1.2 mg/g.

Discussion

In the present investigation the significant depletion in the amount of liver, muscle glycogen in urhul feeding male rat. In the muscle there is significant depletion in the urhul feeding but there is significant incline in the ghrit kumara feeding male rat under 20 days. The present findings are conformity with (Thorpe et al 1964, Susan et al 1989, Singh 1997, Alam & Kumari 2009, Prabhakar & Kumari 2010 and Anwar & Choudhary 2011).

In the present investigation the significant depletion in liver and muscles glycogen on the effect of urhul feeding male rat might be due to hyper activity of the cells of pancreatic islets of langerhans which elaborate great amount of glycogen under 6the influence of urhul.

The glucagon hormone is known to estimate the process of glycogenesis leading to elevate blood glucose level. In the present investigation these significant inclination in muscle of ghrir kumari feeding male rat less than 20 days treatment is due to increase secretion of insulin by β cells of islet (Yadav 1987). Insulin has been known to cause glycogenesis leading to hypoglycaemia in blood and deposited glycogen in liver and muscle. The other possible reason behind the depilation of glycogen in muscle is due to hyper – activity of muscle under the influence of urhul. For the moment it is sufficient that in the process of muscular contraction, glycogen is broken down to pyruvic acid which in strenuous activity may be partly converted into lactic acid. In mammalian muscle it is probable that the lactic acid formed in this way diffuse into the blood particularly if the muscle contract repeatedly without full time of recovery (Thorpe et al 1964). This lactic acid is carried to the liver and is resynthesised to glycogen.

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Extension of Invariant Structure

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Abstract

The aim of this paper is the deep extension towards the findings of structure on algebra which becomes a role model for the way out to reach its goal. Also our main aim is to extension of Frattini's structure

Keywords: Structure, algebra, extension, model, Frattini.

Introduction

(1.1): It is very well known in the group theory that structure plays a vital role in the extension of modern abstract mathematics. It is concerned that in an abelian group every sub group is a normal subgroup and also in a finite abelian group every element generates a normal subgroup. It is also possible that if an element of prime order then we must obtain a normal sub group of prime order and hence every finite abelian group has a normal chain with factors all of prime order then we obtain a composition services.

$$G = G_0 \supset G_1 \supset G_2 \supset \dots \supset G_U \supset 1$$

Also, it has been established than every finitely generated abelian group is a direct product of cyclic groups of prime power order.

Every finite abelian group can be expressed as the direct sum of cyclic groups. If G be any group $\Psi(G)$ the intersection of all its maximal subgroups and if there are no such subgroups then $\Psi(G) = G$.

Hence, $\Psi(G)$ is invariant under all auto morphisms of group G which is also known as Frattini sub group.

Theorem 1.2: Let G be any group. Then $\Phi(G)$ consists precisely of all the non-generators of G .

Proof: Let $c \in G$ be a non-generator. If M is a maximal subgroup, then $c \in M$, for otherwise the subgroup generated by $M \cup \{c\}$ would be larger than M , hence equal to G , and so by definition of c as non-generator, $M = G$, which is false. So c lies in every maximal subgroup, and hence also in $\Phi(G)$.

Converse part, if $c \in \Phi(G)$ and $X \cup \{c\}$ generates G , we have to show that X generates G . if this is not so, then $\langle X \rangle$ is proper in G , and hence, by Zorn's lemma, there is a subgroup M containing X but not c and maximal with those properties. Since $c \notin M$, $M \neq G$, but any subgroup properly containing M must also contain c and hence be equal to G . Hence M is maximal in G , but $c \in \Phi(G) \subseteq M$, which is a contradiction. Hence X generates G and c is indeed a non-generator.

Theorem (1.3): Let G be any group and N be a finitely generated normal subgroup. Then $\Phi(N) \subseteq \Phi(G)$.

Proof: Since $\Phi(N)$ is a characteristic subgroup of N , hence normal in G . Now let M be a maximal subgroup of G and

suppose that $\Phi(N) \not\subseteq M$. Then $\Phi(N)M = G$, hence by the modular law, since $\Phi(N) \subseteq N$.

$$N = \Phi(N)M \cap N = \Phi(N)(M \cap N).$$

So $M \cap N$ together with $\Phi(N)$ generates N ; hence $M \cap N = N$ i.e. $M \supseteq N$ and so $\Phi(N) \subseteq M$, is a contradiction. This follows that $\Phi(N) \subseteq M$ for any maximal M , and so we find that $\Phi(N) \subseteq \Phi(G)$.

Theorem (1.4) : Let G be a finite group and N, K be normal subgroup such that $N \subseteq K \cap \Phi(G)$. If K/N is nilpotent, then SO IS K .

Proof: Let P be a Sylow p – subgroup of K . Then PN / N is a Sylow p – subgroup of K/N because $PN/N \cong P / (P \cap N)$ and this again has p -power order, while its index in K is prime to p . Since K/N is nilpotent, PN / N is normal in K/N , therefore characteristic in K/N and so normal in G/N , hence $PN \triangleleft G$. Now P is a Sylow p -subgroup of PN ; hence we have

$$G = PN.N_G(P) = N.N_G(P)$$

Now $N \subseteq \Phi(G)$ & so $N_G(P) = G$ and so $P \triangleleft G$ hence $P \triangleleft K$ and this holds for all Sylow subgroups of K , so K is nilpotent. Then a finite group G is nilpotent if and only if $G' \subseteq \Phi(G)$.

As the Frattini subgroup may be thought of as a kind of radical, in analogy to the Jacobson radical. Whereas the Jacobson radical is the largest for all the commonly used radicals in a ring, this is not so for the Frattini subgroup. Radicals are not as useful for groups as for rings; this is not so for the Frattini subgroup. Radicals are not as useful for groups as for rings; nevertheless there is another type, first introduced by Hans Fitting in 1938, which plays a vital role in group theory.

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Group Structure with Symplectic Group

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Abstract

As group theory evolved from many folds in the fabric of mathematics. The group of transformation was extended and generalized to the concept of an abstract group by imposing a set of axioms with a single binary operation. The mathematician attempted to describe all geometrics by their group of symmetric. Our aim here is to develop the algebraic structure with symplectic group.

Keywords: Group theory, transformation, abstract group, axioms, binary operation, geometric, structure, symplectic group.

Introduction

(1.1) – If G be a group and $\phi(G)$ the intersection of all its maximal sub groups and there are no subgroups such that $\phi(G) = G$ then it is quite clear that $\phi(G)$ is a subgroup of G . Which is invariant under all automorphism of G and hence it is known as Frattini subgroup. It is also clear than any non-trivial finite group finite group has maximal subgroups so $\phi(G) \subset G$. We know that Frattini subgroups of any finite group is nilpotent.

By a symplectic space over any field K which means a vector space V over K with a regular bilinear form b which is alternating such than $b(X, x) = 0, \forall x \in V$. By linearilization we find that b is autosymmetric as $b(x, y) = -b(y, x)$ relative to basis of V , B is sview symmetric matrix. Since the space is regular so $|A| \neq 0$, A is a matrix symplectic spaces are of importance in the Hamiltonian theory of dynamics & we find that every symplectic space is even basis consists of mutually orthogonal hyperbolic pairs.

Theorem 1. 2

Let G be any group. Then $\phi(G)$ consists precisely of all the non generators of G .

Proof : Let $c \in G$ be a non-generator. If M is a maximal subgroup, then $c \in M$, for otherwise the subgroup generated by $M \cup \{c\}$ be larger than M , hence equal to G , and so by definition of c as non-generator, $M = G$, which is false. So c lies in every maximal subgroup, and hence in $\phi(G)$.

Conversely, if $c \in \phi(G)$ and $X \cup \{c\}$ generates G , we have to show that X generates G . If this is not so, then $\text{gp}\{X\}$ is proper in G , and hence, by Zorn's lemma, there is a subgroup M containing X but not c and maximal with those properties. Since $c \notin M$, $M \neq G$, but any subgroup properly containing M must also contain c and hence be equal to G . So M is maximal in G , but $c \in \phi(G) \subseteq M$, which is a contradiction. Hence X generates G and c is indeed a non-generator.

Theorem 1.3

Let G be a group and U be subgroup. Then $U \subseteq \phi(G)$ whenever U satisfies the following condition :

For any proper subgroup H of G , $\text{gp}\{U, H\} \subset G$ (I)
Where G is finitely generated, this condition is necessary as well as sufficient. And we have $H\phi(G) \subset G$ for every proper subgroup H of G .

Proof : If I holds, then for any maximal subgroup M of (G) , $M \subseteq \text{gp}\{U, M\} \subset G$, hence $M = \text{gp}\{U, M\}$ and so $U \subseteq M$; so $U \subseteq \phi(G)$. If G is finitely generated and if $U \subseteq \phi(G)$, and H , but also $U \subseteq M$, hence $\text{gp}\{U, H\} \subseteq M \subset G$, Hence the theorem proved.

Theorem 1. 4

Let G be any group N be a finitely generated normal subgroup. Then $\phi(N) \subseteq \phi(G)$.

Proof: $\phi(N)$ is a characteristic subgroup of N , hence normal in G . Now let M be a maximal subgroup of G and suppose that $\phi(N) \not\subseteq M$. Then $\phi(N)M = G$, hence by the modular law, since $\phi(N) \subseteq N$,

$$N = \phi(N) M \cap N = \phi(N) (M \cap N).$$

So $M \cap N$ together with $\phi(N)$ generates N ; so $M \cap N = N$, i.e. $M \supseteq N$ and so $\phi(N) \subseteq M$ is a contradiction. Thus follows that $\phi(N) \subseteq M$ for any maximal M , and so we find that $\phi(N) \subseteq \phi(G)$.

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Contributions of the History of Ancient Indian Mathematics on Universe

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Abstract

Indian mathematics emerged in the Indian subcontinent from 1200 BCE until the end of the 18th century. In the classical period of Indian mathematics (400 CE to 1600 CE), important contributions were made by scholars like Aryabhata, Brahmgupta, Mahavira, Bhaskara II, Madhava of Sangamagrama and Nilakantha Somayaji. The decimal number system in worldwide use today was first recorded in Indian mathematics. Indian mathematicians made early contributions to the study of the concept of zero as a number, negative numbers, arithmetic, and algebra. In addition, trigonometry was further advanced in India, and, in particular, the modern definitions of sine and cosine were developed there. These mathematical concepts were transmitted to the Middle East, China, and Europe and led to further developments that now form the foundations of many areas of mathematics. There was an awareness of ancient Indian mathematics in the West since the sixteenth century; historians discuss the Indian mathematical tradition only after the publication of the first translations by Colebrooke in 1817. Its reception cannot be comprehended without accounting for the way new European mathematics was shaped by Renaissance humanist writings. We show by means of a case study on the algebraic solutions to a linear problem how the understanding and appreciation of Indian mathematics was deeply influenced by humanist prejudice that all higher intellectual culture, in particular all science, had risen from Greek soil.

Keywords: Indian mathematics, renaissance, humanism, Greek, Indian influences

Introduction

Indian mathematics emerged in the Indian subcontinent from 1200 BCE until the end of the 18th century. In the classical period of Indian mathematics (400 CE to 1600 CE), important contributions were made by scholars like Aryabhata, Brahmgupta, Mahavira, Bhaskara II, Madhava of Sangamagrama and Nilakantha Somayaji. The decimal number system in worldwide use today was first recorded in Indian mathematics. Indian mathematicians made early contributions to the study of the concept of zero as a number, negative numbers, arithmetic, and algebra. In addition, trigonometry was further advanced in India, and, in particular, the modern definitions of sine and cosine were developed there. These mathematical concepts were transmitted to the Middle East, China, and Europe and led to further developments that now form the foundations of many areas of mathematics. While there was an awareness of ancient Indian mathematics in the West since the sixteenth century, historians discuss the Indian mathematical tradition only after the publication of the first translations by Colebrooke in 1817. Its reception cannot be comprehended without accounting for the way new European mathematics was shaped by Renaissance humanist writings. Western reception of ancient Indian mathematics during the nineteenth century is very much biased by the humanist tradition. Reflections and statements of Western historians

on Indian mathematics can only be fully understood if this context is known and acknowledged. During the Middle Ages mathematics was hardly practiced or appreciated by the intellectual elite. The middle ages knew two traditions of mathematical practice. On the one hand, there was the scholarly tradition of arithmetic theory, taught at universities as part of the quadrivium. The basic text on arithmetic, presented as one of the seven liberal arts, was Boethius's *De Institutione Arithmetica* (Friedlein, 1867). The Boethian arithmetic strongly relies on Nichomachus of Gerasa's *Arithmetica* from the 2nd century. This basically qualitative arithmetic deals with properties of numbers and ratios. All ratios have a name and operations or propositions on ratios are expressed in a purely rhetorical form. The qualitative aspect is well illustrated by the following proposition from Jordanus de Nemore's *De elementis arithmetice artis* (c. 1250, Book IX, proposition LXXI; Busard 1991, 199). During the eleventh century a board game named *Rhythmomachia* was designed to meet with these aesthetic aspirations. Originated as the subject of a competition on the knowledge of Boethian arithmetic amongst cathedral schools in Germany, the game was played until the sixteenth century, when the arithmetic tradition passed into oblivion. Despite its limited applicability, Boethian arithmetic evolved into a specific kind of mathematics, typical for the European Middle Ages, and left its mark on early natural philosophy.

Arithmetical problem solving became much more advanced with the introduction of Arabic algebra through the Latin translations of al Khwarizmi's Algebra by Robert of Chester (c. 1145), Gerard of Cremona (c. 1150) and Guglielmo de Lunis (c. 1215). During the fifteenth century Italian humanists eagerly started collecting editions of Greek mathematics; one of the most industrious was Cardinal Bessarion. In some sense Wallis's Treatise on Algebra (1685) can be considered the first serious historical investigation of the history of algebra. John Wallis was well informed about Arabic writings through Vossius and was one of the first to attribute correctly the name algebra to al - jabr in Kitab fi al - jabr wa'l - muqabala.

He also pointed out the mistaken origin of algebra as Geber's name, which was a common misconception before the seventeenth century. So, while in the seventeenth century no Sanskrit mathematics had yet been introduced into Europe, scholars by then were aware of the existence of Indian algebra. Wallis's view persisted in eighteenth-century historical studies, which reiterated the influence from Indian mathematics. In the early nineteenth century, the English orientalist Henry Thomas Colebrooke, who previously published his Sanskrit Grammar (1805), undertook the task of translating three classics of Indian mathematics, the Brahmasphuta Siddhanta of Brahmagupta (628) and the Lilavati and the Bijaganita of Bhaskara II (1150). At once European historians had something to reflect upon. In a period when mathematics was hardly practiced in Europe and in the Islam regions, there appeared to have existed this Indian tradition in which algebraic problems were solved with multiple unknowns, in which zero and negative quantities were accepted and in which sophisticated methods were used to solve indeterminate methods.

In general, nineteenth-century historians showed an admiration for the Hindu tradition. However, whenever explanations were required, scholars became divided into two opposing camps, Non believers did not grant Indian mathematicians the status of original thought. Indian knowledge must have stemmed from the Greeks, the cradle of Western mathematics, or even mathematics as such. The major non-believer was Moritz Cantor who published an influential four-volume work on the history of mathematics (1880-1908). Cantor (1894, II) takes every opportunity to point out the Greek influences on Hindu algebra. Soon after Kern (1875) published the Sanskrit edition of the Aryabhata's (AB), the French orientalist L'eon Rodet was the first to provide a translation in a Western language (1877, published in Rodet 1879). Rodet wrote several articles and monographs on Indian mathematics and its relation with earlier and later developments in the Arab and Western world, published in the French Journal Asiatiques. He is the scholar who displays the most balanced and subtle views on

the relations between traditions. In particular, his appraisal of Hindu and Arabic algebra as two independent traditions is still of value today. He certainly was a believer. Concerning Aryabhata's inadequate approximation of the volume of a sphere (prop. 7), he writes somewhat cynically that if Aryabhata got his knowledge from the Greeks, then apparently he chose to ignore Archimedes ("Mais elle a, pour l'histoire des mathématiques, d'autant plus de valeur, parce qu'elle nous démontre que si Aryabhata avait reçu quelque enseignement des Grecs, il ignorait au moins les travaux d'Archimède", Rodet 1879, 409). George Thibaut who translated several Sanskrit works on astronomy, such as Varahamihira's Pancasiddhantika (1889), also wrote an article on Indian mathematics and astronomy in the Encyclopedia of Indo-Aryan Research. Concerning influences from Greek mathematics, he takes a middle position. In discussing Hindu algebra he writes that "in all these correspondences does Indian algebra surpass Diophantus" ("In allen diesen Beziehungen erhebt sich die indische algebra erheblich über das von Diophant Geleistete". As on the origins of Indian mathematics, he points out that Indian algebra, especially indeterminate analysis, is closely intertwined with its astronomy. As he argued on the Greek roots of Indian "scientific" astronomy, his evaluation is that Indian mathematics is influenced by the Greeks through astronomy. However, he adds that several arithmetical and algebraic methods are truly Indian.

The first Indian source for a formulation of this rule is from Aryabhata I, as follows: If you know the results obtained by subtracting successively from a sum of quantities each one of these quantities set these results down separately. Add them all together and divide by the number of terms less one. The result will be the sum of all the quantities. The rule is somewhat obscure and difficult to understand without examples, but some observations can be drawn from the formulation which is central to our further discussion. Firstly, the rule is valid for any number of quantities. It is not limited to two or three quantities. Secondly, the sum of all the quantities is unknown and is provided by the rule. Furthermore, and not evident from the rule, as cited above, is that the partial sums relate to the total of all the quantities, except one. For example suppose n amounts ($a_1, a_2, \dots, a_n$) with unknown sum S and with the partial sums ($s_1, s_2, \dots, s_n$) given, where $s_i = S - a_i$, then $S = \frac{\sum_{i=1}^n s_i}{n-1}$

The rule and the problems it applies to should not be confused with a similar problem in which the partial sums of two consecutive quantities are given. For three numbers, the problems are evidently the same, but they diverge for more than three quantities. E.g., for five quantities the corresponding equations are:

$$a_1 + a_2 + a_3 + a_4 = s_1$$

$$a_1 + a_2 = s_1$$

$$a_1 + a_3 + a_4 + a_5 = s_2$$

$$a_1 + a_2 + a_4 + a_5 = s_3 \quad \text{and}$$

$$a_1 + a_2 + a_3 + a_5 = s_4$$

$$a_2 + a_3 + a_4 + a_5 = s_5$$

$$a_2 + a_3 = s_2$$

$$a_3 + a_4 = s_3$$

$$a_4 + a_5 = s_4$$

$$a_5 + a_1 = s_5$$

Let us apply the rule to a simple problem (not discussed by Aryabhata which can be formulated symbolically as

$$x_1 + x_2 = 13$$

$$x_2 + x_3 = 14$$

$$x_1 + x_3 = 15$$

Applying Aryabhata's rule, the solution would be based on the rule for deriving the sum of all three unknown quantities

$$\text{as } x_1 + x_2 + x_3 = \frac{13+14+15}{3-1} = 21$$

This allows us to determine the value of the quantities by subtracting the partial sums from the total with the solution (7, 6, 8).

From the ninth century we find a derived version of the previous problem in Hindu sources. Mahavira gives an elaborate description of the rule in the *Ganitasarasamgraha* (GSS, stanza 233-5, Padmavathamma 2000, 357-9).

The rule for arriving at the value of the money contents of a purse which when added to what is on hand with each of certain persons becomes a specified multiple of the sum of what is on hand with the others:

The quantities obtained by adding one to each of the specified multiple numbers in the problem and then multiplying these sums with each other, giving up in each case the sum relating to the particular specified multiple, are to be reduced to their lowest terms by the removal of common factors. These reduced quantities are then to be added. Thereafter the square root of this resulting sum is to be obtained, from which one is [to be subsequently subtracted. Then the reduced quantities referred to above are to be multiplied by this square root as diminished by one. Then these are to be separately subtracted from the sum of those same reduced quantities. Thus the moneys on hand with each of the several persons are arrived at.

These quantities measuring the moneys on hand have to be added to one another, excluding from the addition in each case the value of the money on the hand of one of the persons and the several sums so obtained are to be written down separately. These are then to be respectively multiplied by the specified multiple quantities mentioned above; from the several products so obtained the already found out values of the moneys on hand are to be separately subtracted. Then the same value of the money in the purse is obtained separately in relation to each of the several moneys on hand. The introductory sentence states that the rule is to be used for determining the value of a purse. The rule is followed by a number of problems that begin as "Four men saw on their way a purse containing money" (ibid. stanza 2451 2, 367). This is the earliest instance, in our investigation of the

sources, in which the popular problem of men finding a purse is discussed. While problems with the same structure and numerical values have been formulated before, the context of men finding a purse seems to have originated in India before 850 AD. Formulations with the purse turn up in Arabic algebra with al-Karkhi's *Fakhri* (c. 1050) and in the *Miftah al-muamalat* of al-Tabari (c. 1075). Fibonacci has many variations of it in the *Liber Abbaci* (1202) and after that it becomes the most common problem in western arithmetic until the later sixteenth century. For an understanding of the rule, let us look at its application to a given problem (GSS, stanza 236-7, pp. 360). Three merchants saw [dropped] on the way a purse [containing money]. One [of them] said [to the others], "If I secure this purse, I shall become twice as rich as both of you with your moneys on hand". Then the second of them said, "I shall become three times as rich". Then the other, the third, said, "I shall become five times as rich". What is the value of the money in the purse, as also the money on hand with each of the three merchants?

We can represent the problem in symbolic equations

$$x + p = 2(y + z)$$

$$y + p = 3(x + z)$$

$$z + p = 5(x + y)$$

Let us apply the recipe of Mahavira to this problem, step by step. By "adding one to [each of the specified] multiple numbers" we have 3, 4 and 6. "Multiplying these sums with each other" we get 72. This has to be "reduced to their lowest terms by the removal of common factors". This least common multiple is 12. The reduced quantities are then 4, 3 and 2 respectively. Adding all three together gives 9, from this the square root is 3. Then the reduced quantities "are to be multiplied by the square root as diminished by one", which is 2. This leads to 8, 6 and 4. The money in hand for each of the merchants now is the difference of these values with the sum of the reduced quantities, being 9. The solution thus is 1, 3 and 5. The rest of the rule is an elaborate way to derive the value of the purse. Using the values in any one of the equations immediately leads to 15 for the value of the purse. Mahavira provides no explanation or derivation of the rule. For a mathematical argument for the validity of the rules see Heeffer (2007a).

We know almost nothing about Thymaridas of Paros, but he is supposed to have lived between 400 and 350 BC (Tannery 1887, 385-6). The only extant witness is Iamblichus, in his comments on the *Introduction to Arithmetic* by Nichomachus of Gerasa. The best known source for The Bloom of Thymaridas is Heath's classic on Greek mathematics. Heath (1921, 94) does not formulate the rule, he only observes that "the rule is very obscurely worded" and writes out the equations. The text from Iamblichus was first published in Holland with a Latin translation by Samuel Tennulius (1668) from the Paris manuscript BNF Gr. 2093. A critical edition,

based on multiple manuscripts was published by Pistelli (1884). Nesselmann (1842, 233) quotes the Greek text and the Latin translation from Tennulius, who translated the method as *florida sententia*. We give here the own literal translation from Pistelli (1884, 62). From this we are also acquainted with the method of the *Epanthema*, passed down to us by Thymaridas. Indeed, when a given quantity divides into determined and unknown parts, and the unknown quantity is paired with each of the others, so will the sum of these pairs, diminished by the sum [of all the quantities] be equal to the unknown quantity in case of three quantities. With four quantities it will be half of it, with five it will be a third, with six, a fourth and so on. The rule is not as obscure as considered by Heath. Let us extract the basic elements of the rule, and compare these with the version of Aryabhata:

1. The rule applies to any number of quantities, as does Aryabhata's.
2. The sum is given in the problem. The rule is described as the division of a known quantity in determined and undetermined parts. In Aryabhata's rule the sum is what is looked for.
3. The partial sums are the sums of the pairs of the unknown part with each of the known quantities. In Aryabhata's rule the partial sums include all the numbers except one.

In short, this rule is different from Aryabhata's in two important aspects. Its intention is to find one unknown part of a determined quantity. Aryabhata's rule is meant for finding the sum of numbers of which the partial sums of all minus one is given. Even in the case of three numbers, when the partial sums are the same, the rules have different applications. To make it clear to the modern era, here is a symbolic version in the general case

$$\begin{aligned}
 x + a_1 + a_2 + \dots + a_{n-1} &= s \\
 x + a_1 &= s_1 \\
 x + a_2 &= s_2 \\
 &\vdots \\
 &\vdots \\
 &\vdots \\
 x + a_{n-1} &= s_{n-1}
 \end{aligned}$$

$$x = \frac{\sum_{i=1}^{n-1} s_i - s}{n-2}$$

Conclusion

The humanist project of reviving ancient Greek science and mathematics played a crucial role in the creation of an identity for the European intellectual tradition. While Greek mathematics was hardly known or practiced before the fifteenth century, humanist mathematicians identified themselves with this tradition. When Regiomontanus declared that algebra was invented by Diophantus, humanist

writers rejected the Arabic roots of algebra, though it was practiced and turned into an independent tradition for two centuries in Italian cities such as Florence and Sienna. The newly created identity of mathematics descending from ancient Greek thinkers blurred historical perception. When Indian algebra and arithmetic was introduced into Europe, the leading historians of the nineteenth-century could only see its alleged relation with Greek mathematics. The Bloom of Thymaridas is an excellent illustration of distorted historical investigation. Apparently nineteenth-century historians found it difficult to accept that mathematics is a human intellectual activity endeavoured across cultures within societies that needed and supported the achievements of mathematical practice. A true history of mathematics should take into account contributions of all origins.

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The Role of Modules in Serial Rings

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Abstract

Injective and Projective Modules in a category play a very crucial role in the development of algebra. By imposing monomorphism to homomorphism on the set then we may get different types of modules. Since serial rings provide the relation between the structure of rings and its categories of modules. Also the quotient ring is always serial. Our aim here is to introduce the new concept of algebraic structure so as to generalized the concept of modules in serial rings.

Keywords: Injective and Projective modules, Category, monomorphism, homomorphism, modules, structure, rings, quotient ring, serial rings.

Introduction

A ring R is called left Artinian if it satisfies the *descending chain condition* on left ideals if every descending chain of left ideals $L_1 \supset L_2 \supset L_3 \supset \dots$ becomes stationary after a finite number of steps, i.e. for some $k \in \mathbf{N}$, we find that $L_{k+1} = L_{k+2} = L_{k+3} = \dots$

Similarly, A ring R is called right Artinian if it satisfies the *descending chain condition* on right ideals if every descending chain of right ideals $L_1 \supset L_2 \supset L_3 \supset \dots$ becomes stationary after a finite number of steps, i.e. for some $k \in \mathbf{N}$, we get $L_{k+1} = L_{k+2} = L_{k+3} = \dots$

An Artinian ring is a ring which is left Artinian as well as right Artinian.

As we find that

(i) A finite direct product of Artinian ring is Artinian.

(ii) A quotient of an Artinian ring is Artinian.

(iii) A finitely generated module over an Artinian ring is Artinian.

In a similar way then a ring R is called left Noetherian if it satisfies the *ascending chain condition* on left ideals if every ascending chain of left ideals $L_1 \subset L_2 \subset L_3 \subset \dots$ becomes stationary after a finite number of steps, i.e. for some $k \in \mathbf{N}$, we get $L_{k+1} = L_{k+2} = L_{k+3} = \dots$

Similarly, A ring R is called right Noetherian if it satisfies the *ascending chain condition* on right ideals if every descending chain of right ideals $L_1 \subset L_2 \subset L_3 \subset \dots$ becomes stationary after a finite number of steps, i.e. for some $k \in \mathbf{N}$, we get $L_{k+1} = L_{k+2} = L_{k+3} = \dots$

A Noetherian ring is a ring which is left Noetherian as well as right Noetherian.

As we find the following propositions which are follows

If R contains a subring which is a division ring, and if R is finite dimensional as a left module over the division ring, then R is Artinian and Noetherian.

Any principal ideal ring is Noetherian.

If a ring R is right Artinian, then R is right Noetherian but the converse is not true. For example, let us consider, $m\mathbb{Z} \subseteq n\mathbb{Z}$

$\Leftrightarrow n/m$ and $m_1\mathbb{Z} \subseteq m_2\mathbb{Z} \subseteq m_3\mathbb{Z} \subseteq \dots \Leftrightarrow m_2/m_1, m_3/m_2, \dots$. The chain $2\mathbb{Z} \supset 4\mathbb{Z} \supset 8\mathbb{Z} \supset \dots 2^n\mathbb{Z} \supset \dots$ is not stationary. So $\mathbb{Z}$ is not Artinian. Thus $\mathbb{Z}$ is Noetherian but not Artinian. A right R -module M is Artinian but it need not be

Noetherian. Since We have the Purifier group $\mathbb{Z}_p^\infty$ where $\mathbb{Q}_p = \frac{\mathbb{Q}}{p^i}$ $\{p^i : (a,p)=1 \text{ and } i \in \mathbf{N} \cup \{0\}\} \hookrightarrow \mathbb{Q}\mathbb{Z} \subseteq \mathbb{Q}_1 p \frac{\mathbb{Q}_p}{\mathbb{Z}} =: \mathbb{Z}_p^\infty$

In $\mathbb{Z}_p^\infty$, there are infinite ascending chains of submodules, so $\mathbb{Z}_p^\infty$ does not satisfy the ascending chain condition and consequently, it is not Noetherian. The module $\mathbb{Z}_\mathbb{Z}$ is Noetherian but not Artinian. Indeed, since every ideal of $\mathbb{Z}$ is principal and therefore finitely generated, it is Noetherian. Since the chain $\mathbb{Z} \supset 2\mathbb{Z} \supset 2^2\mathbb{Z} \supset \dots$ is not stationary, we conclude that $\mathbb{Z}_\mathbb{Z}$ is not Artinian. The Purifier group

$\mathbb{Z}_p^\infty = \frac{\mathbb{Q}_p}{\mathbb{Z}}$ is Artinian but not Noetherian. Now, we are going to

define a suitable mapping $\psi: A \rightarrow B$ which surjective from a ring A to a ring B , then (a) If the ring A is Noetherian then the ring B is Noetherian. (b) If the ring A is Artinian then the ring B is Artinian. A non-empty family M of submodules of a right R -module M satisfies the ascending Chain Condition, i.e. if for every chain $M_1 \subset M_2 \subset \dots M_{n-1} \subset M_n \subset \dots$ of elements in M , there is a positive integer n such $M_{n+i} = M_n$ for $i=1, 2$. A right R -module M is called Noetherian if and only if every nonempty family of submodules of M has a maximal element by inclusion. A ring R is called right Noetherian if and only if R_R is a Noetherian right R -module. The ring R is called Noetherian if and only if it is both right and left Noetherian.

A non - empty family M of submodules of a right R -module M satisfies the Descending Chain Condition, i.e. if for every chain

$M_1 \supset M_2 \supset \dots M_{n-1} \supset M_n \supset \dots$ of elements in M , must terminate after a finite number of steps.

A right R -module M is called Artinian if and only if every nonempty family of submodules of M has a minimal element by inclusion. A ring R is called right Artinian if R_R is an Artinian module. The ring R is called Artinian ring if it is both right and left Artinian. In $\mathbb{Z}_p^\infty$, there are infinite ascending chains of submodules

, so $\mathbb{Z}_p^\infty$ does not satisfy the Ascending chain condition and consequently it is not Noetherian. The module $\mathbb{Z}_\mathbb{Z}$ is Noetherian but not Artinian. Indeed, since every ideal of $\mathbb{Z}$ is principal ideal

and therefore finitely generated. Since the chain $\mathbb{Z} \supset 2\mathbb{Z} \supset 2^2\mathbb{Z} \supset \dots$ is not stationary. We conclude that $\mathbb{Z}_{\mathbb{Z}}$ is not Artinian.

Hence the Prüfer group $\mathbb{Z}_p^\infty = \frac{\mathbb{Q}_p}{\mathbb{Z}}$ is Artinian but not Noetherian.

Theorem 1.2

Let M be a right R -module and let A be its submodule. Then the following statements are equivalent:

- (a) M is Artinian;
- (b) A and M/A are Artinian;
- (c) Any descending chain $A_1 \supset A_2 \supset \dots \supset A_n \supset \dots$ of submodules of M is stationary. This condition is called the descending chain condition.

(d) Every factor module of M is finitely co-generated. $0 \rightarrow Af[X] \xrightarrow{i} Y \xrightarrow{v} Y/A \rightarrow 0$

Let $0 \rightarrow X \rightarrow Y \rightarrow Z \rightarrow 0$ be an exact sequence of right R -modules. Then Y is Noetherian $\Leftrightarrow X$ and Z are Noetherian.

Since the image of Artinian module is also Artinian. Since the finite sum of Artinian submodules of M is also Artinian. As the finite direct sum of Artinian modules of M is also Artinian. As if R is semi-simple, then R is both left and right Artinian.

Let L be a submodule of a module M . Then the following conditions are equivalent:

- (a) M is finitely generated then so is M/L ;
- (b) If L and M/L are finitely generated then so is M .

Theorem 1.3

For any R -module M the following properties are equivalent:

- (a) M is Noetherian;
 - (b) The set of finitely generated submodules of M is Noetherian;
 - (c) Every non-empty set of finitely generated submodules of M has a maximal element;
 - (d) Every submodule of M is finitely generated;
- A right R -module M is called *locally Noetherian* if every finitely generated submodule is Noetherian. The following gives some characterizations of locally Noetherian modules.

Theorem 1.4

(a) Let $0 \rightarrow L \rightarrow M \rightarrow N \rightarrow 0$ be an exact sequence in a right R -module M . Then the following conditions are equivalent:

- (i) If M is Noetherian, then L and N are Noetherian;
- (ii) If L and N are Noetherian, then M is Noetherian;
- (iii) If L is Noetherian and N is locally Noetherian, then M is locally Noetherian;
- (b) The direct sum of locally Noetherian modules is locally Noetherian.

Hence, right R -module M is called *locally Artinian* if every finitely generated submodule is Artinian.

Theorem 1.5

(a) Let $0 \rightarrow L \rightarrow M \rightarrow N \rightarrow 0$ be an exact sequence in a right R -module M . Then the following conditions are equivalent:

- (i) If M is Artinian, then L and N are Artinian;
- (ii) If L and N are Artinian, then M is Artinian;
- (iii) If L is Artinian and N is locally Artinian, then M is locally Artinian;
- (b) The direct sum of locally Artinian modules is locally Artinian.
- (c) If M is a locally Artinian right R -module, then

- (i) Every finitely generated right R -module in $\sigma[M]$ is Artinian;
- (ii) $M/\text{Rad}(M)$ is a semisimple module.

Theorem 1.6

For a right R -module M the following conditions are equivalent:

- (a) M is uniserial;
- (b) The cyclic submodules of M are linearly ordered;
- (c) Any submodule of M has at most one maximal submodule;
- (d) For any finitely generated submodule $0 \neq K \subseteq M$, $K/\text{Rad}(K)$ is simple;
- (e) For every factor module L of M $\text{Soc}(L)$ is simple or zero;

Proof

If (a) $\Rightarrow$ (b) is very clear, because every R -module is uniserial.

If (b) $\Rightarrow$ (a). So let K, L be submodules of M with $K \not\subseteq L$ and $L \not\subseteq K$. Let $x \in K \setminus L$ & $y \in L \setminus K$ then we have, $Rx \subseteq Ry$

or $Ry \subseteq Rx$. In the first case we find that $x \in Ry \subseteq L$ and in the second case $y \in Rx \subseteq K$. Both are contradictory

If (a) $\Rightarrow$ (c) and (a) $\Rightarrow$ (d) $\Rightarrow$ (e) are very very clear as submodules and factor modules of M are again Uniserial.

If (d) $\Rightarrow$ (b). Let us suppose that we may find two cyclic submodules $K, L \subseteq N$ with $K \not\subseteq L$ and $L \not\subseteq K$. Then $(K+L)/$

$(K \cap L) \cong K/(K \cap L) \oplus L/(K \cap L)$ and the factor of $(K+L)/(K \cap L)$ by its radical contains at least two simple summands.

Hence, the factor of $(K+L)$ by its radical also contains at least two simple summands. Which Contradicts (d).

Let (e) $\Rightarrow$ (d). We may show that every non-zero finitely generated submodule $K \subseteq N$ contains only one maximal submodule: If $V_1, V_2 \subseteq K$ are different maximal submodules, then

$K/(V_1 \cap V_2) \cong K/V_1 \oplus K/V_2$ is contained in the socle of $N/(V_1 \cap V_2)$. This is a contradiction to (e).

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Algebraic Structure of Serial Rings and Module

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Abstract

A serial ring is in which left & right free modules are the direct sum of modules with unique finite compositions series. Also semi-simple artinian ring is characterized by each of left modules is a direct sum of simple modules. Any left artinian with only finitely many non-isomorphic finitely generated indecomposable left modules. Our aim here is to discuss for the algebraic structure of serial rings and extension theory.

Keywords: Serial ring, module, semi - simple, artinian ring, non - isomorphic, indecomposable, algebraic structure.

Introduction

Module which satisfy the ascending or descending chain condition for ideals was of fundamental importance in classical algebraic geometry and algebraic number theory. Aho module is regarded as a generalization of a vector space. Since any module over a serial ring is a direct sum of uniserial modules which also uses that every module over an artinian principle ideal ring is a direct sum of uniserial modules. If A is a serial ring then it can be decomposed as $A = A_0 \times A_1 \times A_2 \times A_3$ such that.

- (i) A_0 is semi simple & A_i has no semi simple factor for $i > 0$
- (ii) A_i is an artinian principle ideal ring.

Rings have identities and all modules are unital. If A is a ring, $\text{Rad } A$ is its Jacobson radical. A principal indecomposable A -module is an indecomposable direct summand of A . If M is an A -module then $\text{soc } M$ denotes the sum of all simple submodules of M and $\lambda(M)$ denotes the length of a composition series of M . We find that, if A is left artinian and if M is a uniserial left A -module, with $N = \text{Rad } A$, then $M \supset NM \supset N^2M \supset \dots \supset N^k M = 0$ is the form of a unique composition series for M . We know that a QF (Quasi - Frobenius) ring is an artinian ring which is right and left self injective. A ring A is said to be semiprimary if $\text{Rad } A$ is nilpotent and $A / \text{Rad } A$ is semi - simple artinian.

Theorem 1.2: Let A be a ring each of whose finitely generated left modules is a direct sum of uniserial modules. Then A is serial, and thus every left or right A -module is a direct sum of uniserial modules.

Proof: Since A must be left artinian, so this shows that A is right artinian. If M is any finitely generated right module, then $D(M)$ is a direct sum of uniserial left modules. On the

other hand, if U is a uniserial left module, then there is an exact sequence $X \rightarrow Y \rightarrow U$ with X and Y principal indecomposable A -modules. Hence, $D(U) = \text{Coker}(Y^* \rightarrow X^*)$ is a homomorphic image of X^* , which is a principal indecomposable right A - module. So $D(D(M))$ is a direct sum of homomorphic images of principal indecomposable right A - modules, and it follows by the Krull - Schmidt Theorem. So, A is a real ring.

Definition

Artinian Principal ideal rings. We know that a module of finite length is homogeneous, if its composition factor is all isomorphic to one another.

Theorem 1.3

Let A be a ring with radical N . Then the following are equivalent.

- (1) A is an artinian principal ideal ring.
- (2) A is serial and every ideal of A is a principal left ideal.
- (3) Every left A -module is a direct sum of homogeneous uniserial modules.
- (4) A is serial and, as a left A -module, is the direct sum of homogeneous uniserial modules.

Proof: (1) $\Rightarrow$ (2): If A is an artian principal ideal ring then we have to show that if X is an indecomposable summand of A , then $\text{soc}(X / N^k X)$ is simple for every k . The preimage in A of $\text{soc}(X / N^k)$ is principal, so (X / N^k) is cyclic as an A / N -module. Hence the proof follows.

(2) $\Rightarrow$ (3): Let $A \cong \prod X_i$ as left A -modules, where each X_i is indecomposable. To prove that every indecomposable left A -module is homogeneous, it now suffices to prove that the modules X_i are all homogeneous. As $\text{soc}(A) = \prod \text{soc } X_i$ is

principal. Since it has the same length as A/N , we must have $A/N \cong \text{soc } A$. Let us take a simple A -module S and set $J = \{j \mid \text{soc } X_j \cong S\}$, $T = \prod_{j \in J} \text{soc } X_j$. T is an ideal of A , and A/T is also an artinian principal ideal ring, so again is that $(A/T) / \text{Rad } (A/T) \cong \text{soc } (A/T)$. Now counting the number of simple summands on each side of this equation which are isomorphic to S , we find that for each $j \in J$ such that $\lambda(X_j) \geq 2$, $\text{soc } (X_j / \text{soc } X_j) \cong S \cong \text{soc } X_j$. On continuing in this manner we find that each X_j is homogeneous. Since S is arbitrary follows that every X_i is homogeneous. Hence proved.

(3) $\Rightarrow$ (4): It is quite obvious from the above statement.

(4) $\Rightarrow$ (1): If X and Y are homogeneous uniserial projective modules, then $\text{Hom}_A(X, Y) \neq 0$ if and only if $X \cong Y$. Of course, we may assume that A is indecomposable, and we find that A is a direct sum of isomorphic indecomposable left modules; $A = \prod_{i=1}^n X_i$. Hence A is isomorphic to the ring of $n \times n$ matrices over the local serial ring $\text{End}_A(X_i)$ so the category of A -modules is equivalent to the category of $\text{End}_A(X_i)$ -modules, so every indecomposable right or left A -module is homogeneous.

Hence by the use of symmetry, it is enough to show that any left ideal L is principal. This is equivalent to showing that L/NL is a cyclic A/N -module. But L/NL is a direct summand of $\text{soc } (A/NL)$, so it is enough to show that $\text{soc } (A/NL)$ is a cyclic A/N -module. But A/NL is a direct sum of homogeneous uniserial modules. Hence, $\text{soc}(A/NL) \cong A/N \otimes_A A/NL$ which is cyclic as an A/N -module, since A/NL is cyclic as an A -module.

As indecomposable serial ring, each of whose indecomposable modules is homogeneous, is a full matrix ring over a local serial ring. Also, a local serial ring is trivially being a principal ideal ring. Hence the result has been proved.

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On Algebraic Structures

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Abstract

The well know theorem of Algebra known as Lagrange's theorem is that the order of a subgroup of a finite group is a divisor of the order of that group but the converse is not true our aim here is to discuss about the converse part by imposing distinct algebraic structures. As also it was discussed by Cauchy & Sylow in details.

Keywords: Algebra, order, subgroup, algebraic structure

Introduction

Lagrange's theorem says that the order of a finite group is divided by the order of its subgroup but the converse of the theorem is not true. There are very very few theorems which assure the existence of subgroup of that order in a certain finite groups. If G be a finite abelian group and a prime $p \mid O(G)$, then there is an element a other than the identity $e \in G$ such that $o(a) = p$. If G is an abelian group & every subgroup of an abelian group is normal so H is also normal in G hence the quotient group G/H of G exist & G/H is abelian because G is also abelian. Since $O(G/H) = \frac{O(G)}{O(H)} < o(G)$

$$\therefore O(H) > 1$$

Also if G be a finite group & a prime $p \mid O(G)$ then G is an element other than e of order p . If H is a proper subgroup of G such that $O(H) < O(G) = \sum_a \frac{O(G)}{O(N/a)}$, where the

summation runs over one element a in each conjugate class & $N(a)$ is the normalizer of a . If z is the centre of G then $a \in z \Leftrightarrow N(a) = G \Leftrightarrow O(N(a)) = O(G) \Rightarrow \frac{O(G)}{O[N(a)]} = 1$.

Theorem 1.2

Let G be a finite abelian group, and let p be a prime. If $p \mid o(G)$, then there is an element $a \neq e \in G$ such that $a^p = e$.

Proof: The theorem may be proved by induction method. Let the theorem be true for abelian groups of order less than $o(G)$. From this we shall prove that the result is also true for G . The theorem is true for group G with $o(G) = 1$; i.e. $G = \{e\}$.

If G has only improper subgroups; i.e., G has no subgroups $H \neq (e)$, G , then G must be of prime order since every group of composite order possesses proper subgroups. Now p is a prime number and $p \mid o(G)$. So, $o(G) = p$. Further, we know that every group of prime order is cyclic. Hence, each element $a \neq e$ of G will be a generator of G . Hence excluding e , the remaining $p - 1$ elements $a (\neq e)$ of G are such $a^{o(G)} = e$ i.e., $a^p = e$.

So suppose that G has proper subgroups N , then $N \neq (e)$, G . Now there arise two possibilities: either $p \mid o(N)$ or $p \nmid o(N)$.

If $p \mid o(N)$, by our induction hypothesis, since $o(N) < o(G)$ and N is abelian, $\exists b \in N, b \neq e$ such $b^p = e$. Since $b \in N \subseteq G$, so we have shown that $\exists$ an element $b \in G, b \neq e$ such that $b^p = e$.

G/H So, we assume that $p \nmid o(N)$. Since G is abelian N is a normal subgroup of G , so G/N is a group, Also

$$o(G/N) = \frac{o(G)}{o(N)}$$

$$o(G) = o(G/N)o(N) \dots \dots \dots (1)$$

Since, $p \mid o(G)$ and $p \nmid o(N)$, it follows from (1) that

$p \mid o(G)$ and $p \nmid o(N)$, it follows from (1) that

$$p \mid o(G/N) = \frac{o(G)}{o(N)} < o(G)$$

$$[\therefore o(N) > 1]$$

also, since G is abelian G/N is abelian. So by our induction process. $\exists$ an element $Nb \in G/N$ satisfying $(Nb)^p = N$, the

identity element of G/N , $Nb \neq n$; in the quotient group G/N we have $o(Nb) = p$. We also have

$$(Nb)^p = N, Nb \neq N$$

$$\Rightarrow Nb^p = N, Nb \neq N$$

$$\Rightarrow b^p \in N, b \neq N$$

Now using Lagrange's theorem, we have

$$(b^p)^{o(N)} = e$$

$$\Rightarrow b^{o(N)p} = e$$

$$\Rightarrow (b^{o(N)})^p = e$$

Let $b^{o(N)} = c$. Then certainly $c^p = e$.

In order to show that c is the required element that satisfies the theorem we must finally show that $c \neq e$. However, if $c = e$ then we have

$$\begin{aligned} b^{o(N)} = c, c = e &\Rightarrow b^{o(N)} = e \\ &\Rightarrow Nb^{o(N)} = Ne \\ &\Rightarrow (Nb)^{o(N)} = N \end{aligned}$$

Also in the quotient G/N , $o(Nb) = p$; i.e. $(Nb)^p = N$, $p \mid o(N)$, p is a prime number so

$$Nb = N$$

$$\Rightarrow b \in N, \text{ which is a contradiction to the fact that } b \notin N.$$

So $c \neq e$, $c^p = e$. Hence the theorem holds true.

Theorem 1.3

If G is a finite group, p is a prime number and $p \mid o(G)$, then G has an element a of order p .

Proof: Let G be a group, p be a prime number and $p \mid o(G)$. An element $a \neq e \in G$ satisfying $a^p = e$. To prove its existence by method of induction on $o(G)$; that is the theorem is true for all groups of order less than $o(G)$ and finally show that it is true for G .

If $o(G) = 1$; i.e., $G = \{e\}$ then the theorem is clearly true.

If H is a proper subgroup G , now two cases arise

$$(i) p \mid o(H)$$

$$\text{and } (ii) p \nmid o(H)$$

If there exists a subgroups $H \neq G$ of G such that $p \mid o(H)$, then $o(H) < o(G)$, so by induction the theorem is true for H .

So, $\exists$ an element $a \in H$ such that $o(a) = p$. But

$$a \in H, H \subseteq G \Rightarrow a \in G.$$

Therefore, $\exists$ is element $a \neq e \in G$ such that $o(a) = p$.

Now we assume that $p \nmid o(H)$, where H is any proper subgroup of G . Suppose $Z(G)$ is the centre of G . We now write the class equation for G :

$$o(G) = o(Z(G)) + \sum_{a \notin Z(G)} \frac{o(G)}{o(N(a))}$$

i.e.

$$o(G) = o(Z(G)) + \sum_{N(a) \neq G} \frac{o(G)}{o(N(a))}$$

We know that $N(a)$ is a subgroup of G . If $a \notin Z(G)$, then clearly $N(a) \neq G$. So, $p \nmid o(N(a))$. If $N(a) \neq G$, then

$$c_a = \frac{o(G)}{o(N(a))}$$

$$\Rightarrow o(G) = c_a o(N(a)).$$

Since $p \mid o(G)$ and $p \nmid o(N(a))$, it follows

$$p \mid c_a; \text{ i.e., } p \mid \frac{o(G)}{o(N(a))}, \text{ if } N(a) \neq G \text{ and so}$$

$$p \mid \sum_{N(a) \neq G} \frac{o(G)}{o(N(a))}$$

Again

$$p \mid o(G) \text{ and } p \nmid \sum_{N(a) \neq G} \frac{o(G)}{o(N(a))}$$

$$\Rightarrow p \mid \left(o(G) - \sum_{N(a) \neq G} \frac{o(G)}{o(N(a))} \right)$$

$$\Rightarrow p \mid o(Z(G))$$

Hence, $Z(G)$, is a proper subgroup of G such that $p \mid o(Z(G))$. But, p is not a divisor of any proper subgroup of G , and so $Z(G)$ cannot be a proper subgroup of G that $Z(G) = G$. But then G is abelian. But p is a prime number and so $o(a) = p$. This proves the theorem.

Theorem 1.4

Let G be a finite group and let p be a prime. If $p^m \mid o(G)$, then G has a subgroup of order p^m .

Proof: We prove the theorem by induction method on $o(G)$. If $o(G) = 1$, the result is clear so we assume that the results holds for all groups of order less than $o(G)$. We want to show that the result is valid for G . Let $Z(G)$ denote the centre of G .

If $p \mid o(Z(G))$, then it follows from Cauchy's theorem for abelian group that $Z(G)$ contains an element $b \neq e$ of order p . Let B be the cyclic group generated by b ; i.e., $B = \langle b \rangle$. Since $b \in Z(G)$, B must be normal in G . Moreover, $o(B) = o(b) = p$. Hence we can form the quotient group $\overline{G} = G/B$ and

$$o(\overline{G}) = o(G/B) = \frac{o(G)}{o(B)} = \frac{o(G)}{p}$$

$$\Rightarrow p^{m-1} \mid o(\overline{G})$$

Also, $o(\overline{G}) < o(G)$. Since $p > 1$. By the induction process,

$\overline{G}$ contains a subgroup $\overline{P}$ of order p^{m-1} .

Now, we define a natural mapping $\phi: G \rightarrow G/B$ defined

by $\phi(x) = Bx \forall x \in G$ is a homomorphism of G onto

G/B with kernel B . Let $P = \{x \in G \mid \phi(x) \in \overline{P}\}$. Then

P is a subgroup of G and $\overline{P} \cong P/B$. Hence,

$$o(\overline{P}) = o(P/B) = \frac{o(P)}{o(B)}$$

$$\Rightarrow o(P) = o(\overline{P})o(B) = p^{m-1} \cdot p = p^m$$

Thus, G has a subgroup P of order p^m .

Now consider the case when $p \nmid o(Z(G))$.

The class equation of G is defined as follows

$$o(G) = o(Z(G)) + \sum_{a \notin Z(G)} \frac{o(G)}{o(N(a))}$$

Now, $p \mid o(G)$, but $p \nmid o(Z(G))$. Hence $a \notin Z(G)$ in view

$$\text{of (I)} \quad p \nmid \frac{o(G)}{o(N(a))}$$

For some $a \in G, a \notin Z(G)$. It follows that $p^m \nmid \frac{o(G)}{o(N(a))}$.

But $c_a = \frac{o(G)}{o(N(a))}$. Hence $p^m \nmid c_a$. Further, $c_a = \frac{o(G)}{o(N(a))}$

$$\Rightarrow o(G) = c_a o(N(a))$$

Since $p^m \mid o(G)$ and $p^m \nmid c_a$ it follows that $p^m \mid o(N(a))$ and $o(N(a)) < o(G)$.

$N(a)$ must have a subgroup, say T of order p^m . Again, since T is a subgroup of $N(a)$ and $N(a)$ is a subgroup of G , it follows that G has a subgroup T of order p^m . This proves the theorem.

Theorem 1.5

If A, B are finite subgroups of a group G , then

$$o(AB) = \frac{o(A)o(B)}{o(A \cap B)}$$

Proof: At first we define the suitable mapping

$$T: A \times B \rightarrow A \times Bx^{-1}$$

$$T(axb) = axbx^{-1} \quad \forall axb \in A \times B$$

It is clear that T is well defined as

$$axb = a'x b'axbx^{-1} = a'$$

For T is one-one. Let $a_1x b_1, a_2x b_2 \in A \times B$.

Then

$$\begin{aligned} T(a_1x b_1) = T(a_2x b_2) &\Rightarrow a_1x b_1x^{-1} = a_2x b_2x^{-1} \\ &\Rightarrow a_1x b_1x = a_2x b_2x \quad [\text{by} \end{aligned}$$

right cancellation law in G]

Hence T is one - one.

For T is onto. Let $axbx^{-1}$, be an arbitrary element of $A \times Bx^{-1}$, then $\exists$ an element $axb \in A \times B$ such that

$$T(axb) = axbx^{-1}$$

Hence T is onto.

Hence T is one-one mapping from $A \times B$ onto $A \times Bx^{-1}$. So

$$o(A \times B) = o(A \times Bx^{-1}).$$

Also, Bx^{-1} is a subgroup of G , of order $o(B)$. Hence

$$o(A \times B) = o(A \times Bx^{-1}) =$$

$$\frac{o(A)o(Bx^{-1})}{o(A \cap Bx^{-1})} = \frac{o(A)o(B)}{o(A \cap B)}$$

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The Role of Isomorphism in Fuzzy groups and subgroups

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Abstract

There are various connections between the notion of fuzzy isomorphisms and also the concept of equivalence of fuzzy groups and subgroups. As the term equivalence is much more general than isomorphism of fuzzy group or subgroups. Whenever two fuzzy subgroups are equivalent then they form fuzzy isomorphic but converse need not be true. Our aim here is to discuss about the generalization of isomorphism & equivalence.

Keywords: Isomorphism, equivalence, fuzzy isomorphism, group, subgroup.

Introduction

Since isomorphism is a concept of algebraic structure where as equivalence is an order structure in fuzzy groups or subgroups. Hence, isomorphism & equivalence are not comparable in general but in particular. We may compare both these concepts of course two fuzzy subgroups are equivalent if they maintain the same relative degrees of membership between any pair of elements.

As the matter of equality of fuzzy subgroup is very strong to be of any use in the classification complexity, the complexity generally arises from uncertainty in the form of ambiguity. The probability theory has been an age old and very effective tool to handle uncertainty whereas it can be applied only to situations where characteristic are based on random processes. Uncertainty may arise due to partial information which is not fuzzy reliable. Fuzzy set is an excellent mathematical tool to handle the uncertainty.

As the classical set theory was propounded by German mathematician George cantor, By taking $I = [0, 1]$ is called the unit interval, $\wedge$ & $\vee$ stand for uniform (intersection) & supremum (union). G is taken to be a group. They we may define A fuzzy subset of a non - empty set G is a mapping $G : G \rightarrow I$ G^I denotes the set of all fuzzy subset of G .

Theorem

If f is an isomorphism of G onto H since that $f(\mu) = \nu$ then f is called an isomorphism of μ onto ν .

Since $f(\mu) = \nu$ is a very strong condition because slight change in the membership values of either μ or ν then the definition of isomorphism fails at once. This is so because it is easily deduced from the given definition that the set of membership values of both μ and ν should be identical. Let us verify by example –

If $S = \{e, a, a^2, b, ab, a^2b\}$

$$\& 0 < \epsilon_1 < \epsilon_2 < \frac{1}{2} \& \text{ we}$$

$$\mu(x) = \begin{cases} 1 & \text{if } x = e \\ \frac{1}{2} & \text{if } x = a, a^2 \& \\ \epsilon_1 & \text{otherwise} \end{cases}$$

$$\nu(x) = \begin{cases} 1 & \text{if } x = e \\ \frac{1}{2} & \text{if } x = a, a^2 \& \\ \epsilon_2 & \text{otherwise} \end{cases}$$

If $\epsilon_1 \neq \epsilon_2$ then both fuzzy subgroups will be non-isomorphic. Hence we are going to define in the modified form as follows:

If f is an isomorphism of $\text{supp}(\mu)$ onto $\text{supp}(\nu)$, then f is called a fuzzy isomorphism of μ onto ν if there exists a positive real constant $0 < k < 1$ such that

$$\mu(x) = k\nu(f(x)) \forall x \in \text{Supp}(\mu) \text{ such}$$

that $x \neq e$.

This definition is a slight generalization of the above definition but still has its limitation in the sense that the membership values of μ are either magnified or contracted proportionally to the membership values of ν by one constant of proportionality k . Any reasonable definition of isomorphism between two fuzzy groups must satisfy at least the following two conditions (i) if G and K are isomorphic groups with ϕ an isomorphism of G onto K , then the mapping $\mu \rightarrow \phi(\mu)$ must be one - to - one corresponding fuzzy group μ and $\phi(\mu)$ should be pairwise isomorphic for all μ according to the definition

defined and (ii) it is necessary that under this mapping the set of all normal fuzzy subgroups of G should correspond to the set of all normal fuzzy subgroups of K .

Further we may define as follows: If μ and ν are fuzzy isomorphic if there is a one-to-one correspondence g from $\text{Im}(\mu)$ onto $\text{Im}(\nu)$ such that (i) g is order preserving and (ii) the level sets μ^α and $\nu^{g(\alpha)}$ are isomorphic for all α in the image of μ , (iii) μ and ν have isomorphic supports under ϕ that the fuzzy subgroups are fuzzy isomorphic. This definition is better than the proposed. Even though the definition of fuzzy isomorphism is sharper than Mordeson's definition and Maamba's definition, it lacks the properties that would normally be preserved by isomorphism. The definition of fuzzy isomorphism does not preserve fuzzy normality. Hence these modified definition of isomorphism by adding an extra condition to the above definition.

Theorem 1.3: Contribution of Keychains and pinned - flags

We discuss the concept of keychains to we assume that G is finite group. An $(n+1)$ -tuple $1\lambda_1\lambda_2\dots\lambda_n$ of real numbers in I of the form $1\geq\lambda_1\geq\lambda_2\geq\dots\geq\lambda_{n-1}\geq\lambda_n\geq 0$, is called a keychain of length $n+1$. The λ_i 's are called pins. The number of times a pin is repeated is called the padidity and the set of all padidities of a keychain is called the index of the keychains. So there is a correspondence between integer partition of n and the indices of keychains.

Similarly an increasing maximal chain of $n+1$ subgroups of G starting with trivial subgroup $\{0\}$ is called a flag on G . A pinned-flag is a pair (C, l) where C is a flag on G and l is a keychain, written for fuzzy subgroups as follows:

$$0^1 \subset G_1^{\lambda_1} \subset G_2^{\lambda_2} \dots \subset G_n^{\lambda_n}$$

The correspondence between pinned-flags and equivalence classes of fuzzy subgroups is clear if we associated a fuzzy subgroup μ correspondence to (C, l) on G as follows:

$$\mu(x) = \begin{cases} 1, & x = 0 \\ \lambda_1 & x \in G_1 \setminus \{0\} \\ \lambda_2 & x \in G_2 \setminus G_1 \\ \vdots & \\ \lambda_n & x \in G_n \setminus G_{n-1} \end{cases}$$

Theorem 1.4:

Suppose the pinned - flags correspondence of two fuzzy subgroups μ and ν are given by

$$(C_\mu, l_\mu) : G_0^1 \subset G_1^{\lambda_1} \subset \dots \subset G_n^{\lambda_n}$$

$$(C_\nu, l_\nu) : H_0^1 \subset H_1^{\beta_1} \subset \dots \subset H_n^{\beta_n}$$

Then $\mu \sim \nu$ on G if and if:

- (i) $n = m$;
- (ii) $G_i = H_i$ for $i = 0, 1, \dots, n$; provided the λ_i 's and β_j 's are distinct.
- (iii) $\lambda_i > \lambda_j$ if and only if $\beta_i > \beta_j$ for $1 \leq i, j \leq n$ and the $\lambda_k = 0$ if and only if $\beta_k = 0$ for $1 \leq k \leq n$.

The proof Proposition can be modified to yield a proof for fuzzy subgroups as stated above. It is clear from above that with each equivalence class determined by a fuzzy subgroup μ of G we can associate one and only one pinned flag (C, l)

which is denoted by (C_μ, l_μ) , or simply by C_μ if the keychain l_μ can be determined simply by l_μ if C_μ is known. Conversely every pinned-flag gives rise to only one equivalence class of fuzzy subgroups.

Theorem 1.5 A comparison between equivalence and isomorphism of fuzzy subgroups

Here we compare equivalence with the fuzzy isomorphism equivalence involves fuzzy subgroups of the same fixed group G , while isomorphism may involve fuzzy subgroups of two distinct groups. So, in order to compare the two distinct groups. So, in order to compare the two concepts we have to fix a group G and at fuzzy subgroups of G . Otherwise the concept of equivalence and fuzzy isomorphism cannot be compared.

Now let μ and ν be fuzzy subgroups of G , if μ is equivalent to ν , then it is clear than two fuzzy subgroup are fuzzy isomorphic in terms of the definitions where the isomorphism f is the identity mapping. Conversely it is possible to have two fuzzy subgroups μ and ν to be fuzzy isomorphic without μ being equivalent to ν .

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History of the Ancient Vedic Mathematics and its Application

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Abstract

The ‘Vedas’ are considered ‘divine’ in origin and are assumed to be direct revelations from God. Much of the hype about this topic is based on one single book authored by the Sankaracharya, Jagadguru Swami Sri Bharati Krishna Tirthaji Maharaja titled Vedic Mathematics and published in the year 1965 and reprinted several times since the 1990s. In this the authors probe into Vedic Mathematics and explore whether it is really ‘Vedic’ in origin or ‘Mathematics’ in content. The entire field of Vedic Mathematics is supposedly based on 16 one - to three - word sutras (aphorisms) in Sanskrit, which can solve all modern mathematical problems. In this paper we have gone through different literature on ancient Vedic mathematics and its application. We have discussed the Sutra: Nikhilam Navataścaramam Daśatah and its application.

Keywords: Vedas, Vedic mathematics, sutras and Nikhilam Navataścaramam Daśatah

Introduction

India today has active and excellent schools of research and teaching in mathematics that are at the forefront of modern research in their discipline with some of them recognised as being among the best in the world in their fields of research. It is noteworthy that they have cherished the legacy of distinguished Indian mathematicians like Srinivasa Ramanujam, V. K. Patodi, S. Minakshisundaram, Harish Chandra, K. G. Ramanathan, Hansraj Gupta, Syamdas Mukhopadhyay, Ganesh Prasad, and many others including several living Indian mathematicians. But not one of these schools has lent an iota of legitimacy to ‘Vedic Mathematics’. Nowhere in the world does any school system teach “Vedic Mathematics” or any form of ancient mathematics for that matter as an adjunct to modern mathematical teaching. The bulk of such teaching belongs properly to the teaching of history and in particular the teaching of the history of the sciences. Mantras from the early Vedic period (before 1000 BCE) invoke powers of ten from a hundred all the way up to a trillion, and provide evidence of the use of arithmetic operations such as addition, subtraction, multiplication, fractions, squares, cubes and roots. A 4th Century CE Sanskrit text reports Buddha enumerating numbers up to 10, as well as describing six more numbering systems over and above these, leading to a number equivalent to 10. Given that there are an estimated 10 atoms in the whole universe, this is as close to infinity as any in the ancient world came. It also describes a series of iterations in decreasing size, in order to demonstrate the size of an atom, which comes remarkably close to the actual size of a carbon atom (about 70 trillionths of a metre). Vedic

mathematics is the name given to the ancient system of mathematics, or, to be precise, a unique technique of calculations based on simple rules and principles with which any mathematical problem can be solved – be it arithmetic, algebra, geometry or trigonometry. Vedic mathematics was rediscovered from the ancient Indian scriptures between 1911 and 1918 by Sri Bharati Krishna Tirthaji, a scholar of Sanskrit, mathematics, history and philosophy. He studied these ancient texts for years and after careful investigation, was able to reconstruct a series of mathematical formulae called *sutras*. Bharati Krishna Tirthaji, who was also the former Shankaracharya of Puri, delved into the ancient Vedic texts and established the techniques of this system in his pioneering work, Vedic Mathematics (1965), which is considered the starting point for all work on Vedic mathematics.

Vedic mathematics was immediately hailed as a new alternative system of mathematics when a copy of the book reached London in the late 1960s. Some British mathematicians, including Kenneth Williams, Andrew Nicholas and Jeremy Pickles, took interest in this new system. They extended the introductory material of Bharati Krishna’s book, and delivered lectures on it in London. In 1981, this was collated into a book entitled Introductory Lectures on Vedic Mathematics. A few successive trips to India by Andrew Nicholas between 1981 and 1987 renewed interest in Vedic mathematics and scholars and teachers in India started taking it seriously. A great deal of research is also being carried out on how to develop more powerful and easy applications of the Vedic *sutras* in geometry, calculus and computing. . The system is based on 16 Vedic *sutras* or

aphorisms, which are actually word formulae describing natural ways of solving a whole range of mathematical problems. 16 *Vedic sutras and their corollaries* are shown in table 1.

Sl. No.	Sutras	Sub Sutras or Corollaries
1	Ekādhikena Pūrvena, also a corollary	Ānurūpyena
2	Nikhilam Navataścaramam Daśatah	Śisyate Śesamjnah
3	Ūrdhva - tiryagbhyām	Ādyamādyenantyamant yena
4	Parāvartya Yojayet	Kevalaih Saptakam Gunṣat
5	Sūnyam Samyasamuccaye	Vestanam
6	(Ānurūpye) Sūnyamanyat	Yāvadūnam Tāvadūnam
7	Sankalana - vyavakalanābhyām	Yāvadūnam Tāvadūnīkrtya Vargaṇca Yojayet
8	Puranāpuranābhyām	Antyayordasake' pi
9	Calanā kalanābhyām	Antyayoreva
10	Yāvadūnam	Samuccayagunitah
11	Vyastisamastih	Lopanasthāpanabhyām
12	Śesānyankena Caramena	Vilokanam
13	Sopantyadvayamantyam	Gunitasamuccayah Samuccayagunitah
14	Ekanyūnena Pūrvena	
15	Gunitasamuccaya	
16	Gunakasamuccayah	

16 slokas has been compiled from stray references. Each slokas gives one or more mathematical theories formulated by Bharati Krishna Tirthaji. A major part of the body of mathematical knowledge from the Vedic period that has come down to us is from the Sulvasutras. The Sulvasutras are compositions aimed at providing instruction on the principles involved and procedures of construction of the vedis (altars) and agnis (fireplaces) for performing the yajnas, which were a key feature of the Vedic culture. The fireplaces were constructed in a variety of shapes such as falcons, tortoise, chariot wheels, circular trough with a handle, pyre, etc (depending on the context and purpose of the particular yajna) with sizes of the order of 20 to 25 feet in length and width, and there is a component of the Sulvasutras describing the setting up of such platforms with tiles of moderate sizes, of simple shapes like squares, triangles, and occasionally

special ones like pentagons. Many of the vedis involved, especially for the yajnas for special occasions had dimensions of the order of 50 to 100 feet, and making the overall plan involved being able to draw perpendiculars in that setting. This was accomplished both through the method that is now taught in schools, involving perpendicularity of the line joining the centres of two intersecting circles with the line joining the two points of intersection, as also via the use of the converse of "Pythagoras theorem"; they were familiar with the Pythagoras theorem", and explicit statement of the theorem is found in all the four major Sulvasutras. The Sulvasutras also contain descriptions of various geometric principles and constructions, including procedures for converting a square into a circle with equal area and vice versa and a good approximation to the square root of 2.

The Sulvasutras, like other Vedic knowledge, were transmitted only orally over a long period. There have also been commentaries on some of the Sulvasutras in Sanskrit, but their period remains uncertain. When the first written versions of the Sulvasutras came up is unclear. The text versions with modern commentaries were brought out by European scholars (Thibaut, Burk, van Gelder and others) starting from the second half of the nineteenth century. With regard to genesis of his study of the Sulvasutras Thibaut mentions that the first to direct attention to the importance of the Sulvasutras was Mr. A.C. Burnell, who in his Catalogue of a Collection of Sanskrit Manuscripts, remarks that we must look to the Sulva portions of the Kalpasutras for the earliest beginnings among the Brahman. While the current translations are reasonably complete, some parts have eluded the translators, especially in the case of Manava Sulvasutra which turns out to be terser than the others. New results have been brought to light by R.G. Gupta, Takao Hayashi and perhaps also by others, not recognised by the original translators. Lack of adequate mathematical background on the part of the translators could be one of the factors in this respect. There is a case for a relook on a substantial scale to put the mathematical knowledge in the Sulvasutras on a comprehensive footing. There is also scope for work in the nature of interrelating in a cohesive manner the results described in the various Sulvasutras. The ritual context of the Sulvasutras lends itself also to the issue of interrelating the ritual and mathematical aspects, and correlating with other similar situations from other cultures; for a perspective on this the reader may refer Seidenberg.

Another natural question that suggests itself in the context of the Sulvasutras is whether there are any of the fireplaces from the old times to be found. From the description of the brick construction it would seem that they would have been

too fragile to withstand the elements for long; it should be borne in mind that the purpose involved did not warrant a long-lasting construction. Nevertheless, excavations at an archaeological site at Singhol in Panjab have revealed one large brick platform in the traditional shape of a bird with outstretched wings, dated to be from the second century BCE; it however differs markedly from the numerical specifications described in the Sulvasutras. This leaves open the possibility of finding other sites, though presumably not a very promising one.

Apart from the Sulvasutras, mathematical studies have also been carried out in respect of the Vedas, mainly concerning understanding of the numbers. For a composition with a broad scope, including spiritual and secular, the Rig-Veda shows considerable preoccupation with numbers, with numbers up to 10,000 occurring, and the decimal representation of numbers is seen to be rooted there. The Yajurveda introduces names for powers of 10 upto 10^{12} and various simple properties of numbers are seen to be involved in various contexts for instance. There is scope for further work in understanding the development as a whole; this would involve familiarity with mathematics on the one hand and knowledge of Vedic Sanskrit on the other hand.

The Sutra: Nikhilam Navataścaramam Daśatah

The sutra reads “*Nikhilam Navataścaramam Daśatah*”, which literally translated means: all from 9 and the last from 10”. We shall presently give the detailed explanation presently of the meaning and applications of this cryptical - sounding formula and then give details about the three corollaries.

He has given a very simple multiplication.

Suppose we have to multiply 9 by 7

1. We should take, as base for our calculations that power of 10 which is nearest to the numbers to be multiplied. In this case 10 itself is that power.
(10)
 $9 - 1$
 $7 - 3$
 $6 / 3$
2. Put the numbers 9 and 7 above and below on the left hand side (as shown in the working alongside here on the right hand side margin)
3. Subtract each of them from the base (10) and write down the remainders (1 and 3) on the right hand side with a connecting minus sign (–) between them, to show that the numbers to be multiplied are both of them less than 10.

4. The product will have two parts, one on the left side and one on the right. A vertical dividing line may be drawn for the purpose of demarcation of the two parts.
5. Now, the left hand side digit can be arrived at in one of the 4 ways,
 - a. Subtract the base 10 from the sum of the given numbers (9 and 7 i.e. 16). And put $(16 - 10)$ i.e. 6 as the left hand part of the answer or $9 + 7 - 10 = 6$
 - b. Subtract the sum of two deficiencies $(1 + 3 = 4)$ from the base (10) we get the same answer (6) again or $10 - 1 - 3 = 6$
 - c. Cross subtract deficiency 3 on the second row from the original number 9 in the first row. And you find that you have got $(9 - 3)$ i.e. 6 again $9 - 3 = 6$
 - d. Cross subtract in the converse way (i.e. 1 from 7), and you get 6 again as the left hand side portion of the required answer $7 - 1 = 6$

This availability of the same result in several easy ways is a very common feature of the Vedic system and is great advantage and help to the student as it enables him to test and verify the correctness of his answer step by step.
6. Now vertically multiply the two deficit figures (1 and 3). The product is 3. And this is the right hand side portion of the answer (10)
 $9 - 1$
 $7 - 3$
 $6 / 3$
7. Thus $9 \times 7 = 63$ This method holds well in all cases and is therefore capable of infinite application.

Conclusion

Vedic mathematical methods are derived from ancient systems of computations, now made available to everyone through the great work of Jagadguru Swami Sri Bharati Krisna Tirthaji Maharaja, who published a book on Vedic mathematics in 1965. Compared to conventional mathematical methods, these are computationally faster and easy to perform. In this paper we have shown the second sutras of Bharati Krishna Tirthaji which is applicable for multiplication only. Vedic Mathematics mainly deals with various Vedic mathematical formulas and their applications of carrying out tedious and cumbersome arithmetical operations, and to a very large extent executing them mentally.

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The Study of Orthogonal Group of Space

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Abstract

The aim of this paper is to derive the group of isometries in Clifford algebra with the set of all forms on a field of the Witt group. Further attempts have been made to construct isotropic & anisotropic vector in an inner product space.

Keywords: Group, isometries, anisotropic, vector, inner product, space.

Introduction

At first we are going to define a space which is known as inner product space is that an inner product space over a field K is a pair (V, b) consisting of a finite dimensional vector space V and a symmetric bilinear form b and V . This means there is a mapping $b: V \times V \rightarrow K$ which is symmetric $b(x, y) = b(y, x)$ and is bilinear.

$$b(\alpha x + \beta y, z) = \alpha b(x, z) + \beta b(y, z)$$

$$\text{and } b(x, \alpha y + \beta z) = \alpha b(x, y) + \beta b(x, z)$$

Inner product space is also defined in terms of a quadratic forms by a suitable mapping $q: V \rightarrow K$ s.t

$$q(\alpha x) = \alpha^2 q(x) \quad \forall x \in V \text{ and } \alpha \in K \text{ and}$$

$$q(x+y) - q(x) - q(y) = bq(x, y)$$

is a bilinear form on V . The pair (V, q) is called a quadratic space. It is quite clear that the function bq is symmetric which is known as associated bilinear form which is very useful to define the inner product structure on V . If characteristic $\neq 2$, then every symmetric bilinear form is and then $b = bq$ where $q = \frac{1}{2} b(x, x)$.

$$\text{Hence } b = (x+y, x+y) - b(x, x) - b(y, y) = b(x, y) + b(y, x).$$

When b is symmetric then $b(x, y) = b(y, x)$ then

$$bq(x, y) = q(x+y) - q(x) - q(y) = b(x, y)$$

This shows that there is complete equivalence between quadratic forms and bilinear forms.

If V be an inner product space and relative to the basis $e_1, e_2, e_3, \dots, e_n$ of the space V and b is obtained by $a_{ij} = b(e_i, e_j)$ and if $x = \sum \lambda_i e_i$ and $y = \sum \mu_i e_i$ then by linearity property $b(x, y) = \sum a_{ij} \lambda_i \mu_j$

1.2 An inner product space is singular iff the matrix of its form is singular.

Proof:

Let us take V be a regular space and U be any subspace. If $V_1, V_2, \dots, V_n$ is a basis of V and also such that $V_1, V_2, \dots, V_n$ is a basis of V and also such that $V_1, V_2, \dots, V_n$ is a basis of U then a vector $x = \sum c_i V_i$ is orthogonal to U iff $b(x, V_j) = \sum c_i (V_i, V_j) = 0 \dots (1)$

where $J = 1, 2, 3, \dots, r$

As $n \times n$ matrix $b(V_i, V_j)$ is regular, so $n \times r$ matrix consists of first r columns has rank r . Hence I has rank r and the space of solutions has dimension in r . V then $\dim U + \dim U^\perp = \dim V$ and also $U^{\perp\perp} = U$.

Proof:

It is very much obvious that $\dim U + \dim U^\perp = \dim U^\perp + \dim U^{\perp\perp}$

So $\dim U^{\perp\perp} = \dim U$. Also since $U \subseteq U^{\perp\perp}$ and hence $U^{\perp\perp} = U$.

It is for regular space, the determinant of the matrix although it is not invariant but an invariant of the space then the relation is

$$\det A^1 = \det A (\det P)^2$$

Where $(\det p)^2$ may take the value of any non zero square in K . It is denoted by K^{X^2} and hence each regular inner product space V gives a unique element of the factor group K^X/K^{X^2} which is known as the determinant of the space V .

1.3 If V be an inner product space then the isometries of V with itself are known as orthogonal transformation and they form a group which is known as orthogonal group denoted by $O(V)$. The elements of $O(V)$ are the vector space automorphisms θ of V so that $q(x\theta) = q(x)$, $\forall x \in V$.

If q has the matrix A then $O(V)$ possesses all invertible matrices X s.t.

$$X A X^T = A$$

If U is regular then $(\det X)^2 = 1 \Rightarrow \det X = \pm 1$.

Also the special orthogonal group $SO(V)$ is defined as the subgroup of $O(V)$ is defined as the subgroup of $O(V)$ containing all isometries of $\det 1$. Its elements are the proper orthogonal transformations known as rotations. It is clear that isometric spaces have isomorphic orthogonal groups. If $\beta : V \rightarrow V_1$ is an isometry then the mapping $\theta \rightarrow \beta^{-1} \theta \beta$ is an isomorphism from $O(V)$ into $O(V_1)$.

A non-zero vector $x \in V$ is called isotropic if $q(x) = 0$ otherwise x is anisotropic. A subspace U of V is isotropic if it contains an isotropic vector. If (V, q) be a quadratic space then from above $\beta(x, y) = q(x + y) - q(x) - q(y)$ is a symmetric bilinear form. Now, we may define the bilinear form $b(x, y) = \frac{1}{2} \beta(x, y) \Rightarrow b(x, x) = q(x)$.

Moreover if (V, q) be a quadratic space over K then Clifford algebra for (V, q) is a K -algebra C with a linear mapping $\mu : V \rightarrow C$ is admissible s.t. for any algebra A with a map

$\lambda : V \rightarrow A, \exists$ a unique homomorphism $\lambda^1 : C \rightarrow A$ s.t., $\lambda = \mu \lambda^1$.

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The Category of Projective Modules over Polynomial Rings

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Abstract

Projective modules form a generalization of free modules which is important in homological algebra. There has been a functor between categories of modules. As for any sequence, each functor takes zero maps to 380 maps. The aim of the paper is to confine over the different – categories applicable on different types of modules.

Keyword - Projective Module, Functor, homological algebra categories of modules and sequence.

Introduction

Since linear algebra deals with fields and vector spaces. We are concerned with the generalization to rings & modules over them. Whereas a vector space over a given field is obtained upto isomorphism by its division so there is a large variety for modules. The homological treatment of module theory requires the motions of projective & injective modules.

The theory of vector spaces over a field derives its selective simplicity from the fact than every vector space has a basis. This holds even for vector spaces over skew fields, whether it is finite dimensional or not. But it fails for general rings. This leads to the notion of free module. As linearly independent generating set is called a basis and a module is called free if it has a basis which means that any basis of a free module is a minimal generating set which means that a generating set such that has no proper sub set generate the whole module. As for any rings R, and R-modules M said to be simple if $M \neq 0$ & M has no sub modules other than 0 & M. Since over a field, the only simple module is a one dimensional vector space. A module which can be expressed to a direct sum of simple module is known as semi simple.

Theorem 1.2: Let R be any ring. Then for any set I there is a left R-module F_I with a mapping $\varphi: I \rightarrow F_I$ which is universal for mappings into left R-modules, i.e. given any mapping $f: I \rightarrow M$ into a left R-module, there is a unique homomorphism $f': F_I \rightarrow M$ such that $f = \varphi f'$. The module F_I is in fact free on the image of I under φ .

Proof: Let us first define F_I as the free left R-module on a basis $(\mu_\lambda)(\lambda \in I)^\varphi$. By given any mapping

$f: I \rightarrow M$, if there is a homomorphism $f': F_I \rightarrow M$ of the required form, then clearly $u_\lambda f' = \lambda f$; hence we must have

$$\left(\sum \alpha_\lambda u_\lambda\right) f' = \sum \alpha_\lambda (\lambda f) \quad \dots\dots\dots 1$$

This shows that at most one such homomorphism can exist. Now it is easily tested that Equation I indeed defines a homomorphism f' with the desired properties.

Let us take an exact sequence

$$0 \rightarrow M' \xrightarrow{\alpha} M \xrightarrow{\beta} M'' \rightarrow 0, \dots\dots\dots II$$

in which M'' is free is split exact.

Proof: We have to show that II split it is enough to find a left inverse of β . Let I be a basis of M'' ; since β is surjective, we can for each $\lambda \in I$ find $x_\lambda \in M$ mapped to λ by β .

The mapping $\lambda \mapsto x_\lambda$ extends to a homomorphism $\beta': M'' \rightarrow M$. Hence the sequence II splits.

Theorem 1.3: Let R be any ring and M be an R-module. Then there is a free R- module F with a submodule G such that $M \cong F/G$.

Proof : Let M be a left R-module say, with generating family $(a_\lambda), (\lambda \in \gamma)$. By taking a free left R-module F with basis $u_\lambda (\lambda \in I)$; the mapping $\lambda \mapsto a_\lambda$ extends to a homomorphism $f: \sum \alpha_\lambda u_\lambda \mapsto \sum \alpha_\lambda a_\lambda$ which is surjective, because the a_λ generate M hence $G = \ker f$ is a submodule of F such that $M \cong F/G$.

It is clear that a module M is finitely generated iff it can be written as F/G, where it is free on a finite set. If $M \cong F/G$, where F is free and F, G are both finitely

generated we say that M is finitely presented; if we merely known that G is finitely generated, M is said to finitely related.

We now compare the cardinals of different bases of a free module. In the case of vector spaces this cardinal is uniquely determined and is free the dimension of the space. We find that for most of the rings encountered here the cardinal is again unique, but this does not hold for all. We therefore consider briefly under what circumstances exceptions can occur.

A free R -module F with a basis of cardinal γ is said to have rank γ , if all its bases have the same rank, F is said to have unique rank. In the first place we have that free module which is not finitely generated always has unique rank. This result actually holds in a somewhat more general form. For any subset X of all R -module M we denote by RX the submodule of M generated by X .

Theorem 1.4: Given any ring R , let M be a left R -module with a minimal generating set X . If X is finite, of cardinal α , then any generating set of M has cardinal least α . In particular, M is not finitely generated and any two minimal generating have the same cardinal.

Proof: Let Y be any generating set of M , of cardinal β . Every $\gamma \in Y$ is affine combination of a finite number of elements from X , hence there is a finite sub X_γ of X such that $\gamma \in RX_\gamma$. We have that

$$X = \bigcup_{\gamma \in Y} X_\gamma \quad \dots\dots\dots \text{III}$$

It is clear that $\bigcup X_\gamma \subseteq X$, and $R(\bigcup X_\gamma)$ is a submodule containing Y and hence equal to M . So, $\bigcup X_\gamma$ is a generating set, equal to X , by the minimality of the letter. If Y were finite III we would express X as a finite union of finite sets. Which controls the fact that X is infinite. Hence Y must be infinite and from III and we find that

$$\alpha = |X| \leq \sum |X_\gamma| \leq \aleph_0 \beta = \beta$$

This shows that Y has cardinal at lest α . If Y is also minimal, then by interchanging the roles of X and Y we see that $\beta \leq \alpha$, hence $\beta = \alpha$.

Which splits, thus we have

$$R^m \cong H \oplus R^n$$

In terms of matrices, let B be the $m \times n$ matrix describing the mapping β in and let A be the $n \times m$ matrix describing the

splitting homomorphism, so that $AB = I_n$. We can now restate the conditions for IBN and weak finiteness:

Theorem 1. 5: (i) For any ring R , the following conditions are equivalent:

- (a) R has invariant basis number;
- (b) For any $m, n, R^m \cong R^n \Rightarrow m = n$
- (c) For any m, n , if $A \in R^{n \times m}, B \in R^{m \times n}$ and $AB = I, BA = I$, then $m = n$.
- (ii)
 - a) R is weakly finite
 - b) for any $n, R^n \cong H \oplus R^n \Rightarrow H = 0$;
 - c) for any n , if $A, B \in R_n$ and $AB = I$, then $BA = I$.

Moreover, any non-trivial weakly finite ring has IBN.

Proof : (i) Since (b) is practically a restatement of the definition. The isomorphism $R^m \cong R^n$ is described by matrices as in (c), so the condition for IBN is that $m = n$.

(ii) To say that an n -element generating set of R^n is linerly independent just as to the assertion that $H = 0$, and the condition $H = 0$ holds precisely when $\beta\beta' = 1$ i.e. $BA = I$.

Finally, if IBN fails and $R \neq 0$, then we have $R^m \cong R^n$ for some $m \neq n$, say $m > n$. then $R^m \cong R^{m-n} \oplus R^n$ and by (ii) (b) it follows that R is no weakly finite.

It is clear that the trivial ring is weakly finite, but does not have IBN. For example of rings with IBN that are not weakly finite. We say that a ring R with the property $ab = 1 \Rightarrow ba = 1$ is also called directly finite.

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Structure in Algebra with Cyclotomic Group

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Abstract

Our aim here is to present a comprehensive amount between the Galois structure of the group of units in a number field. Since the algebraic thoughts come from integral representation theory. Also it has been derived by means of Galois theory is than an equation is soluble by radicals of an insolubility of the aquatic.

Keywords: Galois structure, automorphism, binary operation, homomorphism, field theory.

Introduction

Galois Theory is an excellent composite of the theory of group with the theory of algebraic field extensions and has a wide application in the union of equation. This theory is the study of three mathematical objects on irreducible polynomials $f(x)$, the splitting field K and the set of automorphism of K over the field F . Since the set of all automorphism of F forms a group with respect to the composite composition.

Galois Theory gives a beautiful interplay of group and field theory. Also the fundamental theory of Galois theorem establish a one – to – one correspondence between the set of subfields of the splitting field E of a separable polynomial in $F(x)$ and the set of subgroups of the group of F -automorphism of E . The fascinating outcome of this correspondence is the development of a mechanism which transforms problem regarding sub fields of field into work amendable problems about sub groups of groups. He also deduced a new style term cyclotomic polynomial and has a great impact on the group theory.

Theorem 1.2.

Let E be a Galois extension of a field F of characteristic zero. Let K be any subfield of E containing F . Then the mapping $K \mapsto G(E/K)$ set up a one-to-one correspondence from the set of subfield of E containing F to the subgroups of $G(E/F)$ such that

- (i) $K = E_{G(E/K)}$
- (ii) For any subgroup H of $G(E/F)$, $H = G(E/E_H)$.
- (iii) $[E:K] = |G(E/K)|$, $[K:F] = \text{index of } G(E/K) \text{ in } G(E/F)$.
- (iv) K is normal extension of F if and only if $G(E/K)$ is a normal subgroup of $G(E/F)$.

- (v) If K is a normal extension of F , then $G(K/F) = G(E/F)/G(E/K)$.

Proof: (i) Since E is Galois extension of F , it is normal extension of F , so E is splitting field of some polynomial $f(x)$ over F . Again since K is subfield of E containing F , E is splitting field of $f(x)$ over K . By definition of normality, it follows that E is a normal extension of K . Because E is Galois extension of F , it is also a finite separable extension. K is the fixed field of $G(E/K)$, that is $K = E_{G(E/K)}$. This proves part (i)

(ii) E is a finite separable extension and we have given that H is a subgroup of $G(E/F)$. Thus, it follows that $H = G(E/E_H)$. This proves (ii).

(iii) Because E is a normal extension of F and also K , we have

$$[E:F] = |G(E/F)| \text{ and } [E:K] = |G(E/K)|$$

So the transitivity of finite extensions

$$[E:F] = [E:K] [K:F]$$

Gives $|G(E/F)| = |G(E/K)| [K:F]$

$$\Rightarrow [K:F] = \frac{|G(E/F)|}{|G(E/K)|}$$

Hence, $[K:F] = \text{index of } G(E/K) \text{ in } G(E/F)$.

This proves (iii)

(iv) Let $\bar{F}$ be an algebraic closure of F containing E . By definition of normality we know that K is a normal extension of F if and only if each embedding $\sigma: K \rightarrow \bar{F}$, which keeps each element of F fixed, maps K onto K .

We have that K is a normal extension of F if and only if for each $\sigma \in G(E/F)$, $\sigma(K) = K$ clearly, if K is a normal extension of F and $\sigma \in G(E/F)$ then σ restricted to K , that is $\sigma|_K: K \rightarrow E$ is an embedding of K into E , and hence, into $\bar{F}$ because $\bar{F}$ contains E .

So, we have $\sigma(K) = K$. Conversely, let $\sigma: K \rightarrow \bar{F}$ be an embedding that keeps each element of F fixed. So σ can be extended to $\sigma^*: E \rightarrow \bar{F}$. But since E is a normal extension of F , it follows that $\sigma^*(E) = E$. So $\sigma^*(E) = E$ is an F -automorphism of E . Hence $\sigma^* \in G(E/F)$. So $\sigma^*(K) = K$. Hence $\sigma(K) = K$ because σ^* is an extension of σ . We have already shown that if $\sigma: K \rightarrow \bar{F}$ is an embedding that keeps each element of F fixed, then $\sigma(K) = K$. Hence K is normal extension of F . This proves the statement.

Hence, K is a normal extension of F if and only if for all $\sigma \in G(E/F)$ and $k \in K$, $\sigma(k) \in K$.

Again suppose that K be a normal extension of F , and let σ be an arbitrary element of $G(E/F)$. Then for all $k \in K$, $\sigma(k) \in K$. Further, for all $\tau \in G(E/K)$, we have

$$\tau(\sigma(k)) = \sigma(k)$$

$$\Rightarrow (\sigma^{-1}\tau\sigma)(k) = k \text{ for all } k \in K$$

Thus $\sigma \in G(E/F)$, $\tau \in G(E/K) \Rightarrow \sigma^{-1}\tau\sigma \in G(E/K)$.

Therefore, $G(E/K)$ is a normal subgroup of $G(E/F)$.

Conversely, suppose that $G(E/K)$ is a normal subgroup of $G(E/F)$. Then for all

$$\sigma \in G(E/F), \tau \in G(E/K) \Rightarrow \sigma^{-1}\tau\sigma \in G(E/K).$$

$$\Rightarrow \sigma^{-1}\tau\sigma = k \text{ for all } k \in K$$

Thus, $\tau(\sigma(k)) = \sigma(k)$ for all $\tau \in G(E/K)$, for all $k \in K$. As explained earlier, E is a normal extension of K . Hence K is the fixed field of $G(E/K)$.

Thus

$$\tau(\sigma(k)) = \sigma(k) \Rightarrow \sigma(k) \in K.$$

$$\forall \sigma \in G(E/F) \text{ and } k \in K.$$

Therefore, K is a normal extension of F .

This proves (iv)

(v) Let K be a normal extension of F . Then, $\sigma \in G(E/F)$, $\sigma(K) = K$. So σ induces an automorphism $\sigma^*: K \rightarrow K$ defined by

$$\sigma^*(k) = \sigma(k), k \in K$$

Now we define a suitable mapping

$$f: G/F \rightarrow G(K/F)$$

Such that $f(\sigma) = \sigma^*$, $\sigma \in G(E/F)$.

Let $\sigma_1\sigma_2 \in G(E/F)$ Then

$$(\sigma_1\sigma_2)^*(k) = \sigma_1^*(\sigma_2^*(k)) = \sigma_1^*(\sigma_2(k)) = \sigma_1(\sigma_2(k)) = (\sigma_1\sigma_2)(k)$$

$$\text{So } \sigma_1^*\sigma_2^* = \sigma_1\sigma_2$$

$$\text{Again } \sigma_1, \sigma_2 \in G(E/F) \Rightarrow \sigma_1\sigma_2 \in G(E/F)$$

So,

$$(\sigma_1, \sigma_2)^*(k) = (\sigma_1, \sigma_2)(k) \forall k \in K$$

$$\Rightarrow (\sigma_1, \sigma_2)^* = \sigma_1, \sigma_2$$

$$\text{Hence } (\sigma_1, \sigma_2)^* = \sigma_1^*, \sigma_2^*$$

$$\Rightarrow f(\sigma_1\sigma_2) = f(\sigma_1)f(\sigma_2) \forall \sigma_1, \sigma_2 \in G(E/F)$$

Thus f is a homomorphism of $G(E/F)$ into $G(K/F)$. Now

$$\text{Ker } f = \{\sigma \in G(E/F) \mid f(\sigma) = \text{identity}\}$$

But $\sigma^* = \text{identity}$ if and only if $\sigma^*(k) = k$ for all $k \in K$.

That is, $\sigma(k) = k$ for all $k \in K$, so $\sigma \in G(E/K)$.

Hence $\text{Ker } f = G(E/K)$.

Then by the fundamental theorem of homomorphism

$$\frac{G(E/F)}{G(E/K)} = \text{Im } f \subset G(K/F) \quad \dots\dots\dots (1)$$

By (iii) we have

$$\frac{G(E/F)}{G(E/K)} = [K:F] \quad \dots\dots\dots (2)$$

Also, because K is normal over F , we have

$$|G(K/F)| = [K:F]. \quad \dots\dots\dots (3)$$

By (2) and (3), we get

$$\frac{G(E/F)}{G(E/K)} = |G(K/F)| \quad \dots\dots\dots (4)$$

By (1) $\frac{G(E/F)}{G(E/K)}$ is isomorphic to $\text{Im } f$ which is a subgroup of $G(K/F)$.

But, by (4) order of the group $\frac{G(E/F)}{G(E/K)}$ is equal to the order of the group $G(K/F)$.

So, the $\text{Im } f$ in $G(K/F)$ must be $G(K/F)$ itself.

Hence, by (1), we get

$$\frac{G(E/F)}{G(E/K)} = (K/F) \text{ which is the complete proof.}$$

Theorem 1.3

If $f(x) \in F[x]$ has r distinct roots in its splitting field E over F , then the Galois group $G(E/F)$ of $f(x)$ is a subgroup of the symmetric group S_r .

Proof: Let us suppose a polynomial $f(x) = a_0 + a_1x + \dots + a_nx^n \in F[x]$ of degree n have r distinct roots $\alpha_1, \alpha_2, \dots, \alpha_r$, where $1 < r \leq n$. We know that for each $\sigma \in G(E/F)$, $\sigma(\alpha_i)$ is again a root of $f(x)$. also, if $\alpha_i \neq \alpha_j$ then $\sigma(\alpha_i) \neq \sigma(\alpha_j)$.

So $\sigma(\alpha_1), \sigma(\alpha_2), \dots, \sigma(\alpha_r)$ is a permutation of $\alpha_1, \alpha_2, \dots, \alpha_r$. Set $\phi_\sigma(\alpha_i) = \sigma(\alpha_i)$, $i=1, \dots, r$. Then it is clear that $\phi_\sigma \in S_r$ the symmetric group of degree r , and we have mapping $f: G(E/F) \rightarrow S_r$ given by

$$f(\sigma) = \phi_\sigma \text{ for all } \sigma \in G(E/F). \dots\dots\dots (1)$$

IF $\sigma, \eta \in G(E/F)$, then

$$\phi_{\sigma\eta}(\alpha_i) = (\sigma\eta)(\alpha_i) = \sigma(\eta(\alpha_i)) = \sigma(\phi_\eta(\alpha_i)) = \phi_\sigma(\phi_\eta(\alpha_i))$$

for all α_i . Hence $\phi_{\sigma\eta} = \phi_\sigma \phi_\eta$. This implies, by (i), that

$$f(\sigma\eta) = f(\sigma)f(\eta) \text{ for all } \sigma, \eta \in G(E/F).$$

Hence, f is a homomorphism.

Next, we show that f is 1-1. To effect this, let $\phi_\sigma = \text{identity}$. Then

$$\phi_\sigma(\alpha_i) = \alpha_i; \text{ that is } \sigma(\alpha_i) = \alpha_i \text{ for all } i = 1, \dots, r.$$

Thus, $\sigma = \text{identity}$, since $E = F(\alpha_1, \dots, \alpha_r)$

Hence, f is an embedding of $G(E/F)$ into S_r .

Theorem 1.4

Every polynomial $f(x) \in C$ factors into linear factors in $C[x]$

Proof: Let $f(x) = a_0 + a_1x + \dots + a_nx^n \in C[x]$. Again then

$$g(x) = (x^2 + 1)(a_0 + a_1x + \dots + a_nx^n) \\ (\bar{a}_0 + \bar{a}_1x + \dots + \bar{a}_nx^n).$$

If $c_1, c_2 \in C$, then we find that $c_1\bar{c}_1 = |c_1|^2 \in R$ and $c_1\bar{c}_2 + \bar{c}_1c_2 = 2\text{Re } c_1\bar{c}_2 \in R$. It is clear that $g(x) \in R[x]$. Let E be splitting field of $g(x)$ over R .

Then $R \subset C \subset E$

At first there does not exist a subfield K of E containing C such that $[K:C] = 2$. Suppose such a subfield exists. Since K is a finite separable extension of C . $K = C(\alpha)$ for some $\alpha \in K$. If $p(x) \in C[x]$ is the minimal polynomial of α over C , then $\deg p(x) = 2$.

So, let $p(x) = x^2 + 2ax + b$. Then

$$p(x) = (x+a)^2 - (a^2-b) \\ = (x+a - \sqrt{a^2-b})(x+a + \sqrt{a^2-b}),$$

That is, $p(x)$ splits into linear factors in C , which is a contradiction because $p(x)$ is irreducible over C .

Let $G = G(E/R)$ be the Galois group of $g(x)$ over R . Let $|G| = 2^m q$, where q is an odd integer. Let H be the 2-sylow subgroup of G , and let L be the corresponding subfield of E . We have the following diagram which exhibit the foregoing consideration.

$$E \leftrightarrow \{e\}$$

$$L \leftrightarrow H$$

$$C \leftrightarrow G(E/C)$$

$$R \leftrightarrow G(E/R)$$

Then $[E:L] = 2^m$. But

$$|G| = [E:R] = [E:L][L:R] = 2^m [L:R]$$

$\Rightarrow [L:R] = q$, because $[G] = 2^m q$.

Also since L is a finite separable extension of R , $L = R(\beta)$ for some $\beta \in L$. So the minimal polynomial

$$q(x) = b_0 + b_1x + \dots + b_qx^q \in R[x]$$

of β is of degree q . But this is impossible unless $q=1$, because by the intermediate value theorem of Calculus every equation of odd degree over the real's has a real root. hence $q = 1$, so we have $[E:R] = 2^m$. But we know that

$$[E:R] = [E : C] [C : R] = 2 [E : C]$$

$$\Rightarrow [E : C] = 2^{m-1}$$

Then the subgroup $G(E/C)$ of $G(E/R)$ is of order 2^{m-1} . If $m > 1$, let H_1 be the subgroup of $G(E/C)$ of order 2^{m-2} , and let L_1 be the corresponding subfield of E . It is clear that $L_1 \supseteq C$. Then we find that $[E : L_1] = 2^{m-2}$, and so $[L_1 : C] = 2$ which is a contradiction.

Hence, we must have $m = 1$. Then $[E : R] = 2$.

But $[C : R] = 2$ and $E \supseteq C$ implies that $E = C$. This proves the theorem

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Motivation Plays an Important Role in Alcoholism Treatment: A Brief Review

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Abstract

Motivation plays an important role in alcoholism treatment by influencing patients to seek, complete, and comply with treatment as well as make successful long - term changes in their drinking. Both alcohol abusing and alcohol - dependent people can be classified into different “stages of change” in terms of their readiness to alter their drinking behaviour. In this paper, the concept of motivation and its influence on behaviour change, the role of motivation in alcoholism treatment and treatment methods designed specifically to influence motivation has been studied. The results indicate that motivation - based approaches can increase patient motivation to leave the habit.

Keywords: Motivation, alcoholism treatment, drinking habit and URICA scale

Introduction

Motivation is an important first step toward any action or change in behaviour. Sayings such as “You can lead a horse to water, but you can’t make it drink” reflect the fact that people generally will not perform desired behaviours unless or until they are motivated to do so. Until recently, many alcoholism treatment professionals used this approach when treating alcoholic patients, contending that interventions were useless until the alcohol - dependent patient was self - motivated to change his or her drinking behaviour.

During the past several years, however, researchers and clinicians have shown increased interest in the concept of motivation and the role that motivation plays in recovery from alcohol problems.

Researchers have outlined a series of stages of change to describe the process that a person goes through when making a behavioural change. Those stages—precontemplation (i.e., not yet considering change), contemplation (i.e., considering change but not taking action), preparation (i.e., planning to change), action (i.e., making changes in one’s behaviour), and maintenance (i.e., changing one’s lifestyle to maintain new behaviour)—offer a new perspective on motivation and the process of behaviour change (DiClemente and Prochaska 1998; Prochaska et al. 1992). Recognizing that patients vary in their motivation or readiness to change, researchers has designed interventions and treatments to enhance motivation (DiClemente et al. 1992; Higgins and Budney 1993; Miller and Rollnick 1991; Miller et al. 1992; Stitzer et al. 1993).

In this paper, the concept of motivation and its influence on behaviour change, the role of motivation in alcoholism treatment and treatment methods designed specifically to influence motivation has been studied.

Stages of Change and the Trans - theoretical Model of Change

A number of studies (Carney and Kivlahan 1995; DiClemente and Hughes 1990) have demonstrated that people with alcohol and other drug problems that seek or participate in treatment differ significantly in their levels of motivation to change. Using an alcohol version of the University of Rhode Island Change Assessment (URICA) scale, DiClemente and Hughes (1990) reported on the various stages of change among alcohol - dependent patients seeking outpatient treatment. Patients in the precontemplation stage were more likely to deny that they had a drinking problem, Patients in the preparation and action stages were more likely to admit that they had a drinking problem, stating, for example, “I am actively working on my alcohol problem” and “I have a problem and I really think I should work on it.” Using the same scale, Carney and Kivlahan (1995) found similar profiles among a large group of substance abusing veterans. The same variations in motivation have been found in other treatment-seeking populations, including inpatient substance abusers (Isenhardt 1994) and polydrug users in methadone maintenance treatment (Belding et al. 1995). Recognizing these differences was the first step to evaluating how differences in motivation affect participation in treatment programs and drinking outcomes. Assessment of motivation presents a significant challenge. External influences and pressures, as well as internal thoughts and feelings, contribute to a person’s motivation both to consider and implement a change in behaviour (Cunningham et al. 1994). Evaluating a person’s motivation requires assessment of the person’s attitudes and intentions, confidence and commitment, and decision making ability figure (1).

Researchers have attempted to measure motivation in several different ways, including querying patients about their intentions and plans to change and asking multiple questions reflecting the different stages of change. Clinicians and researchers are attempting to intervene earlier with problem drinkers and design programs to recruit and motivate unmotivated patients. Such programs are designed to address specific tasks and obstacles that arise at the different stages of change. To move from the precontemplation stage, the patient must admit to having an alcohol problem and recognize the need to change his or her drinking behaviour.

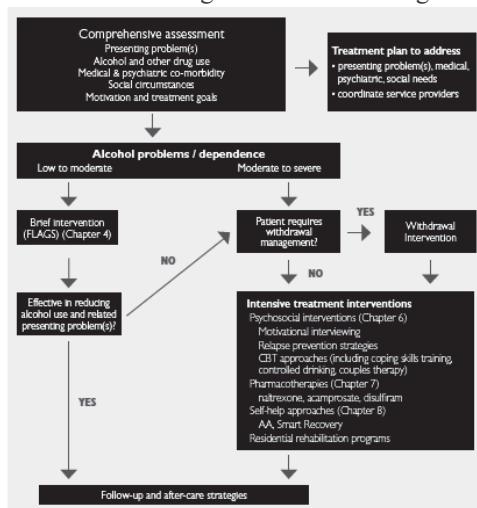


Figure 1: Major points of comprehensive assessment and treatment planning for alcohol use disorders.

Assessment of motivation presents a significant challenge. External influences and pressures, as well as internal thoughts and feelings, contribute to a person's motivation both to consider and implement a change in behaviour (Cunningham et al. 1994). Evaluating a person's motivation requires assessment of the person's attitudes and intentions, confidence and commitment, and decision making ability. Researchers have attempted to measure motivation in several different ways, including querying patients about their intentions and plans to change and asking multiple questions reflecting the different stages of change. As an outcome of the revised perspective on the concept of motivation, clinicians and researchers are attempting to intervene earlier with problem drinkers and design programs to recruit and motivate unmotivated patients.

Such programs are designed to address specific tasks and obstacles that arise at the different stages of change. In the preparation stage, the patient increases his or her commitment to change and plans to take action. In the action stage, the patient develops specific behavioural strategies to change his or her drinking behaviour. Finally, in the maintenance stage, the patient strives to avoid relapse by developing a lifestyle that supports the changes in his or her

drinking. For successful recovery, patient motivation is important throughout the entire process, although it is an especially important focus during the first three stages.

Motivation appears to be a critical dimension in influencing patients to seek, comply with, and complete treatment as well as to make successful long-term changes in their drinking habit. Studies among substance-abusing patients have demonstrated the importance of motivation for treatment in predicting treatment participation and recovery (DeLeon et al. 1997). Motivation for changing problem behaviours like drinking, however, is not synonymous with motivation for participating in treatment. Many patients enter treatment under pressure from other people.

Although these patients may attend treatment, they may not be ready to change their drinking behavior and may not actively participate in treatment. Most substance abuse treatment programs and self-help initiatives are designed to assist patients who are ready to take action and address their problems.

Depending on the type of program and the intensity of the examination procedures before admission, however, people who are not ready to change or who are in the early stages of change are often admitted into these programs. Therefore, most clinicians need to know how to handle unmotivated or reluctant patients who enter treatment and who are, at best, ambivalent about changing their behaviour. Traditional approaches to treating unmotivated patients with alcohol problems often use aggressive and confrontational strategies in response to the patients' denial. In one widely used approach, a team of family members, friends, and colleagues unite to confront the drinker and convince him or her that alcoholism treatment is necessary.

Ryan and colleagues (1995) found that among people who received outpatient alcohol treatment, internal motivation (as assessed by a treatment motivation questionnaire) was related positively to both treatment involvement and retention. Among the study subjects, the outpatients with high levels of both internal and external motivation had the highest treatment retention and treatment attendance outcomes. Irrespective of their level of external motivation, outpatients with low internal motivation had the worst treatment outcomes. Finally, patients with more severe alcohol problems generally had greater internal motivation for treatment. The severity of the patient's alcohol problems enhances internal motivation, presumably because the problem severity increases distress and thus influences decision making. Although internal motivation appears to be more effective for long term success, external motivation seems to promote short-term abstinence from alcohol and other drugs. Relying solely on external pressure and incentives to influence a patient to modify his or her drinking

behaviour, however, can be difficult. A study of 263 inpatients in alcohol treatment found that patients whose motivations to enter treatment were related to current external threats (e.g., threatened loss of job, driver's license, or spouse) experienced better treatment outcomes than did patients without such motivating factors (Krampen 1989). However, patients who had experienced one of the aforementioned losses in the past before entering treatment had poorer treatment outcomes than patients who had not experienced such losses. Because of the limited effectiveness of external motivators, treatment providers face the challenge of shifting patients' motivation from external to internal incentives.

Motivational Treatment Approaches

Treatment approaches designed to enhance patients' intrinsic motivation include brief intervention, motivational interviewing, and motivational enhancement therapy (MET).

1. Brief Motivational Intervention

Brief (i.e., single - session) motivational intervention uses straight forward advice and information on the negative consequences of alcohol abuse to motivate patients to reduce or stop drinking. This treatment approach has generally been viewed as more relevant for problem drinkers who are not yet alcohol dependent than for alcohol - dependent drinkers. Brief interventions vary in duration from one to four sessions, with each session lasting from 10 to 60 minutes. The goal of brief intervention is often reduced drinking rather than abstinence. Physicians or other treatment providers, advise patients on the need to reduce their alcohol consumption and offer feedback on the effects of the patients' drinking. The feedback is designed to increase patient motivation to reduce or stop drinking. Unlike more traditional treatment approaches, this technique does not involve overtly confrontational tactics but rather consists of a respected professional giving the patient advice and providing personally motivating information (Miller and Rollnick 1991). Researchers are still investigating whether certain types of patients benefit more from brief intervention than do other patients. Such research could suggest whether these interventions can benefit patients with more severe problems or additional psychopathology. At present, brief interventions appear to work best to motivate middle - aged excessive drinkers with more serious, multiple problems to enter and participate in more extensive treatment.

2. Motivational Interviewing

For less motivated patients with alcohol problems, a motivation-enhancing technique known as motivational interviewing (MI) may be more beneficial than either self-help books or cognitive behavioural interventions (Heather et al. 1993). Based on motivational psychology and the stages-

of-change model, MI focuses on enhancing and facilitating the patient's internal motivation to change. This approach assumes that the patient is responsible for changing his or her addictive behaviour and recognizes ambivalence as a natural part of the process. In contrast to confrontational approaches, MI is designed to assist patients in working through their ambivalence and in moving toward positive behavioural change. The MI therapist uses various techniques to help increase the patient's motivation to change his or her behaviour. The MI therapist emphasizes the patient's personal choice regarding change, deemphasizes diagnostic labels, and avoids arguing with and confronting the patient. The MET approach was specifically developed for Project MATCH. MET combines MI techniques with the brevity of a less intensive intervention. MET consists of four treatment sessions over 12 weeks preceded by an extensive assessment. In the first session the therapist provides the patient with clear, structured, personalized feedback concerning his or her drinking frequency (number of drinking days per month), drinking intensity (number of drinks per drinking occasion), typical level of intoxication, risk for negative consequences of alcohol use, results of liver function and neurological tests, and risk factors for alcohol problems (e.g., familial risk and tolerance symptoms). This information comes from scores on various measures and diagnostic tests that the patient has completed before the session.

The patient's scores are then compared with the scores of a reference group of patients to increase the patient's awareness of the extent to which alcohol has affected his or her life and to motivate the patient to change his or her drinking behaviour. The patient receives a copy of the feedback report to take home. Patients' readiness to change at the start of treatment had a significant impact on their success in quitting and reducing drinking throughout the 3 years after treatment. Figure 2 shows alcohol withdrawal syndrome progression.

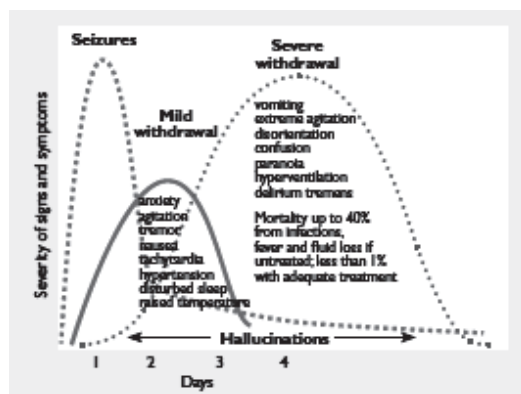


Figure 2: Alcohol withdrawal syndrome progression

Conclusion

Motivation, a key element in treatment and recovery, influences a patient's progression through the stages of change - from considering change, to making the decision to change, to following the planned action into sustained recovery. Current research and treatment initiatives reflect the increased focus on the role of motivation in alcoholism treatment. Efforts to intervene effectively with alcohol problems in primary care settings, court diversion programs, prison programs, and more traditional inpatient and outpatient treatment programs must address client motivation. In fact, as early identification and intervention programs become more proactive and aggressive, the importance of addressing patient motivation will only increase. In general, motivated patients enter and attend treatment at higher rates than do less motivated patients. However, some extrinsically motivated patients may attend treatment regularly but be reluctant to participate in the treatment program. Other minimally motivated patients may attend and participate to some degree but fail to make substantial changes or sustain changes made in treatment. Both the type and intensity of the patient's motivation for change are important potential moderators of treatment participation and recovery success. An increasing number of treatment strategies and programs are being used to address the patients' motivational needs. Some programs have established groups or initial program components to examine and increase motivation. Research indicates that motivation - based approaches can increase patient motivation and improve drinking outcomes, researchers and clinicians still have much to learn about how to influence patient motivation whether intrinsic or extrinsic. The next few years should yield interesting and important information about the feasibility and effectiveness of methods to motivate and move patients through the stages of change in all types of alcohol and other drug treatments, whether brief or more intensive.

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