

*PRACTICAL WORKBOOK*

*FOR*

*QUANTUM AND STATISTICAL*

*PHYSICS*

*What nation expects from its technocrats is their indigenous innovations evolving into marketable products which would boost the economy and living standards of fellow citizens. This is the sole mantra for evolution of developing country like ours.*

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## ***STUDENT'S INFORMATION***

Name of Student:

Branch (Section):

Roll Number:

Practical Group (Subgroup):

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## **FOREWORD**

Physics and Mathematics form the backbone of engineering studies and it inspires the student for strong conceptualization of basic principles so as to have better insight and keep update of rapidly growing technology. The subject of physics is related to not only getting conversant with theoretical concepts and their applications, rather experimentation also plays a vital role, which the juvenile engineering undergraduates need to appreciate. Laboratory work inculcates many virtues in a student who works religiously. The rational approach with alertness, zeal, innovation and discipline uplifts the mental faculties and also mindset to deal with any day to day problem.

The introductory laboratories, at the under-graduate level, are usually designed to serve objectives such as (a) demonstration of physical concepts or phenomena (b) familiarization with equipment and techniques and (c) systematic training to plan, devise and perform experiments. The first objective related to the demonstration of physical ideas or phenomena not only helps greatly in its understanding but also provides an insight about the order of magnitude of various physical quantities involved. However the demonstration can never be an alternative to proper explanation of the phenomenon where all physical quantities are dealt more closely and realistically. The second objective served by such laboratories is perhaps more important. In the laboratory work, many instruments and equipment are used by students. Although no course can acquaint a student with all kinds of instruments, but they do help in gaining experience as well as developing a scientific approach to handle them in general. The third objective plays a vital role in training a student to plan, devise and perform experiment. Any experiment that is

devised and subsequently performed must adhere to important basics such as: (i) it must always be planned such that precision of measured physical quantities is appropriate to the purpose (ii) all necessary steps must be taken to eliminate systematic errors in technique and instruments employed (iii) measurements and calculations must be recorded accurately, clearly and concisely (iv) critical analysis of results is necessary to draw correct conclusions and also estimate the precision of the final result.

To do well in laboratory work, a student must learn and develop the confidence to take initiative. To get the best out of a physics laboratory, one must utilize lab hours efficiently to have threadbare clarity of details of activities. The approach of the student must be to understand all possible details related to measurement and equipment employed which should be then properly correlated to theory or phenomenon being studied. The sensitization of the student to account for precautions and sources of errors appropriately and judiciously plays a vital role. It should be kept in mind that the results of the experiments must be evaluated promptly and discussed with the instructor. No experiment is complete till academic gains from it are analyzed in terms of technique, experimental skill and its applications.

With these words of advice and caution, I welcome the student community to take up this workbook and provide feedback necessary for addition or corrections and updation.

- J.K.Goswamy

## **ACKNOWLEDGEMENTS**

This workbook has been prepared as a tool accessible to a community of students, teachers and instructors for the smooth conduct of physics experiments in the laboratory. It has been prepared keeping in view the syllabus of practicals in Physics of Materials as approved by Board of Studies for Applied Sciences of Panjab University, Chandigarh. The preparation of this workbook was inspired by the objectives to (a) provide a ready reference for various aspects of different prescribed experiments, (b) lay down a prescription for smooth running of physics laboratory, (c) make available a workbook to students wherein observations and experiences, while performing the experiments in the laboratory, can be recorded.

The workbook comprises of three sections which address data analysis, measuring instruments and experiments related to physics of materials. The Section A, related to the data analysis, introduces the student to the spirit of physical measurements, errors associated with them and analysis of such errors. Further the mathematical tools such as significant digits, graphical method and curve fitting procedures which help in judicious representation and sensitive interpretation of data have been discussed. The subsequent section B introduces various common laboratory tools and equipment used in the experiments. This section is expected to help students to have greater insight about aspects of instrumentation, thereby sharpening their skills. The section C introduces experiments related to physics of materials. In each of the experimental activity, theory, apparatus, procedure of performance, precautions cum sources of errors have been spelt out explicitly. Further readymade tables for recording observations and subsequently

evaluating the results have been provided. At the end of each experiment, some activities related to the concerned experiment have been suggested so as to broaden the perspective of students.

During the preparation of this manuscript, the author mainly relied on his experience of teaching at the undergraduate level. In addition to the own experience, the critical reviews, discussions and guidance offered by Prof. Devinder Mehta, Prof. S.K.Tripathi, Prof. C.N. Kumar and Prof. G.S.Saini (all working as faculty in the Department of Physics, Panjab University, Chandigarh) during the preparation of this workbook has helped greatly in shaping this document. The cooperation and suggestions extended by my colleagues Dr. Monika Randhawa, Dr. Shuchi Gupta, Dr. Jyoti Sood, Dr. Geetu, Dr. Mamta Sharma, Dr Suresh and Dr. Sunil have been of immense value. Last but not the least, the continuous feedback from students has played a vital role in bringing this workbook to its present form. I hope that my students, working in their laboratory, will keep on providing useful suggestions for ever remaining scope of improvement in various aspects of this workbook.

- **J.K.Goswamy**

## **TEN GOLDEN RULES FOR LABORATORY WORK**

1. Laboratory is a temple with work ethics and doesn't bar anyone from working.
2. Punctuality and regularity does strengthen the laboratory basics of students.
3. Time spent efficiently enhances the confidence and sharpens experimental skills.
4. Be peaceful and converse in brief while working in the laboratory.
5. Learn to work in small groups of two (or three) students in a coherent manner.
6. Learn the correct way to handle the equipment and record observations.
7. Take care to minimize personal, gross and systematic errors. Mistakes do not find a place in the laboratory work.
8. Keep an alternate check on each recorded observation. This defines genuity of an experimenter.
9. Evaluate results and draw conclusions without delay. Never forget to discuss them with your instructor.
10. Laboratory rewards those who strive for knowledge and its extension.

# FORMAT FOR WRITING EXPERIMENT ON NOTEBOOK

LEFT HAND SIDE OF NOTEBOOK (Write only with pencil only)	RIGHT HAND SIDE OF NOTEBOOK (Write with pen only)
<ul style="list-style-type: none"><li>• <i>Objective of Experiment.</i></li><li>• <i>Formula used with meaning of various symbols.</i></li><li>• <i>Circuit or block diagram (No sketch of equipment).</i></li><li>• <i>Observations with emphasis on units.</i></li><li>• <i>Results and Error analysis with emphasis on units.</i></li><li>•</li></ul>	<ul style="list-style-type: none"><li>• <i>Objective of Experiment.</i></li><li>• <i>Apparatus required without description of equipment.</i></li><li>• <i>Formula used with meaning of various symbols.</i></li><li>• <i>Precautions and Sources of errors.</i></li><li>• <i>Conclusions drawn from the experiment.</i></li><li>•</li></ul>

# RECORD OF EXPERIMENTS PERFORMED BY STUDENT

Name of the Student:

Roll No.:

Branch (Semester):

Section:

Practical Group:

Sub-Group:

S.No.	Experimental Activity	Date of		Teacher's	
		Allotment	Completion	Awards	Signatures
1.	Frank-Hertz Set-up.				
2.	Planck's Constant.				
3.	Photoresistor.				
4.	Photovoltaic cell.				
5.	Rutherford scattering experiment				
6.	Probability law and statistics				
7.	Probability using coins				
8	Michelson Interferometer				
9	Millikan oil drop experiment				
10	Thomson's e/m ratio				
11	Hall Effect				
12	Four Probe Method				
13	Ultrasonic interferometer				

# DATA ANALYSIS

*(Forms the basic strength of a sound experimentalist)*

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# **A1: PHYSICAL MEASUREMENTS AND ERROR ANALYSIS**

## **A1.1: The Physical Observation and its Analysis**

The modern science and technology is the output of scientific experiments involving measurements. Well designed experiments followed by careful analysis produce a body of scientific facts which are unquestionable. While science does contain a large body of well established facts, still there are many phenomena which are not well understood primarily because of the lack of sufficiently well designed experiments coupled with properly analyzed results. Many concepts and quantities can be established by two or more very different types of investigations. The regularity of nature is such that, when two different scientists carry out different types of experiments to establish a given concept, they must arrive at the same conclusion provided that they are sufficiently competent and each of them carries out a sufficiently careful analysis of his data. Nevertheless and unfortunately scientific journals and books contain many instances where one statement related to a given concept conflicts the other. Such difficulties emphasize the phrases such as "sufficiently competent scientists" and "sufficiently careful analysis". Careful analysis sometimes reveals that, although two types of experiments were designed to measure same quantity, they do in fact measure different quantities.

## **A1.2: The Physical Observation and the Error**

The subject of physics, in the aforementioned spirit, establishes the facts on the basis of experimental observations, which involve direct or indirect measurement of various physical quantities. Laboratory work in physics is primarily concerned with measurement of physical quantities. Many people think that the physical measurements, that physicist makes in his laboratory, are exact and consequently the physics is an exact science. It is not true and no physical measurement can ever be exact. In general, the accuracy of a measurement is limited by the degree of refinement of the instrument and the skill of the observer. Even with the highly skilled observer using the most refined instrument, the measurement can't be exact in the absolute sense. This is so, because every observation in a measurement is a product of human judgment and sophistication of instrument used and these can never be a subject of mathematical exactness.

If we have a close look at the basics of physical measurement, then a little consideration will show that every physical measurement reduces, essentially, to reading a scale provided with the instrument being used. The scale is already calibrated in the appropriate units of the physical quantity being measured. The task of the observer is to judge that a specific point or mark is nearer to a particular graduation mark than the other one. For example, the points A, B and C are marked on a scale, shown in the figure A1.1. The points A and B will be read as 2.2 while C will be read as 2.5.

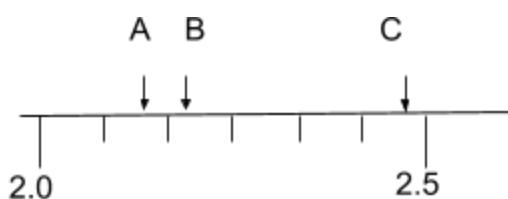


Fig. A1.1

If the specified mark lies halfway between two graduations, then the one with the higher value is recorded. Making an observation in this manner, an observer makes a maximum possible error of half a scale division and this can be recorded in the observation with the  $\pm$  sign. Thus the observation A, as seen in the fig.A1.1, will be represented as  $2.2 \pm 0.05$ . Such a notation expresses the limit within which the value of observation may lie.

### **A1.3: Possible Error and the Least Count**

It should be realized that the measured value of any physical quantity is the difference of two observations – the initial (or zero) reading and the final reading. For example the length of an object is determined by reading a meter scale at both ends of the object. If the position of two ends of the object is recorded as  $2.4 \pm 0.05$  cm and  $9.2 \pm 0.05$  cm respectively, then the length of the object is 6.8cm. If it is realized that initial and final positions of the object can lie anywhere between 2.35cm to 2.45cm and 9.15cm to 9.25cm, then the actual length of the object will lie anywhere between 6.7 to 6.9cm. This length, in notation of errors, can be expressed as  $(6.8 \pm 0.1)$  cm. It should be noted that the measurement of length has an uncertainty of 0.1cm which is equal to one scale division or the least count. This is actually true in the case of every physical measurement that the possible error in the measured value is equal to the least count of the measuring instrument used. In addition to this inherent uncertainty, there may creep further uncertainties in the measured value because of numerous other causes.

### **A1.4: Error and its Representation**

The complete statement of the measured value of a physical quantity should include not only the magnitude of the physical quantity but also an estimate of the uncertainty in that

magnitude. Technically the error is an uncertainty in the magnitude of the measured value of the physical quantity. The uncertainty in the measured value of a physical quantity is defined relative to a *reference or true value* which is the result of (i) theoretical prediction or (ii) measurement made by a sophisticated instrument or (iii) average evaluated from a very large number of measurements.

A discrepancy, on the other hand, is the difference between the two measured values of a physical quantity. It is worth mentioning here that the errors creep into a measurement due to causes of varied nature and must not always relate to the observer's skills. The errors in different kinds of physical measurements can be represented in the following different manner:

- (i) **Absolute Error** is the magnitude of the difference between the true value of the quantity and the individual measurement. If true or standard value is known, then we have:

$$\text{Absolute Error} = \Delta a = a - a_m \quad (A1.1)$$

where  $a_m$  is the measured value and  $a$  is the true or standard value. If true value is not known, then the arithmetic mean of large number of measurements is taken as the true value.

- (ii) **Relative Error** is the ratio of absolute error in the physical measurement to the standard or true value of physical quantity being measured.

$$\text{Relative Error} = \frac{a - a_m}{a} \quad (A1.2)$$

- (iii) **Percentage Error** is the form of the relative error when expressed as a percentage.

$$\text{Percentage Error} = \left( \frac{a - a_m}{a} \right) \times 100 \quad (A1.3)$$

## A1.5: Classification of Errors

The error in a measurement of a physical quantity can result due to many different causes and these lead to the classification of errors:

- (a) **Least Count Error:** This is the error associated with the least amount of measurement that an instrument can make. This error can be reduced by using a sophisticated or refined instrument capable of measuring the physical quantity.
- (b) **Setting Errors:** In some experiments, it is not possible to make a desired adjustment (or setting) with sufficient precision because the experimental set-up lacks the desired sensitivity. Such errors are incurred commonly while obtaining no parallax situation in optical experiments or achieving balance point in the null experiments. These errors are purely due to instrument and they can be estimated by finding out how small change in the measured quantity can the given arrangement detect or measure. Setting errors are always greater than the least count of the experimental set-up in use.
- (c) **Systematic Errors:** These errors arise due to definite discoverable phenomena. Their cause can be traced and subsequently accounted optimally. Usually such errors produce the same kind of uncertainty in all the measurements i.e. either it lead to higher or lower values of observation in proportional amounts. Systematic errors can arise due to the following reasons:

- (i) *Error due to Alteration in the Measured Quantity:* An error is introduced when the quantity is being altered by the very act of its measurement. For example, the current measured by an ammeter gets altered by the finite resistance of the ammeter itself. Ideally a quantity should be so measured that its magnitude is minimally altered by the method employed. A more practical way is to measure the altered magnitude of the quantity and then correct it for the disturbance. Thus if the resistance of the ammeter is known, then value of measured current can be corrected appropriately.
- (ii) *Instrument Error:* These are caused by faulty or inaccurate instruments. For example an instrument may have erroneous or worn off calibration. The repeated measurements with the same instrument will fail to reveal the presence of such errors. These are better detected by change of two similar instruments or by use of different methods to measure the same physical quantity.
- (iii) *Errors due to Imperfect Technique:* These errors occur by the very nature of experimental arrangement used. For example heat loss due to radiation in calorimetry experiments or weight loss of the body due to buoyancy of surrounding medium. When known to be present, suitable correction can be applied to account for such errors.
- (iv) *Errors due to External Conditions:* The external environmental conditions such as temperature, pressure, mechanical vibrations, line voltage fluctuations may have considerable effect on a measurement. These factors, if not properly controlled, influence the results in two ways: (a) calibration of the instrument may get affected or (b) the quantity being measured, may suffer a change. Elaborate means of controls

like constant temperature baths, humidity controllers, voltage stabilizers, shockproof mountings etc are, therefore used to achieve the desired accuracy in highly sophisticated experiments.

- (v) *Personal Errors:* An experimenter himself may unknowingly contribute errors to his results because of his habits and peculiarities of taking an observation. Many people suffer from the bias of regarding the first reading as correct and not paying much attention to the subsequent readings. Personal errors can be detected if the measurements are repeated by different observers.

The ability to detect the presence of systematic errors is of great importance in achieving the desired accuracy in the measurement. These errors are often hidden and are not revealed by merely making repetitive measurements. A thorough understanding, on the part of the experimenter, of all the features of the measurement being carried out is essential. He should be aware of the alternate techniques suitable for that measurement. Most important of all, he should be vigilant and inquisitive at every stage of the experiment to look for the factors that could possibly influence his results.

- (d) **Random Errors:** Even after elaborate precautions are taken to reduce the systematic errors so that they are small enough to be treated as negligible, it is found that the repeated measurement of a quantity yield values which are slightly different from each other. This shows that some kind of error still exists. These are known as random errors as they occur in random or irregular fashion, completely devoid of any regular pattern. These errors arise due to haphazard combination of large number of small

effects. To minimize random errors, we take large number of measurements of the same physical quantity and evaluate their arithmetic mean. Statistical methods are used for dealing with random or chance errors.

(e) **Gross Errors:** These are caused by the outright carelessness and sloppy habits of the experimenter. They have no place in the experiment and must be avoided by exercising due care. These errors arise due to sheer carelessness of the experimenter while reading the instrument or recording the observations or overlooking the prescribed precautions and sources of errors or committing the computational mistakes.

### A1.6: Precision vs. Accuracy

It is very important for students to understand the terms precision and accuracy as they are usually mistaken. The word precision relates to the random error distribution in repeated measurement of the same physical quantity. This term is often referred to a particular experiment or particular type of experiment. The word accuracy is related to the existence of systematic errors-differences between laboratories.

### A1.7: Propagation of Errors

The errors in the directly measured physical quantities result in erroneous values of other quantities which are derived from them through subsequent computations. This phenomenon is called propagation of errors. The physical quantities derived from directly measurable physical quantities involve certain basic arithmetic operations such as

addition, subtraction, multiplication, division and exponentiation. The maximum possible error in resultant quantity can be computed as follows:

**1. Addition or Subtraction:** If two measured physical quantities  $A = a \pm \Delta a$  and  $B = b \pm \Delta b$  are added or subtracted to obtain the resultant quantity C. Then maximum possible absolute error in the value of C is the sum of the absolute errors in the measured values of A and B.

$$\begin{aligned} &\text{If } C = A \pm B \\ &\text{then } \Delta c = \Delta a + \Delta b \end{aligned} \quad (A1.4)$$

**2. Multiplication or Division:** If two measured physical quantities  $A = a \pm \Delta a$  and  $B = b \pm \Delta b$  are multiplied or divided to obtain another quantity C. Then maximum possible relative error in the value of C is the sum of the relative errors in the measured values of A and B.

$$\begin{aligned} &\text{If } C = A \cdot B \quad \text{or} \quad C = \frac{A}{B} \\ &\text{then } \frac{\Delta c}{c} = \frac{\Delta a}{a} + \frac{\Delta b}{b} \end{aligned} \quad (A1.5)$$

**3. Exponentiation:** If we have a measured physical quantity  $A = a \pm \Delta a$  and another quantity C defined as  $C = A^n$ , then relative error in C is n times the relative error in A.

$$\begin{aligned} &\text{If } C = A^n \\ &\text{then } \frac{\Delta c}{c} = n \frac{\Delta a}{a} \end{aligned} \quad (A1.6)$$

**4. Trigonometric Functions:** If a physical quantity is expressed as trigonometric functions, then error will be of function and not of the measured angle. Let's consider that during measurement of angle  $\theta$ , an absolute error of  $\Delta\theta$  is induced. Then we can proceed in different situations as follows:

(i) If the physical quantity is expressed as  $Z = \tan\theta$ , then we have

$$\begin{aligned}\Delta Z &= \Delta\theta \sec^2 \theta \\ \frac{\Delta Z}{Z} &= \frac{\Delta\theta \sec^2 \theta}{\tan \theta} = \frac{2\Delta\theta}{2\sin\theta\cos\theta} \\ \frac{\Delta Z}{Z} &= \frac{2\Delta\theta}{\sin 2\theta}\end{aligned}\tag{A1.7}$$

(ii) If the physical quantity is expressed as  $Z = \sin\theta$ , then we have on similar lines

$$\frac{\Delta Z}{Z} = \frac{\Delta\theta}{\tan\theta}\tag{A1.8}$$

(iii) If the physical quantity is expressed as  $Z = \cos\theta$ , then we have on similar lines

$$\frac{\Delta Z}{Z} = \frac{\Delta\theta}{\cot\theta}\tag{A1.9}$$

## A1.8: Statistical Treatment of Errors

Let's consider that  $n$  measurements of a physical quantity are made which are all equally reliable. Their measured values are  $X_1, X_2, X_3, \dots, X_n$ . The arithmetic mean of this set of measured values will be given as

$$\bar{X} = \frac{1}{n} \sum_{i=1}^n X_i$$

The true value is defined as the mean of infinite set of measurements made under constant conditions and is denoted as  $\mu$ . As it is not practically possible to make infinite number of measurements of a physical quantity, so the number of observations must be sufficiently large.

The precision with which a physical quantity is measured depends inversely upon the deviation or dispersion of set of measured values  $X_i$  about their mean value  $\bar{X}$ . If values are widely spread, the precision is defined to be low. The deviation of the individual measurement from the mean value is defined as  $\delta_i = X_i - \bar{X}$ . The average or mean deviation can be expressed as:

$$\bar{\delta} = \frac{1}{n} \sum_{i=1}^n \delta_i$$

The standard deviation is defined as the square root of mean squared deviation and is denoted by

$$\sigma = \sqrt{\frac{\sum_{i=1}^n (X_i - \bar{X})^2}{(n-1)}} = \sqrt{\frac{\sum_{i=1}^n \delta_i^2}{n-1}}$$

The standard error can now be defined as

$$\sigma_m = \frac{\sigma}{\sqrt{n}}$$

The normal law of errors predicts the probability that mean value from a finite number of observations  $n$  may be in the interval  $\mu \pm \sigma_m$  is 68% and probability that it may lie in the interval  $\mu \pm 2\sigma_m$  is 95%.

**Degree of Accuracy:** In an experiment, all observations must be taken to the same degree of accuracy. It is of no use taking some observation to a much higher degree of accuracy than others because it will never improve the accuracy of final results. The degree of accuracy of the final results is guided by the least accurate observation. It does not imply that observations should not be measured accurately but a careful sense of proportion must be used. It should be taken care that average of certain number of measured values of a physical quantity can never be expressed with an accuracy better than least count of the equipment used.

## PROBLEM SET A1

**Problem A1.1:** Find the equivalent resistance of two resistors  $R_1 = (200 \pm 2)\Omega$  and  $R_2 = (300 \pm 6)\Omega$ , connected in parallel.

**Problem A1.2:** Calculate the value and maximum possible error of the refractive index of the material of a prism from the following observations: Angle of prism =  $A = 60^\circ \pm 0.5^\circ$ ,

Angle of minimum Deviation =  $D = 36^\circ \pm 0.5^\circ$

**Problem A1.3:** A current  $I = (10.0 \pm 0.5)$  A flows through a resistor  $R = (50 \pm 5)\Omega$ . Calculate the power dissipated in the resistance with the limits of possible error.

**Problem A1.4:** Find the focal length of a lens from the readings  $u=(20.1\pm 0.2)\text{cm}$  and  $v=(50.1\pm 0.5)\text{cm}$ .

## A2: DATA REPRESENTATION: SIGNIFICANT FIGURES

### A2.1: Concept of Significant Figures

In the previous chapter, the concept of error, its representation and classification followed by propagation of errors as a result of arithmetic compounding have been discussed. Any measurement of a physical quantity is depicted by a notation wherein the actual magnitude of physical quantity is specified along with the associated uncertainty. This representation, in general, is cumbersome when referred time and again. To avoid this inconvenience, a universal convention is adopted, where the measured component is depicted but the uncertainty, resulting from least count of the instrument, is dropped unless required specifically. If the length of an object, measured with a meter scale, is  $(6.8 \pm 0.1)$  cm, then it can be written as 6.8 cm. The measured value is interpreted as: (i) the length has been measured using a meter scale capable of measurement up to one-tenth of a centimeter (ii) the first digit 6 is certain (iii) the second digit 8 is doubtful and remains uncertain within one unit.

The measured value of a physical quantity depicts the accuracy of measurement which in turn is indicative of the refinement of the instrument used for the purpose. *The number of digits, which enter into the measured value through the actual measurement and on which*

*the observer has confidence, define the significant figures.* The measurement associated with greater number of significant figures is considered to be more accurate and reliable.

There are certain rules which aid in ascertaining the number of significant digits in a depicted measured value of the physical quantity. These rules are:

I<sup>st</sup> Rule: All non-zero digits are significant.

II<sup>nd</sup> Rule: Zeroes occurring between two non-zero digits are always significant.

III<sup>rd</sup> Rule: Zeroes to the left of the leftmost non-zero digit are never significant.

IV<sup>th</sup> Rule: Zeroes to the right of the rightmost digit are significant only if they come from some physical measurement.

## A2.2: Arithmetic of Significant Figures

So far we have been talking about the significant figures of quantities that can be directly measured with the help of an appropriate instrument, for example length, mass, time, temperature etc. What about the physical quantities which are computed using the directly measured physical quantities. Since the computation involves the arithmetic operations between numbers and here these operations deal with physically measured quantities, so our ideas need revision as far as significant figures are concerned. Certain conventions, stated below, have been generally accepted to make sure that we do not write meaningless-cum-surplus digits in our final results.

- Interpretation:** Unless otherwise stated it may be safely assumed that the last digit in the statement of a measured quantity is uncertain by one unit.

**2. Significant Figures and Conversion of Units:** Expressing the measured value of a physical quantity in different units will not alter the number of significant digits. For this purpose, the conversion to other units should be accompanied with multiples or sub-multiples of 10. For example, it is better to write a length measurement of 50.2m as  $5.02 \times 10^3$ cm or  $5.02 \times 10^{-2}$ km instead of 5020cm or 0.0502km. This is preferred as it preserves the number of significant digits.

**3. Rounding off:** Sometimes it is necessary to round off a given number to the one having lesser number of digits by discarding the unwanted digits. For this purpose, the following is the convention;

- (i) If the last digit is less than 5, then it is deleted without increment to the preceding digit. For example 15.654 is rounded off to two decimal places as 15.65.
- (ii) If the last digit is greater than 5, then its deletion is followed by incrementing the preceding digit by one. For example 7.687 is rounded off to two decimal places as 7.69.
- (iii) If the last digit is 5, then its deletion is followed by incrementing the preceding digit only if it turns from odd to even. For example 7.355 is rounded off to two decimal places as 7.36 by 7.345 is rounded off as 7.34.

**4. Addition/Subtraction Rule:** In the process of addition or subtraction of two measured values of the same physical quantity, the number of significant digits in the resultant is the smallest number of significant decimal places used in the sum or difference. For example the sum of 0.86cm +1.122cm+ 17.8cm results in length

19.782cm. Since the smallest number of decimal places is one, hence the resultant length will be rounded off to 19.8cm.

5. **Multiplication/Quotient Rule:** In multiplication or division of measured values of two physical quantities, the number of significant digits in the product or quotient is the same as the smallest number of digits in any one of these factors. For example if the length and breadth are 1.152cm and 20.6cm respectively, then area will be  $23.7312\text{cm}^2$ . As per the rule, the area will be expressed up to three significant digits i.e.  $23.7\text{cm}^2$ .
6. **Significant Digits in Multi-step Computations:** When a final step is to be obtained following several intermediate steps of calculations, it is advisable that an extra digit be carried in each value through all the computation steps till the final result. This extra digit is rounded off after the completion of calculations.

## PROBLEM SET A2

**Problem A2.1:** Indicate the number of significant figures in the following:

501, 0.5709, 0.00690,  $7.3 \times 10^4$ , 97.1000

**Problem A2.2:** Round off the following numbers as indicated:

- (i) 1526167 to 6 digits. (ii)  $1.996 \times 10^4$  to 3 digits. (iii)  $2.5946 \times 10^{-3}$  to 2 digits.
- (iv) 0.7995 to 1 digit. (v) 0.00385 to 2 digits.

**Problem A2.3:** Simplify the following with due regards to significant figures:

- (i)  $2 \times 10^{-5} + 2 \times 10^{-6}$  (ii)  $9.15 + 3.879 + 6.547 - 4.55$
- (ii)  $3.8 \times 0.125$  (iv)  $0.0036 \times 4.0 \times 10^4$
- (v)  $(3.75 - 0.501)^2$  (vi)  $\sqrt{(3.5 - 3.31)} \sqrt{(3.5 - 3.31)}$

# A3: DATA REPRESENTATION: GRAPHICAL ANALYSIS

## A3.1: What is a Graph?

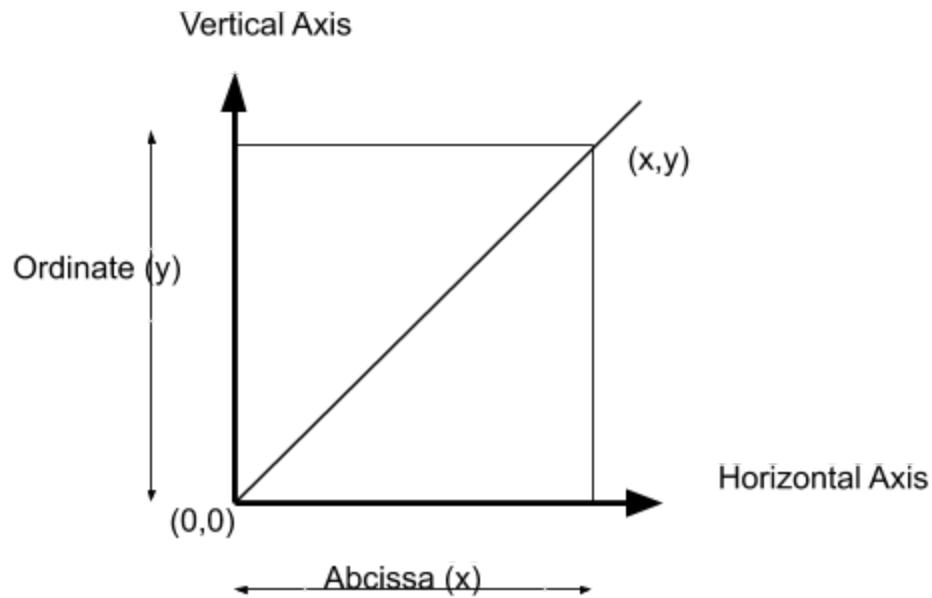
We know that the physical laws express relationship between quantities. Any such relationship can be depicted in one of the following convenient ways:

- (i) in words combined properly to frame a statement.
- (ii) in the form of a mathematical equation.
- (iii) by pictorial representation.

The relations are effectively expressed pictorially in the form of graphs. Graphical representation is not only a convenient way of reporting experimental data, but plays an important role in deducing useful information from it.

The experimental data is usually depicted in a plane bounded by two rectangular (Cartesian) coordinate axes. One of the two variable physical quantities can be varied at discretion and is usually referred to as *independent variable*. The value of the other physical quantity depends on (or is deduced from) the independent variable and is commonly called *dependent variable*. The dependent variable is mostly experimentally observed corresponding to certain value of independent variable. The horizontal axis, called the x-axis (*abscissa*), is used to represent the independent variable. The vertical axis, called the y-axis (*ordinate*), is used for the dependent variable. The location of a point on the graph is defined by its coordinates (x,y) relative to a defined origin (0,0) which

is usually defined at the intersection of two axes. The specimen graph is shown in the fig.A3.1 where two axes, origin and a plotted point are depicted.



**Figure A3.1**

## A3.2: Plotting of a graph

The graphical representation of data helps in pictorial view of relation between two physical quantities. Further in certain cases, the graphical variation is used to deduce one or more physical quantities. Consequently it becomes mandatory to list necessary features for a good graphical representation:

1. From the given set of experimental data, find the independent and dependent variables. The independent variable must be represented along the x-axis while the

dependent variable is plotted along the y-axis. The unit of each quantity must be specified (usually in parentheses next to or below the axis label).

2. For plotting a good graph, one must carefully determine the range of each of the variables and count the number of big squares (1cm x 1cm) available for each variable along their respective axes so as to choose a convenient scale. This choice is essential for good representation of data as the choice of wider scale amplifies the irregularities thereby affecting the shape of curve. However too narrow a scale restricts the degree of accuracy of depicted observation points. The scale chosen should also take care of the fact that the plotted points could be read easily. Further it must fit the data within at least one half to an entire graph sheet.
4. Data points must be clearly visible. They should either be encircled or a cross should be put neatly for their clear depiction.
5. The plotted points should always be joined by a regular free hand curve. The zigzag line connecting successive points is incorrect way to plot a curve through the representative points.
6. Experimental points along a graph almost never form a perfect line because experimental data is associated with uncertainties originating from varied sources. The relationship between the quantities represented on a graph is described by the smooth curve which best fits the data. This is called a best-fit curve. The curve should not necessarily pass through each data point, but should fit the general area of significance of the data, with equal numbers of points falling symmetrically above

and below the line. Note that the best-fit curve for any particular data set may not necessarily be a straight line.

6. One should prefer at least six to eight observations for plotting a graph. This enables the experimenter to ascertain the shape of the graph.
7. The quantities to be plotted along x- and y-axis should be labeled clearly at the top of the graph in a tabular form.

### A3.3: Linearization of Graph

Often the data obtained in an experiment are linearly related. In other words, the data will fit the theoretical equation  $y = mx + c$  where m defines the slope of line (a constant of proportionality between the variables x and y) and c is an algebraic offset or also called

y-intercept. The slope is a measure of the steepness of the line, and is given by  $m = \frac{\Delta y}{\Delta x}$ .

Note that graphs are not always straight lines. Even without the ability to fit the data to a straight line, however, a careful analysis of graph can reveal a wealth of information about the relationship between the physical quantities. With a little mathematical manipulation, one can recast the data into the form of a linear relationship. Whenever practicable, variables should be so chosen as to yield a straight line graph. Apart from the fact that the straight line is the simplest curve and therefore reduces the labor of drawing a smooth curve, it has several advantages; the computations are minimized, experimental errors are readily disclosed and finally, it leads to the formulation of a single equation representing the entire experimental data. As there are no standard rules to follow, it requires the

ingenuity of an experimenter to get a straight line graph from a given data. Broadly speaking, two types of situations are encountered:

**(a) Where the functional relations between measured quantities are already known**

For example the mirror formula is given as:

$$\frac{1}{u} + \frac{1}{v} = \frac{1}{f} \quad (A3.1)$$

One may plot  $x = \frac{1}{u}$  on the x-axis and  $y = \frac{1}{v}$  on y-axis to get a straight line whose

equation will be  $y = c - x$  with  $c = \frac{1}{f}$ . This equation has equal intercept on both the

axes. The above relationship could also be expressed in linear form as  $uv = f(u + v)$

where  $uv$  will be plotted on x-axis while  $(u + v)$  will be plotted on y-axis. The slope of

line will give the focal length  $f$  and such a line will pass through the origin.

**(b) When the Relation between the measured variables is unknown**

The graphical analysis offers a powerful tool to search for relationship between the measured quantities. Suppose we have obtained a set of values of a physical quantity  $v$  for different values of  $u$ . If we desire to obtain relation between the physical quantities, then we plot graph between  $v$  and  $u$  and proceed as follows;

- (i) If the points lie haphazardly and no smooth curve can be drawn, then these two quantities bear no relation and are independent of each other.
- (ii) If a smooth curve can be drawn, the relationship can be written as:

$$v = cu^n \Rightarrow \log v = \log c + n \log u \quad (A3.2)$$

Hence a plot of  $\log v$  vs  $\log u$  will yield a straight line and by measuring the slope and intercept, exact relation can be obtained. Special types of graph papers are available which facilitate the plotting of logarithm of quantities. In such graph papers, the markings along one or both axes are not uniform. When both the axes are marked logarithmically on a graph sheet, it is called log-log graph paper. It is used to plot the relation depicted in eq. (A3.2)

- (iii) If the above plot does not yield a straight line, the possible relation that can be tried is:

$$\begin{aligned} v &= ce^{bu} \\ \Rightarrow \log v &= \log c + bu \log e \\ &= \log c + (0.4343b)u \end{aligned} \quad (A3.3)$$

These types of relations are plotted on semi-log graph paper where y-axis is marked logarithmically.

### A3.4: Limitations of Graphical Analysis

Although graphical technique is good for visualizing the results of an experiment, yet it is not the best method to find the value of constants. A graph enables us to see how closely the plotted points lie on the expected curve defining the relation between the concerned physical quantities. If certain points lie far away they are rejected in the subsequent analysis. The greatest difficulty lies in not only drawing a mean line but also deciding which of the points lie on this line. In certain situations, there may be possible to draw

more than one line which appears to be equally good with respect to the plotted points. However each of these lines will yield a different set of constants. In such a situation, the graphical analysis must be supplemented with more accurate techniques involving fitting procedures.

## A4: DATA REPRESENTATION: LEAST SQUARES FITTING

### A4.1: Need of Fitting the Experimental Data

The observations obtained from an experiment may have one physical quantity ( $y$ ) varying linearly with another physical quantity ( $x$ ). If we take  $n$  observations, then we have  $n$  such sets  $(x_i, y_i)$  which when depicted graphically, will, in most cases, not yield a straight line passing through all the points. Certain observations sets  $(x_i, y_i)$  do not lie on a straight line due to errors creeping into the observations due to instrumental uncertainties, procedural flaws, limitations of technique employed and many other (un)known causes. Consequently the information inferred from the graphical representation of the experimental data will have errors propagated into them. In this situation, it becomes necessary to draw the best possible line so as to derive physical quantities with minimal uncertainties. This necessitates the need to understand the prescription of least squares fitting of the experimental data.

### A4.2: Method of Least Squares

Our data consists of measurements  $(x_i, y_i)$  of an independent variable  $X$  and a dependent variable  $y$ . We wish to fit them with an equation of the form

$$y = a + bx \quad (A4.1)$$

by determining the values of the coefficients  $a$  and  $b$  such that the discrepancy, between the values of our measurements and the corresponding values of  $y = f(x)$  given by

equation (A4.1), is minimized. These coefficients can't be determined exactly from finite number of observations. The criteria for minimizing the discrepancy and optimizing the estimates of coefficients  $a$  and  $b$  is given below:

(a) **Criteria of minimization:** For well chosen coefficients, the deviation

$$\Delta y_i = y_i - a - bx_i \quad (A4.2)$$

must be small. The sum of these deviations is not a good measure of how well we have approximated the data with our calculated straight line because the large positive deviations can be balanced by large negative ones to yield a small sum even when the fit is pretty bad. To avoid this, we consider the sum of squares of deviations as the criteria for minimization.

(b) **Method of Maximum Likelihood:** The data consists of a sample of observations extracted from a parent distribution, which determines the probability of making any particular observation. Let us define the parent coefficients  $a_0$  and  $b_0$  such that the actual relationship is

$$y(x) = a_0 + b_0 x \quad (A4.3)$$

We define the quantity  $\chi^2$  as

$$\chi^2 = \sum \left( \frac{\Delta y_i}{\sigma} \right)^2 = \sum \left( \frac{y_i - a - bx_i}{\sigma} \right)^2 \quad (A4.4)$$

and consider it to be the appropriate measure of goodness of the test. The optimum fit to the data is obtained by minimizing the  $\chi^2$ .

(c) **Minimizing  $\chi^2$**  : For minimizing the  $\chi^2$ , we equate its partial derivatives w.r.t. a and b to zero. Hence we get:

$$\frac{\partial}{\partial a}(\chi^2) = 0 \quad ; \quad \frac{\partial}{\partial b}(\chi^2) = 0$$

The results of above derivatives on rearrangement give

$$\sum y_i = \sum a + \sum bx_i = aN + \sum bx_i \quad (A4.5)$$

$$\sum x_i y_i = \sum ax_i + \sum bx_i^2 = a \sum x_i + \sum bx_i^2 \quad (A4.6)$$

where N is number of observation sets. We solve equations (A4.5) and (A4.6) for the coefficients a and b. Same formalism can be extended for least squares fitting of higher order polynomials. For such polynomials, the  $\chi^2$  will be minimized by differentiation w.r.t all constant coefficients associated with different powers of independent variables.

## PROBLEM SET A4

**Problem 4.1:** Experimental data for determination of the temperature T along a metal rod suspended between two constant temperature baths as a function of position of x along the rod. Calculate a and b using above given formalism.

Trial (i)	Position (x <sub>i</sub> )	Temperature (T <sub>i</sub> )	T <sub>i</sub> <sup>2</sup>	x <sub>i</sub> T <sub>i</sub>	x <sub>i</sub> <sup>2</sup>
1	1.0	16			
2	2.0	18			
3	3.0	37			
4	4.0	44			

5	5.0	58			
6	6.0	62			
7	7.0	64			
8	8.0	70			
9	9.0	99			
	$\sum x_i =$	$\sum T_i =$	$\sum T_i^2 =$	$\sum x_i T_i =$	$\sum x_i^2 =$

**Problem A4.2:** The volume occupied by the gas as a function of its pressure at a constant temperature was studied experimentally. The observations of pressure and inverse of volume occupied are given below. Obtain the least squares fitted curve and hence calculate its fitting coefficients.

<b>x<sub>i</sub>=Pressure</b>	<b>y<sub>i</sub>=1 / Volume</b>
88.75	0.062
86.25	0.060
82.65	0.058
78.75	0.055
72.75	0.051
69.25	0.049
66.25	0.047
64.00	0.045

## **SECTION B: LABORATORY INSTRUMENTATION**

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## B1: MEASUREMENT OF SMALL LENGTHS

While working in the physics laboratory, the student uses many instruments for routine measurements related to length, time, mass and other physical quantities. The knowledge of these instruments becomes essential for their judicious usage. This chapter is dedicated to the understanding of these equipments, with emphasis on physical concepts related to their construction and usage, which plays a vital role in the growth of a student as an experimenter.

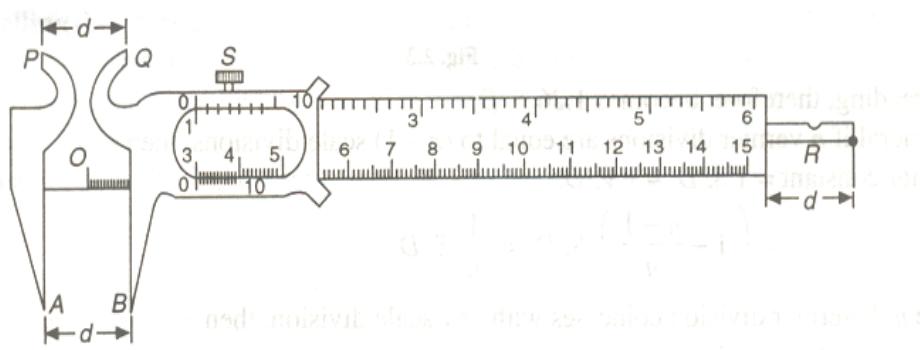
In the experiments, the common fundamental physical quantities being measured are length and angle which can often be quite small. The instruments like vernier calliper, micrometer screw gauge and spherometer are commonly employed for such measurements.

### B1.1: Measurement of Length and Angle

The measurement of length is necessary in almost all experiments in physics. Where great accuracy is not required, a meter rod can be used conveniently to measure all distances correct up to a millimeter. To measure a length correct up to a fraction of a millimeter, the divisions must be further subdivided. This is, however, not practical beyond a certain limit because it is difficult to see the small divisions clearly. Moreover, closer are the divisions on the scale, greater the inaccuracy in its graduation. Hence for measurements up to  $1/10^{\text{th}}$  or  $1/100^{\text{th}}$  of a millimeter, the various devices such as vernier callipers, screw gauge and spherometer are commonly employed.

### B1.1.1: Vernier Callipers

**Construction:** A vernier calliper consists of a rectangular steel bar, called the main scale, graduated in inches on one edge and in centimeters on other edge as shown below in the fig. B1.1.



**Fig.B1.1: Schematic diagram of Vernier calipers.**

A small scale, called the vernier scale, slides over the main scale. The instrument has two jaws of which one (say jaw A) is fixed at the end of the rectangular bar on the zero side, while the other (say jaw B) can slide on the main scale. Each jaw is at a right angle to the main scale and the movable jaw can be fixed at any position by the screw S. The two vernier scales are attached to the movable jaw as shown in the fig. B1.1. Usually when the two jaws are touching each other, the zero of the vernier scale coincides with that of the main scale. In some form of the instruments, the jaws also project in the upper part as shown at P and Q. These projecting jaws are used to measure the internal diameters of

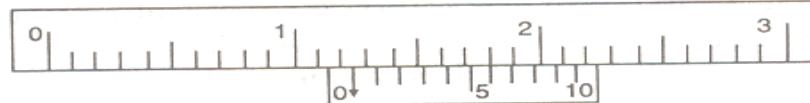
the tube or hollow samples with circular cross-section. The movable jaw also carries a thin rectangular rod R which is used to measure the depth of vessels.

**Vernier Constant:** To understand how the fraction of the smallest scale division is read accurately with the help of vernier, consider a vernier scale having 10 equal divisions. Let these 10 vernier divisions coincide with 9 divisions of the main scale.

$$1VSD = \frac{9}{10} MSD$$

$$VC = 1MSD - 1VSD = \frac{1}{10} MSD$$

The difference between one main scale division (MSD) and one vernier scale division (VSD) is called vernier constant (VC). Now suppose that the zero of vernier scale lies between 1.2 and 1.3 of the main scale as shown below in fig. B1.2.



**Fig. B1.2: Diagram showing the method to read the vernier caliper scale.**

It is clear that the 6<sup>th</sup> vernier division coincides with a scale division. If we denote the fraction after 1.2 by x, then

$$1.2 + x + 6VSD = 1.2 + 6MSD$$

$$x = 6(MSD - VSD) = 6VC = \frac{6}{10} MSD$$

The reading therefore becomes 1.26MSD.

**Circular Vernier in Spectrometers:** A spectrometer is used to measure refractive index by measuring the angle of prism and angle of minimum deviation. In this case 60 vernier scale divisions (VSD) coincide with 59 circular scale divisions (CSD) and the magnitude of each circular scale division is  $0.5^\circ$ . Hence Vernier constant can be expressed as:

$$1VSD = \frac{59}{60} CSD$$

$$\therefore VC = 1CSD - 1VSD = \frac{1}{60} CSD = \frac{1'}{2} = 30''$$

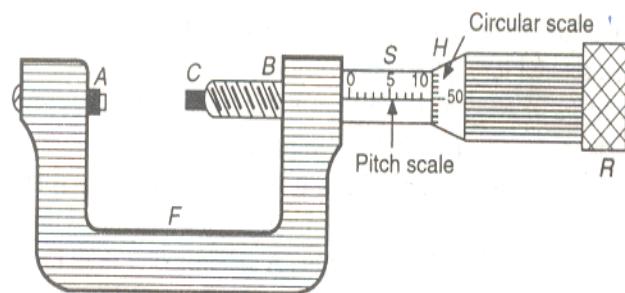
**Circular Vernier in Polarimeters:** A polarimeter is used for studying polarization of light and to determine specific rotation of optically active liquids or solutions. In this case each division on the circular scale is  $1^\circ$  and 10VSD coincide with 9CSD. Hence vernier constant is evaluated as:

$$1VSD = \frac{9}{10} CSD$$

$$\therefore VC = 1CSD - 1VSD = \frac{1}{10} CSD = 6'$$

### B1.1.2: Screw Gauge

**Construction:** It is based upon the principle of a screw and consists of a U-shaped frame F having a fixed end at A as shown in the fig. B1.3.



**Fig. B1.3: The Schematic Diagram of screw gauge.**

Through the other end B passes a fine and an accurately cut screw of a uniform pitch. A cap fits on to the screw and carries on its inner edge H, a circular scale 50 or 100 equal division marks, which is called head scale. This scale is used to measure the fraction of a revolution.

The number of complete revolutions can be read on the pitch scale S which is graduated on the nut parallel on the axis of the screw. The reading of the circular scale is taken against a line known as the reference line. A screw gauge is used to measure accurately small thicknesses or diameter of the wires. In some screw gauges, the screw head is provided with a ratchet arrangement R. When the studs A and C are in contact either with each other or with some other object placed in between, the ratchet slips over the screw cap without moving the screw forward. This avoids undue pressure on the studs or on the object.

**Pitch of the Screw:** Every screw is characterized by a quantity called pitch which is defined as the distance between two consecutive threads taken parallel to its axis and is measured by the distance through which the screw moves forward or backward when one complete rotation is given to the circular cap. To find the pitch of the screw, we proceed as follows: (a) Rotate the circular scale H and coincide its zero mark with the reference line. Note the reading on the pitch scale (b) Give a fixed number of complete rotations to the circular scale and again note the reading on the pitch scale. (c) The pitch of the screw is given as the ratio of distance moved by the screw on the pitch scale to the number of complete rotations given to the circular scale. The pitch of screw is usually 0.5mm or 1mm.

**Least Count:** The least count of a screw gauge is defined as the distance through which the screw moves forward or backward when the circular cap is rotated through one division on the circular scale. The least count of the screw gauge is defined as:

$$\text{Least Count} = \frac{\text{Pitch of the screw}}{\text{Number of divisions on the circular scale}}$$

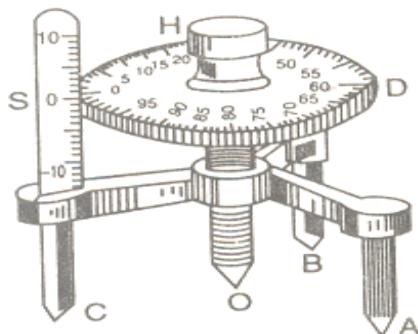
**Zero Correction:** If we bring the jaws A and C in contact without applying any undue pressure, the zero of the circular scale may not coincide with the reference line. In such a case there is a zero error. In some instruments, the zero mark goes beyond the reference line while in others it is left behind. To find the zero correction, count the number of divisions on the circular scale by which the zero mark has advanced beyond or left behind the reference line. This number multiplied by the least count is the zero correction. If the zero of the circular scale has advanced beyond the reference line, then zero correction is positive but if it is left behind the reference line, it is negative in nature.

**Backlash Error:** In instruments such as screw gauge or spherometer, in which the principle of screw is used to measure small distance, the accurately cut screw moves in or out of the nut. There is always a misfit or play between the two which increases with constant usage. Due to this play, when the direction of motion of the screw is reversed, the axial motion of the screw does not take place until the screw has been rotated through a certain angle. This lag between the linear motion of the screw and circular motion given to its head is called backlash error. This error can be avoided by turning the screw in the same direction while taking the observation.

### B1.1.3: Spherometer

A spherometer is an instrument used to measure very small distances. It also works on the principle of a micrometer screw. It is generally used to measure the thickness of thin plate and for determining the radius of curvature of spherical surfaces such as a lens or a mirror.

**Construction:** It consists of metal framework supported on three fixed legs of equal length. The ends of three legs lie on the three vertices of an equilateral triangle. An accurately cut screw works through a threaded hole at the centre of the framework. The screw terminates at the top into the milled head and carries a large graduated disc as shown in the fig.B1.4. The lower end of the screw forms the central leg of the instrument. A small vertical scale marked in millimeters or half-millimetres is fixed at one end of the frame with graduations close to those of circular disc. The edge of the circular disc is divided into a large number of equal divisions, generally 50 or 100.



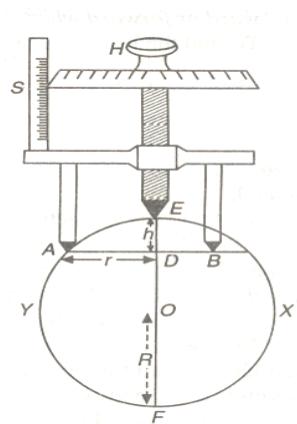
**Figure B1.4: Schematic Diagram of a spherometer.**

**Least Count:** The least count of the spherometer is the smallest distance that can be measured with it. It is the distance moved by the screw when it is turned through one division on the circular scale. The least count of the spherometer can be determined as

follows: (a) Bring the zero of the circular scale against a division marked on the main scale  
 (b) Give some fixed number of complete rotations to the screw and find the distance moved on the main scale (c) Divide this distance moved by the number of complete rotations given to find the pitch of the screw (d) Divide the pitch of the screw by the number of divisions on the circular scale to obtain the least count of the spherometer.

**Radius of Curvature of the Spherical Surface:** The spherometer can be used to find the thickness of the glass slip or radius of curvature of spherical surface. The theory underlying the measurement is illustrated below:

Let AEXFY be the sphere of which the given spherical surface forms a part. Consider a vertical section passing through one of the legs say A, the tip of the screw E and centre of the sphere O. It will be circle as shown in the fig. B1.5. The third leg of the spherometer will not be visible in this position of the spherometer. The tip of the central screw touches the spherical surface at E and if the surface were plane it would touch at D. Thus  $DE = h$  represents the height of the central screw above the plane of the outer legs ABC.

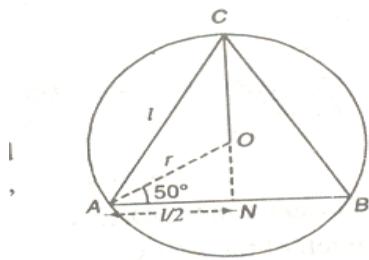


**Figure B1.5: Measurement of depth of curvature.**

The section of the sphere through ABC will be a circle of radius AD=r. From the geometry of the circle, we have

$$\begin{aligned}AD^2 &= ED \cdot DF \\R &= \frac{r^2}{2h} + \frac{h}{2}\end{aligned}\Rightarrow r^2 = h(2R - h) \quad (B1.1)$$

If L is the length of the side of the equilateral triangle ABC formed by joining the tips of three outer legs, then as shown in the fig. B1.6 below, we will have



**Figure B1.6.**

$$\frac{L}{2} = r \cos 30 = \frac{r\sqrt{3}}{2} \quad r = \frac{L}{\sqrt{3}} \quad (B1.2)$$

Putting (B1.2) in (B1.1), we get :

$$R = \frac{L^2}{6h} + \frac{h}{2} \quad (B1.3)$$

Thus the value of R can be calculated from the measurements of L and h.

## **B2:LABORATORY ELECTRICAL MEASUREMENTS**

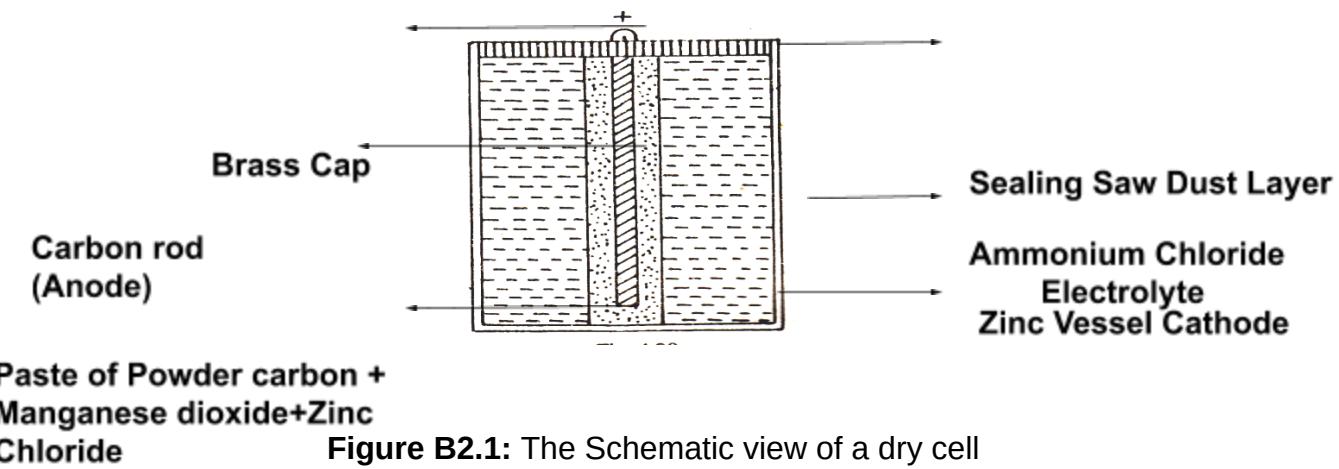
The undergraduate laboratories undertake many experimental activities which involve measurement of electric current, voltages and resistances. All these experiments utilize set-ups which are constituted by electric/magnetic circuit elements (such as resistors, inductors capacitors, electromagnets and coils), measuring instruments (such as galvanometer, tangent galvanometer, ammeter, voltmeter, magnetic compass, search coils etc) and sources of electrical energy (such as cell, battery, battery eliminator, constant voltage or current supply). These circuit elements may be in different forms and it is necessary to understand them from construction and working point of view for their justified usage. This chapter is dedicated to making aware the students of basics of electrical and magnetic circuit elements.

### **B2.1: Sources of Electric Energy**

#### **B2.1.1: Dry Cell**

The dry cell is easily portable and convenient source of direct current requirements for electrical circuits used in a laboratory. It is a modified and miniaturised form of Lechlanche cell. The schematic diagram of dry cell is shown in the figure B2.1. It consists of a cylindrical zinc vessel containing a moist paste (electrolyte of the cell) comprising of plaster of paris, flour sal ammoniac (ammonium chloride) and zinc chloride. The purpose of adding zinc chloride to the paste is to maintain the moisture as it is hygroscopic in

nature. Zinc vessel acts as a negative electrode and ammonium chloride is the exciting fluid. The anode of the cell is a carbon rod placed in the middle of the zinc vessel.

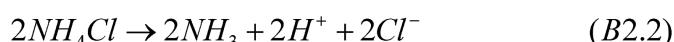


**Figure B2.1:** The Schematic view of a dry cell

The carbon rod is surrounded by a mixture of manganese dioxide and powdered carbon.

The manganese dioxide acts as a depolarizer while the powdered carbon reduces the internal resistance of the cell. The carbon rod is fitted with a brass cap which acts as a positive electrode. The carbon rod is effectively insulated from the zinc vessel by means of a tar paper washer. The top of the cell has a layer of saw dust which acts as a bed for the sealing layer of bitumen. The gases formed by chemical action escape via vents in this layer.

When the current is drawn from the cell by closing the external circuit, the chemical reactions that take place are

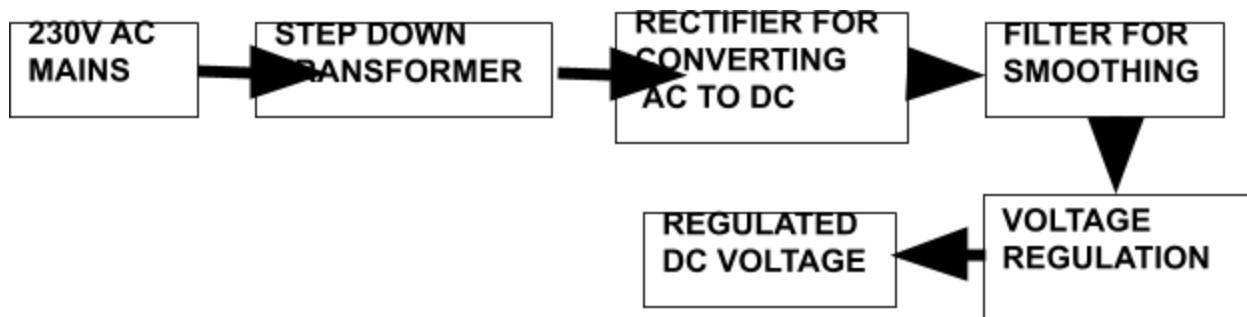


Irrespective of the size of cell, the emf of this cell is 1.5V. It is advisable to draw not more than 0.25A of current from this cell. On drawing excessive current, the polarization process

is set in. On continued use of this cell, the cell gets polarized but recovers if allowed to rest for some time. The internal resistance of this cell varies from 0.1 to slightly more than  $1\Omega$  which also depends upon ageing of cell.

### B2.1.2: Constant Voltage Power Supply

The constant voltage power supply is commonly used in the laboratory as a source of regulated DC voltage. There are different types of power supplies, most of which are designed to convert high voltage from AC mains to a suitable voltage output for electronic circuits and other devices. The figure B2.2 depicts a block diagram of common 5V power supply indicating its major components, each of which performs a specific function.

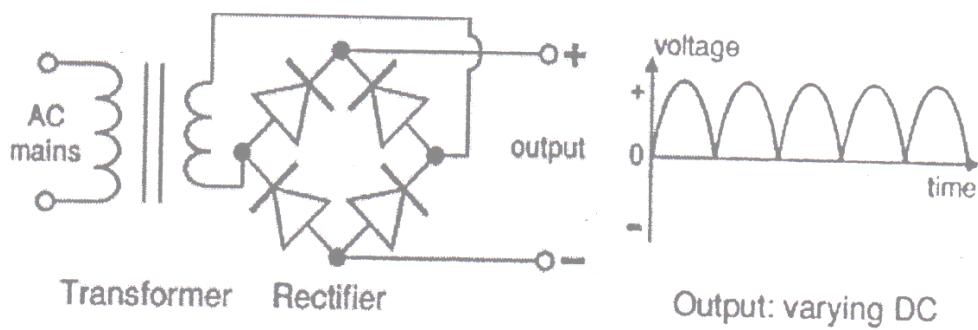


**Figure B2.1:** Block diagram of the components of constant voltage power supply.

The quality of the output depends upon the various components used.

**Transformer only:** These power supplies are used in those experiments where a lower AC voltage is required. This lower AC voltage is derived by using an appropriate step-down transformer. These are the most primitive form of the power supplies.

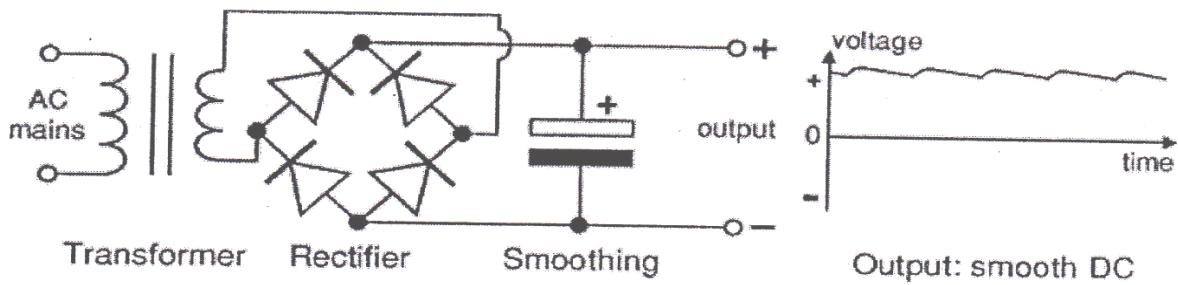
**Transformer plus Rectifier:** Such power supplies have output voltage which is pulsating DC and is preferred in those experiments where fixed DC voltage is not a requirement in stricter sense. The stepped down voltage is supplied as input to the bridge rectifier, comprising of four diodes, which uses both the half cycles of the AC input to generate pulsating DC output. The bridge rectifier is rated by the maximum current it permits and the maximum reverse voltage that it can withstand. This type of power supply circuit is shown in the figure B2.2 along with its output voltage.



**Figure B2.2:** The circuit diagram and output voltage of a transformer plus rectifier.

**Unregulated DC Voltage Power Supply:** The fixed DC can be obtained from the pulsating DC voltage from a bridge rectifier, a filter circuit (which may in the form of RC, inductive, capacitive, pi-circuit) is used. The filter circuit smoothens the pulsations to yield an output which more or less resembles the fixed DC voltage. The smoothing action of the filter circuit is accomplished by using an electrolytic capacitor which acts as a reservoir and supplies current when output from the rectifier is falling. The capacitor charges quickly near the peak of the input AC wave and discharges as it supplies the current to the output.

The circuit diagram of an unregulated power supply along with its output voltage is shown in the figure B2.3. Such power supplies, when connected to external circuit leads to fall in voltage due to loading and termed as unregulated power supply.



**Figure B2.4:** The circuit diagram and output voltage of constant power supply.

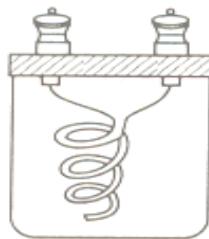
**Constant Voltage Regulated Power Supply:** If the output of unregulated voltage supply is connected to regulator circuit then the resulting voltage does not show any fall when connected to the external circuit. Voltage regulator IC are available in the market with fixed (typically 5V, 12V and 15V) or variable voltage. These are rated by the maximum current they allow. Most regulators have automatic protection against the flow of excessive current (overload protection) or overheating (thermal protection).

## B2.2: Common Electrical Instruments

### B2.2.1: Resistance Box

**Resistance Coil:** To measure the resistance of a conductor, it is compared with a known standard resistance in the form of a coil. The resistance of a standard coil should not vary with temperature. Such coils are made of thin insulated wires of alloys like constantan,

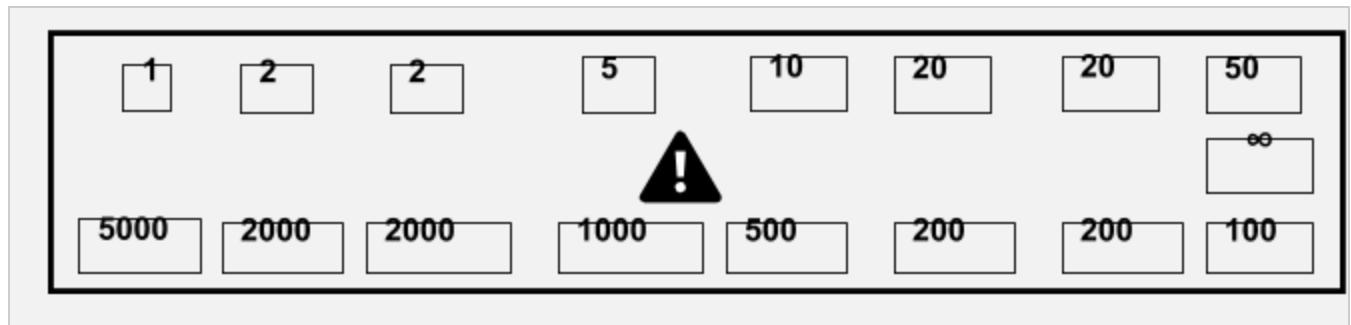
eureka, manganin, german silver etc which are characterized by (i) very low temperature coefficient of resistance to avoid any practical temperature variation of resistance (ii) very high resistivity so as to consume smaller lengths of wire. To prepare a non-inductive resistance, the calculated length of wire is double folded and wound carefully on a reel. The schematic view of a resistance coil sealed in an insulating box is shown in the figure B2.5.



**Figure B2.5:** The schematic diagram of a resistance coil sealed in an insulating box.

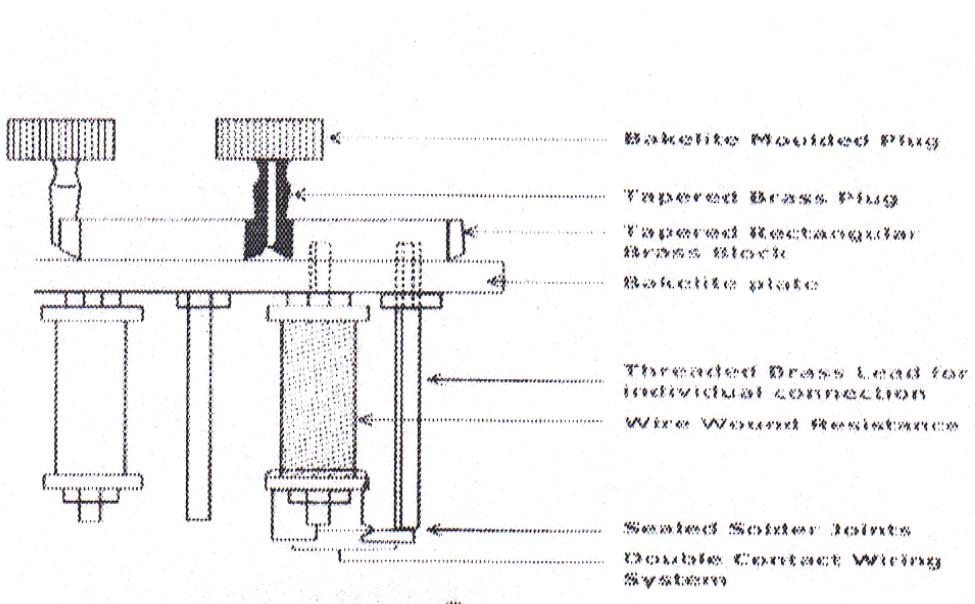
The two free ends of the wire are soldered to two terminals as shown in the figure B2.5. The terminals are fixed to ebonite plate and whole arrangement is enclosed inside a box.

**Resistance Box:** The resistance box is a device for introducing a known resistance in the electrical circuit. It consists of a wooden box having an ebonite top (an insulator material) on which are fixed stout brass pieces separated by a small distance. A conical hole is drilled in between two stout brass pieces into which fits a plug key. Standard non-inductive resistance coils are placed in the box. The two ends of the coil are soldered to the two adjoining brass pieces and value of the resistance is marked at the top. The arrangement of various coils in a resistance box is shown in the figure B2.6.



**Figure B2.6:** Arrangement of resistance coils in a resistance box.

The resistance gets introduced when the plug key between two brass pieces, to which the ends of the resistance coil are connected, is removed. When the plug key is in position, the current passes through the brass pieces which have practically no resistance owing to their large area of cross-section offered to the flow of current. The figure B2.7 shown below depicts the schematic connections of a resistance coil to the adjoining brass pieces and appearance of resistance box.



**Figure B2.7:** The schematic connection of a resistance coil to brass pieces,

### B2.2.2: Rheostat

It is an adjustable resistance placed in electrical circuits to regulate the flow of electric current. It consists of a porcelain hollow tube on which is wound certain length of insulated resistance wire connected at either end to a terminal. Over the porcelain tube is fixed a bar carrying a terminal  $T_1$  at one end and a sliding contact  $S$  as shown in the figure B2.8 below. The length and diameter of the wire depends upon the maximum resistance to be introduced and the maximum current to be passed. To use a rheostat as a variable resistance, one contact is made with the terminal  $T_1$  fixed to the bar while others made variable with one of the two terminals  $T_2$  to which the ends of the wire are connected. The resistance in the circuit can be varied by moving sliding contact along the bar. The maximum resistance that can be introduced in the circuit and the maximum current that can be passed through it are generally marked on the instrument.



**Figure B2.8:** The schematic picture of a rheostat.

### B2.2.3: Keys

These are used for opening or closing an electrical circuit. There are three different kinds of keys, which find use in electric circuit making.

The **plug key** has a base of ebonite sheet having two brass bars mounted on it and a conical hole drilled between them. If the metallic plug is inserted in this hole, the current flow starts and circuit gets closed. In the same manner, the removal of plug opens the circuit. Another such type is the **two way key** is used to change the path of current. This is essentially a combination of two plug keys.

Third kind of key commonly used is the **tapping key** in which a ebonite base sheet has two terminals in the form of a short metallic bar fixed at one end while at a little distance is fixed a flat, flexible and long piece of metallic plate hinged to ebonite plate. The two ends of the electric circuit are connected to these terminals. When long metallic bar is bent to bring it in contact with other terminal, the circuit gets closed. If left to itself, this long metallic bar retracts back and circuit opens. This key is used for sending current momentarily.

Fourth kind of key is the **reversible key** which can be used to reverse the direction of current in the circuit by changing the polarity of the voltage source.

## B2.3: Common Measuring Instruments

### B2.3.1: Weston Type Galvanometer

It is a common type of moving coil galvanometer used in laboratory for detection and measurement of electric current. It is based on the principle that a current carrying coil when placed in magnetic field suffers deflection due to torque exerted on it. The

magnitude of this deflection depends upon (i) number of turns in the coil (ii) current passed through the coil and (iii) strength of external magnetic field. When current is passed through the coil placed in radial magnetic field, due to permanent magnet with concave shaped pole pieces, it suffers deflection torque. The coil deflects under the influence of this torque until the opposing restoring torque counterbalances it. In Weston type galvanometer, the coil made of thin insulated wire is mounted on pivots. The coil is kept in position between concave pole pieces of a strong permanent magnet by two phosphor bronze springs. The two springs are soldered to the two ends of the coil and serves as leads for current. A light pointer is fixed rigidly to the coil and its movement over the scale measures the deflection. This galvanometer is endowed with following advantages: (a) as the coil is placed in strong magnetic field due to permanent magnet, the magnetic field due to earth does not influence its deflection. Consequently this instrument can be placed at any position unlike tangent galvanometer. (b) The strength of current is directly proportional to the deflection and hence the scale is linear. (c) It is dead beat instrument and comes to position of rest instantaneously. This is achieved by placing the soft iron core between the pole pieces of the permanent magnet with a narrow air gap for the free motion of the coil. The motion of the coil due to the flow of current in it produces an induced current in the core which opposes the motion of the coil and makes it dead beat.

### **B2.3.2: Ammeter**

The galvanometer is a low resistance device which can, in general, sense the current or measure small currents. The passage of large current through the galvanometer can cause damage because of (i) large deflection torque exerted or (ii) excessive heating of

the coil. The galvanometer is modified so as to measure large currents flowing through it.

For this purpose, a shunt resistor is placed in parallel combination to the galvanometer.

Let's suppose that we are to convert a galvanometer of resistance  $G$  and full scale deflection current  $I_g$  to an ammeter of range zero to  $I$  amperes. When current  $I$  is allowed to pass through this modified circuit, then the maximum current  $I_g$  (that generates full scale deflection in galvanometer) flows through the galvanometer while remaining passes through the shunt resistor. Hence mathematically we have:

$$I_g G = (I - I_g)R \Rightarrow I = \left[ \frac{G + R}{G} \right] I_g \quad (B2.4)$$

The value of shunt resistance determines the range of ammeter. The ammeter is a low resistance instrument having resistance of a few ohms.

### B2.3.3: Voltmeter

The voltmeter is an instrument used to measure potential difference between two points in a given circuit. The galvanometer, being a sensitive instrument, is modified so as to measure the potential difference. The ideal voltmeters are those which draw practically no or very small current in measuring voltage between two points in a circuit. The galvanometer is modified by connecting a high resistance of suitable value in series with the galvanometer.

Let's suppose that we have to convert a galvanometer of resistance  $G$  and full scale deflection current  $I_g$  to a voltmeter of range zero to  $V$ . If we apply a potential difference  $V$

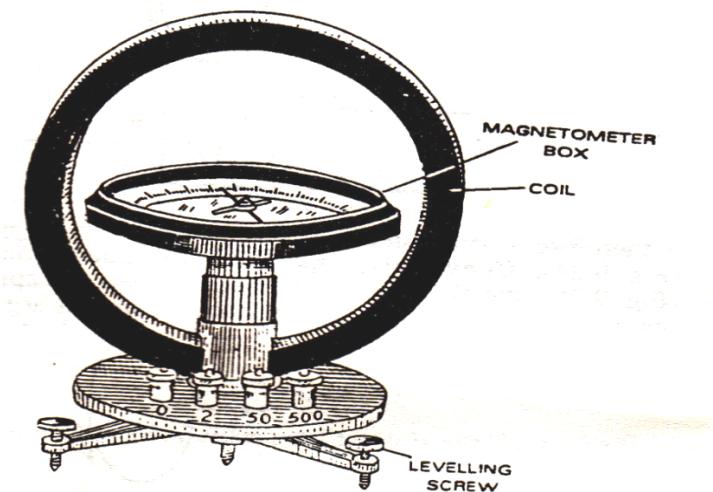
across the series combination of galvanometer and high resistance, then the current  $I_g$  is flows, then we can write as:

$$V = I_g(G + R) \quad (B2.5)$$

The voltmeter is high resistance instrument.

#### B2.3.4: Tangent Galvanometer

**Construction:** It is used for measuring the electric current. The schematic view of the tangent galvanometer is shown in the figure B2.9.



**Figure B2.9:** The schematic view of a tangent galvanometer,

It consists of a circular coil of large number of turns of insulated copper wire wound on a vertical circular frame of brass or wood. There are generally three separate coils of 2, 50 and 500 turns. The ends of these coils are connected to suitably marked terminals fixed at the horizontal base or platform of the instrument. The vertical frame is mounted on a horizontal platform and is provided with three leveling screws for alignment. The vertical

frame can be rotated about vertical diameter as its axis. A magnetic compass needle is pivoted at the centre of the vertical circular coil. Since the magnetic field due to earth is uniform over a very small area so it is necessary to take a small magnetic needle which remains suspended in the region of uniform magnetic field even when it is in the deflected position. The magnetic needle can rotate freely in a horizontal plane over a horizontal circular scale graduated in degrees. This scale is divided into four quadrants i.e.  $0^\circ$ - $90^\circ$ ,  $90^\circ$ - $0^\circ$ ,  $0^\circ$ - $90^\circ$  and  $90^\circ$ - $0^\circ$ . To read the deflection of the needle on a scale, a long and light aluminium pointer is attached at a right angle to the magnetic needle. In order to avoid parallax error, a plane mirror is fixed below the horizontal scale. Both the needle and the scale are enclosed in a compass box with a glass cover. The compass box saves the magnetic needle from the effect of air current.

**Working:** For making measurements with the tangent galvanometer, its coil is made vertical by using three leveling screws attached with the horizontal platform. The vertical coil is then rotated so that it gets oriented parallel to the magnetic needle and hence lie the plane of magnetic meridian. Now the compass box is rotated so that the aluminium pointer reads  $0^\circ$ - $0^\circ$  at two ends.

When the current is passed through the coil, it generates a magnetic field ( $B$ ) in a direction perpendicular to the plane of coil and also orthogonal to the horizontal component of magnetic field due to earth ( $H$ ). The magnetic needle is under the influence of two magnetic fields which are mutually perpendicular. These two fields result in torques of opposing sense on the magnetic needle which deflects and attains equilibrium at a new orientation (say the deflection is  $\theta$ ). At this orientation, we have:

$$\mu B \cos \theta = \mu H \cos(90 - \theta) \Rightarrow B = H \tan \theta \quad (B2.6)$$

Here B is the magnetic field due to current carrying coil of radius r at its centre and is given as:

$$B = \frac{\mu_o n I}{2r} = GI \quad (B2.7)$$

Putting equation (C1.7) in (C1.6), we get :

$$GI = H \tan \theta \Rightarrow I = \frac{H}{G} \tan \theta \quad (B2.8)$$

We define  $K = \frac{H}{G}$  = Reduction factor of galvanometer

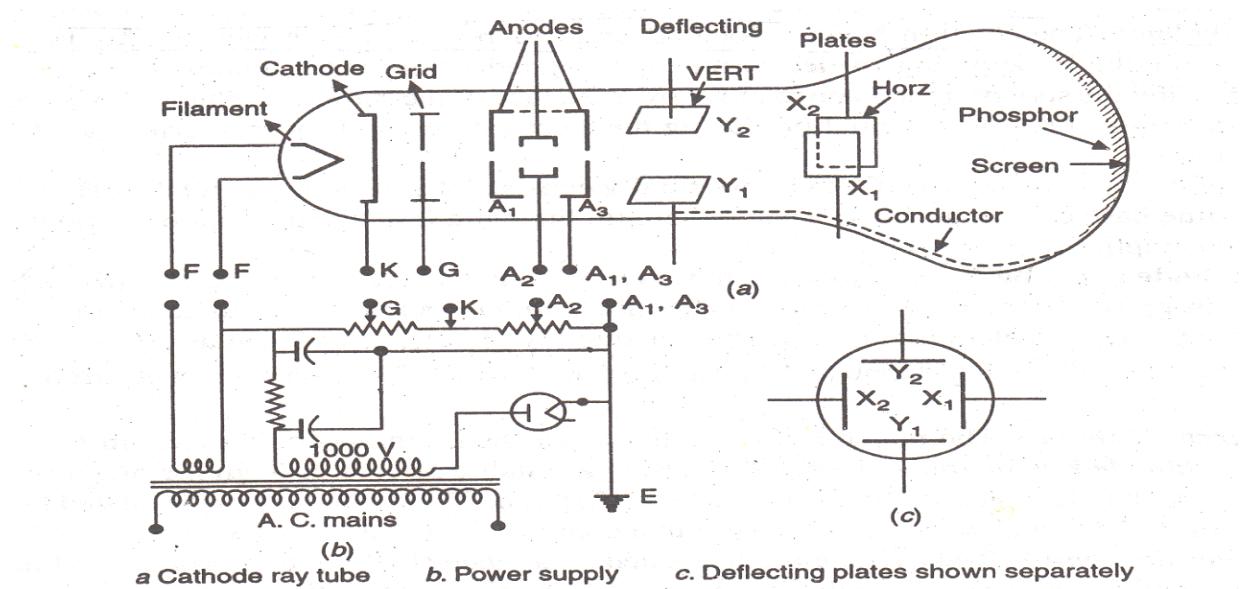
$$I = K \tan \theta \quad (B2.9)$$

The electric current is directly proportional to the tangent of angle of deflection. This explains as to why this instrument is called the tangent galvanometer.

### B2.3.5: Cathode Ray Oscilloscope (CRO)

The cathode ray oscilloscope is an instrument which can record instantaneous value of rapidly varying voltages. This instrument employs a finely collimated electron beam which is made to impinge upon the fluorescent screen after passing through applied electric and magnetic fields. The deflection caused by these applied fields results in tracing of a visible pattern by the electron beam on the fluorescent screen. This traced pattern yields information about potential due to the electric field or current causing the magnetic field. The CRO consists of a cathode ray tube (CRT), a sweep or time base circuit, a synchronization circuit, high and low voltage power supplies, horizontal and vertical amplifiers. These components are described in the following subsections:

**(a) Cathode Ray Tube:** The cathode ray tube (CRT) is an evacuated glass tube of the shape shown in the figure B2.10. It comprises of the following units:



**Figure B2.10:** The schematic picture of CRT showing various parts.

- (i) *Electron Emitter:* The electron beam is obtained by indirect heating of a cathode (K) by passing a suitable current through filament (FF). The cathode surface is coated with MgO layer so as to increase the yield of electrons by thermionic emission.
- (ii) *Intensity Control:* The cathode is surrounded by a metal shield in the form of a cylinder with hole at the end farther from the cathode. This cylindrical electrode or the grid (G) is maintained at a negative potential with respect to the cathode so that it repels the electrons so as to collimate them to a fine beam coming out of hole. The variation of negative potential on the grid controls the electron flux emitted from the cathode and thus helps in controlling the beam intensity or brightness. This control is usually labeled intensity or brightness on the front panel of CRO.

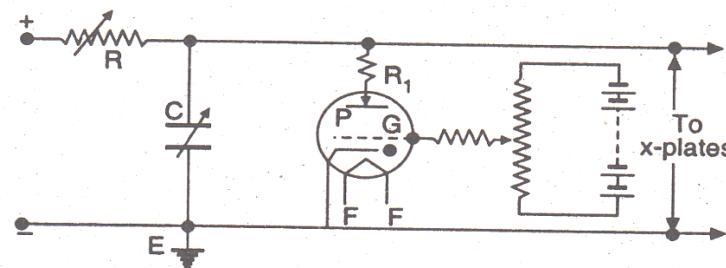
- (iii) *Focus:* The electron beam, leaving the grid as fine collimated beam, passes through a series of electrodes referred to as first, second and third anodes ( $A_1$ ,  $A_2$ ,  $A_3$ ) which are maintained at a positive potential with respect to cathode. These electrodes act like an electron lens system which focuses the electron beam on to the fluorescent screen. Generally, the anodes  $A_1$  and  $A_3$  are connected together and have highest positive potential with respect to the cathode while the anode  $A_2$  has a lower potential than  $A_1$  and  $A_3$  but greater than that of cathode. The potential of the anode  $A_2$  can be varied and thus the spot of light can be brought to focus on the screen. The potentiometer knob controlling the potential of  $A_2$  is labeled as focus on the front panel of the CRO.
- (iv) *Deflection of Electron Beam:* The common form of CRT uses electrostatic deflection arrangement which comprises of two sets of parallel plates: (a) horizontal plates ( $Y_1 Y_2$ ) representing a capacitor which produces electric field in the vertical direction. This electric field deflects the electron beam up and down (b) vertical plates ( $X_1 X_2$ ) produces electric field in horizontal direction. This electric field tends to deflect the electron beam left or right. These deflection plates are mounted between the anode  $A_3$  and the screen. If the electric fields are simultaneously set up between vertical and the horizontal deflection plates, the resultant displacement of the electron will be in accordance with the vector sum of the vertical and horizontal displacements. Hence the spot can be focussed to a predetermined position by applying suitable DC potential to each of the two set of deflection plates. The variable DC potential for this purpose is regulated by control knobs labeled Horizontal Positioning (Hor. Pos.) and

Vertical Positioning (Vert. Pos.). One plate of each pair is connected to the anode  $A_3$  so that there is no electric field between the plates and the electron gun. Finally the anode  $A_3$  is grounded so that the deflector plates are not in the vicinity of the high potentials which would cause the spot to behave in an erratic manner. The cathode and other electrodes are at negative potential w.r.t the earth.

- (v) **Screen:** The screen is coated with fluorescent material such as zinc orthosilicate and a brilliant spot is visible on the screen due to fluorescence. The inside of the neck of the tube is coated with aquadog or some other conducting material thereby connecting the screen to final plate  $A_3$  so that the electron beam is in free flight from anode  $A_3$  to screen with the intervening space free of electric or magnetic fields. As the electrons strike the screen, they leak off to anode  $A_3$  and finally through the power supply to the anode.

- (b) **Time Base Circuit:** The CRO has the ability to measure short time intervals and record variation of voltage with time. If an alternating potential difference is applied to the plates  $Y_1Y_2$ , the spot on the screen will trace out a vertical line proportional to the peak value of the applied voltage. If the waveform of the alternating voltage is to be obtained, the spot must be made to move horizontally along x-direction as well as move up and down along y-axis so that the pattern may spread out on the screen. This is done by connecting the horizontal deflecting plates  $X_1X_2$  to a source of voltage that rises gradually at a constant rate to the maximum value and then suddenly drop back to zero. Such a voltage is said to have saw-tooth shape. It causes the beam to move horizontally across the screen at a uniform speed and then

sends back to its starting point. The electronic circuit producing the saw-tooth voltage is called sweep or time base circuit. The time base is said to be linear if it produces uniform movement of the spot. The sweep or time base circuit helps us to obtain a pattern on the screen which is exactly the same as the curve of varying voltage applied across the vertical deflecting plates  $Y_1 Y_2$  as a function of time. The only condition that has to be fulfilled is that the period of sawtooth voltage must be an exact multiple of the period of applied voltage to be studied. For example, suppose the frequency of applied alternating voltage applied to Y-plates is 50Hz, then if the frequency of sawtooth voltage is 50Hz, then one complete wave of the applied voltage will appear on the screen. However if the frequency of sawtooth voltage is 100Hz, then two complete waves of the applied voltage will appear on the screen. A simple sweep or time base circuit makes use of a gas filled triode thyratron and this circuit is shown in the figure B2.11.



**Figure B2.11:** The circuit diagram of a time base circuit.

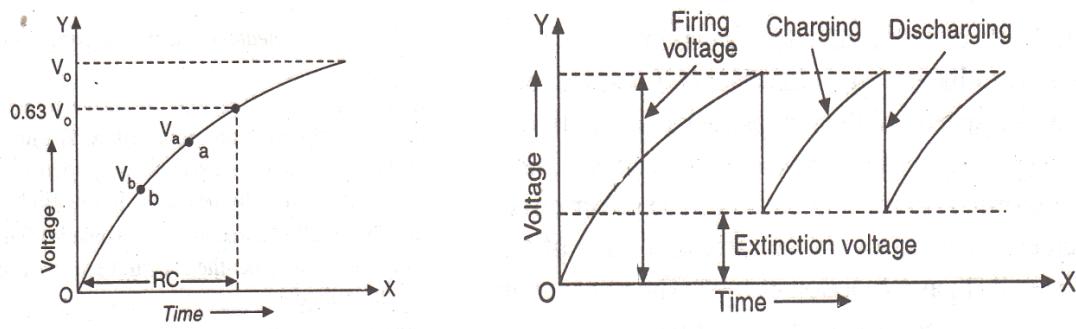
The capacitor is charged through a high resistance  $R$  by connecting it to DC source of constant emf  $V_0$ . With the thyratron connected in parallel to the capacitor, the voltage across the capacitor rises from zero to  $V_c$  according to the equation given as:

$$V_c = V_0(1 - e^{-t/RC}) \quad (B2.10)$$

The curve for the voltage  $V_c$  plotted against time is shown in the figure B2.12. The quantity  $RC$  is the time constant which is defined as the time taken for the voltage across capacitor to grow to 63.2% of the peak value.

When the capacitor voltage is sufficiently high to fire the thyratron, the capacitor discharges through thyratron till its voltage falls to extinction value. At this stage discharge stops and capacitor starts recharging. This process repeats itself and the voltage pattern similar to the edge of saw and hence the name saw tooth is obtained.

If the X-plates are connected across the capacitor, the spot moves uniformly across



**Figure B2.12:** Saw tooth trace generated by charging and discharging of capacitor.

the screen during the period of charging of capacitor and then suddenly flies back to the starting point during the discharge period. The time required for a complete cycle of build up of charge and then discharge of capacitor can be adjusted by changing the time constant  $RC$ . The length of sweep depends upon the value of firing and extinction voltage. In order to study waveforms of different frequencies, the sweep frequency can be suitably adjusted with the help of two controls, which are : (i)

*Range Switch* giving different ranges of frequencies in coarse step by changing the capacitance C and (ii) *Frequency selector* which fine tunes the sweep frequency, within the limits of any one setting of the range switch, by gradual variation of resistance value.

- (c) **Synchronisation:** When two events occur simultaneously, they are said to be synchronized. The input voltage signal applied to Y-plates, if it is periodic and synchronized with the sweep signal, it will appear *stationary or locked in step*. When the beginning of the signal is caused to appear at the beginning of the sweep trace, the wave pattern will remain stationary. However if there is a slight lack of synchronization, the pattern will start moving on the screen. To keep the trace steady without frequent recourse to the fine frequency control, a portion of the input signal is fed to the sawtooth generator and helps to lock the generator in step with the vertical input frequency. The *synch* control varies the signal feed to the sweep oscillator and a slight adjustment of the *synch* control knob helps to freeze the pattern on the screen. When synchronization is achieved, the trace is repeated again and again in the same position and thus appear stationary on the screen.
- (d) **Power Supply:** The power supply provides the high voltage for the cathode ray tube (1-15kV), filament voltages for vacuum tubes and low voltages (1-150V).
- (e) **Amplifiers:** To amplify the weak signals to increase the deflection in the vertical and horizontal directions without distortion of the incoming signals, the X- and Y-plates producing the horizontal and vertical deflections are connected to suitable amplifiers. Regulating knobs provided on the front panel of CRO are respectively labeled as

*horizontal gain* (HOR. GAIN) and *vertical gain* (VERT. GAIN). With the help of these the gain or amplification of the signal may be varied for the horizontal and vertical plates. The horizontal amplifier can also amplify the sweep circuit oscillations and provided with a switch so that it can be connected to the internal sweep circuits or external signal.

# **SECTION C: Quantum and Statistical Physics**

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# C1: FRANK-HERTZ SET UP

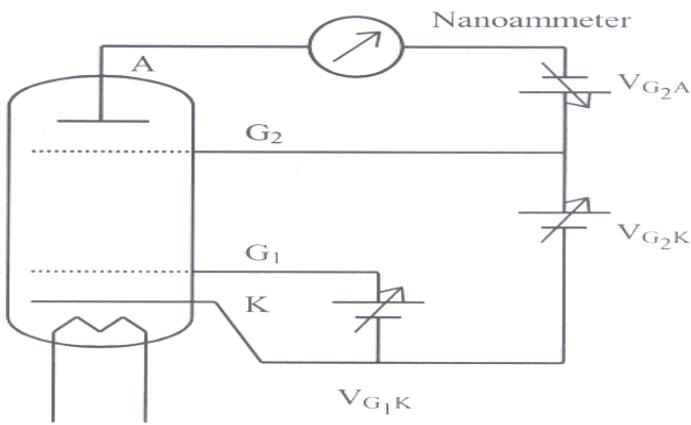
**C1.1: Objective:** To determine quantized excitation energy in argon using Frank-Hertz set-up.

## C1.2: Apparatus (Frank-Hertz Set-Up)

This setup is single table top portable model. The central part of this set-up is a tetrode tube which is filled with Argon gas at low pressure. The remaining circuitry comprises of power supplies capable of providing variable voltages across different electrodes of the tetrode tube. In addition to these variable power supplies, the set up has multi-range ammeter and voltmeter to measure different currents and voltages in the tetrode tube. Further terminals are provided to feed output to CRO generating the trace. The different components of the set-up are described in the subsections to follow:

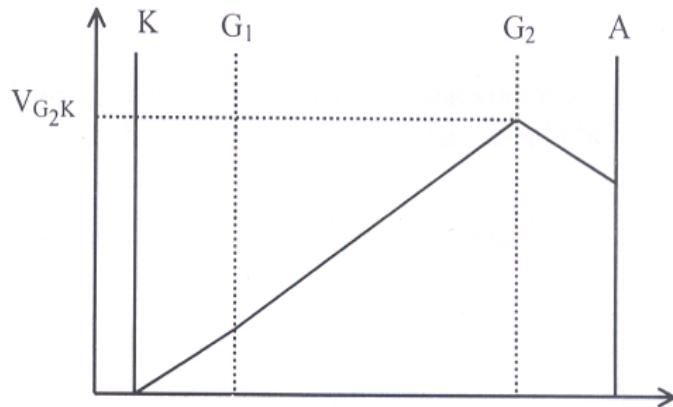
### C1.2.1: Tetrode Tube

The tetrode tube is a vacuum tube comprising of four electrodes and in our case it is filled with argon gas at low pressure. The associated biasing circuit to provide potential difference across different portions of tetrode tube is shown in the figure C1.1.



**Figure C1.1:** The tetrode tube showing the biasing voltages across various electrodes.

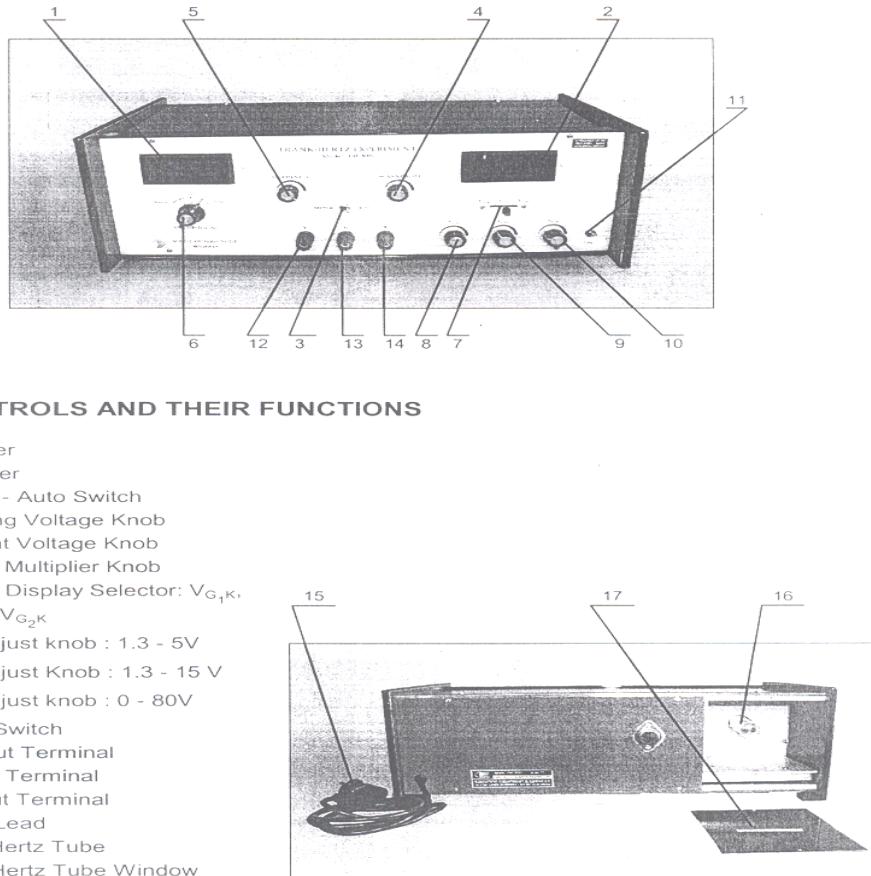
The four electrodes in this tube are (i) cathode which is heated through heat radiated by the filament and is maintained at a negative potential (ii) two grids  $G_1$  and  $G_2$  which are maintained at progressively accelerating voltage with respect to the cathode and (iii) collector or anode which is maintained at a retarding potential with respect to the grid  $G_2$ . The potential distribution among different electrodes is depicted in the figure C1.2.



**Figure C1.2:** The potential distribution among various electrodes of the tetrode tube.

**C1.2.2: Front Panel:** It comprises of various voltage controls, measuring instruments of voltage and current and a pair of terminals for viewing the output voltage on CRO screen (see section B1.3.5. for details of construction and working). The schematic view of the front and rear panels is shown in the figure C1.3.

- (a) Control knob to continuously vary power supply to the filament through variation of voltage in the range 2.6-3.4V.
- (b) Control knob through which the accelerating potential difference between the cathode and the first grid ( $V_{G1K}$ ) can be varied continuously in the range 1.3-5V.
- (c) Control knob through which the accelerating potential difference between cathode and second grid ( $V_{G2K}$ ) can be varied continuously in the range 0-95V.
- (d) Control knob for applying retarding potential difference between anode and second grid ( $V_{G2A}$ ) which can be varied continuously in the range 1.3-12V
- (e) Sawtooth waveform for CRO display having scanning voltage 0-95V and scanning frequency  $115\pm20\text{Hz}$ .
- (f) Multi Range digital voltmeter (Range: 0-100V with 3.5 digit, 7-segment digital display).
- (g) Multirange digital ammeter (Ranges: 0-100 $\mu\text{A}$  ; 0-10  $\mu\text{A}$  ; 0-1  $\mu\text{A}$  with 3.5 digit 7-segment digital display)



**Figure C1.3:** The schematic view of front and rear panel of the Frank-Hertz set-up.

### C1.3: Theory

#### C1.3.1: Basic Objective of Experiment

The early spectroscopic experiments provided convincing evidences that atoms, when excited through some means to excited states, subsequently de-excite to lower states emitting photons of discrete frequency. This was first proved theoretically by Bohr's model of the hydrogen atom and the frequency of radiation  $\nu$  is related to the difference in energy of involved levels through  $\Delta E = h\nu$ . In earlier experiments on spectroscopy of atoms, the excitation of electrons to higher excited states was accomplished by thermal or

electrical means. Further experiments demonstrated that the absorption of radiation by atomic vapors also occurred only for discrete frequencies. The mechanism of energy transfer, involved in excitation of atoms in vapor, is through inelastic scattering of electrons from the entire atom. If the bombarded atom does not get ionized, then almost the entire kinetic energy of the bombarding electron can be transferred to the atomic system. It is then expected that transfer of energy to atomic electrons, irrespective of the mechanism of excitation, should always be in discrete amounts. Frank and Hertz in 1914 performed experiment to verify that it is possible to excite atoms by low energy bombardment and the energy transferred from electrons to the atoms is always discrete in nature.

### **C1.3.2: Working of Frank-Hertz Set-Up**

The cathode is indirectly heated by a filament thereby resulting in emission of electrons through the process of thermionic emission. The electronic current can be controlled through the power supply to the filament for which a control knob is provided on the front panel. The grid  $G_1$  is maintained at slightly higher potential ( $V_{G1K}$ ) than the cathode so that the electrons emitted by the cathode do not result in space charge accumulation and modify the voltage distribution between various electrodes.

If the potential difference  $V_{G1K}$  crosses the threshold for exciting first level of argon atom (in the vapor state), some of the electrons in the beam can undergo inelastic collisions with argon atoms thereby exciting them to the first excited state. In this process, such electrons will lose most of their kinetic energy. To detect the excitation of atoms in the vapor, it is possible to observe, for example, the radiation emitted when the atoms de-excites to the

ground state or the change in absorption of the given spectral line or some other related phenomenon; however the most sensitive technique consists in observing the electron beam itself. If the electrons have been accelerated to a potential just equal to the energy of the first excited state, some of them will excite atoms of the vapor and as a consequence will lose almost all their kinetic energy. Now if some retarding potential exists before collector region, the electrons that have scattered in-elastically will be unable to overcome it and thus will not reach the anode.

The grid-2 ( $G_2$ ) is maintained at potential difference ( $V_{G2K}$ ) so as to accelerate the electron beam in the space between two grids and cause inelastic excitation of argon atoms to the first excited state. However if the accelerating potential  $V_{G2K}$  reaches a value twice that of first excitation potential, it is possible for an electron to excite an Argon atom halfway between two grids, lose all its energy, and then gain anew enough energy to excite a second atom and end with practically zero energy at grid-2. Thus it is not possible to overcome the retarding potential between grid-2 and anode, giving rise to second dip in the current. The advantage of this set-up is that the current dips are pronounced and it is easy to obtain five-fold or even higher multiplicity in the excitation of the first level. However it is impossible to obtain excitations of higher levels.

It is evident that the density of the atomic vapor through which the electron beam passes greatly affects the current. Low densities increase the electron current but dips are shallow. High densities decrease the electron current but dips are more pronounced. Another important point is that in principle the experiment must be performed with monoatomic gas since if the molecular vapors are bombarded, it is possible for the

electrons to transfer energy to molecular energy levels which form almost a continuum. Some of the preferred elements for Frank-Hertz experiment are mercury, neon and argon.

#### C1.4: Procedure

1. Before the power is switched ON make sure that all control knobs are at their minimum position and current multiplier knob is at  $10^{-9}$  position.
2. Turn the Manual-Auto switch to Manual and check that scanning voltage knob is at its minimum position.
3. Set the following parameters by turning different control knobs: Filament voltage = Mid-range position;  $V_{G1K}=1.5V$ ;  $V_{G2A}=7.5V$ ;  $V_{G2K}=0V$ ; Current multiplier= $10^{-9}A$ .
4. Rotate  $V_{G2K}$  knob and observe the variation of plate current with the increase of  $V_{G2K}$ . The current reading would show maxima and minima periodically. The magnitude of maxima could be adjusted by varying filament voltage and current multiplier. Now take the systematic readings,  $V_{G2K}$  vs. plate current. For better resolution the readings must be taken at an interval of 1V. The readings of  $V_{G2K}$  vs. plate current should be taken up to the potential value  $V_{G2K}=80V$ . Plot the graph with plate current on Y-axis and accelerating voltage  $V_{G2K}$  at X-axis.
5. Determine excitation potential and excitation energy of the first excited state of Argon atom by noting the voltage difference between consecutive current minima. These voltage differences yield excitation potentials which are averaged to get the final value. The excitation energy is given by multiplying the excitation potential with charge of electron. In general the excitation potential expressed in electron volt units (eV) is the excitation energy.

## C1.5: Observations

**Table C1.1:** Tabulation of electron current as a function of voltage  $V_{G2K}$ .

$V_{G1K} =$

$V_{G2A} =$

Current Multiplier setting =

S.No.	$V_{G2K}$	Current (nA)	S.No.	$V_{G2K}$	Current (nA)	S.No.	$V_{G2K}$	Current (nA)
1.			26.			51.		
2.			27.			52.		
3.			28.			53.		
4.			29.			54.		
5.			30.			55.		
6.			31.			56.		
7.			32.			57.		
8.			33.			58.		
9.			34.			59.		
10.			35.			60.		
11.			36.			61.		
12.			37.			62.		
13.			38.			63.		
14.			39.			64.		

15.			40.			65.		
16.			41.			66.		
17.			42.			67.		
18.			43.			68.		
19.			44.			69.		
20.			45.			70.		
21.			46.			71.		
22.			47.			72.		
23.			48.			73.		
24,			49.			74.		
25.			50.			75.		

Excitation Potential of Argon atom =

Excitation energy of 1st excited state of Argon atom =

## C1.6: Precautions

1. Before taking the systematic readings, gradually increase the value of  $V_{G2K}$ . Adjust the filament voltage if required such that maxima readings are about 1000 in  $10^{-9}$  range. This will ensure that reading could be taken in the same range.
2. During the experiment (in manual mode), when the voltage is over 60V, be attentive over the output current indicator. If the ammeter reading increases suddenly, decrease the voltage at once to avoid damage of the tube.

3. Whenever the filament voltage is changed, allow 3-4 minutes for its stabilization.
4. To avoid the damage of Frank-Hertz tube, make sure before switching the power ON:

Manual-Auto Switch Position	Manual
Scanning Voltage Knob	Minimum position
Filament Voltage Knob	Minimum position
VG1K, VG2K, VG2A Knobs	Minimum position

### **C1.7: Some related activities**

1. To find the frequency of radiation being emitted by Argon atoms.
2. To check the effect of variation of VG2A on the dips and peaks obtained in the graph.

### **C1.8: Some suggested questions for viva.**

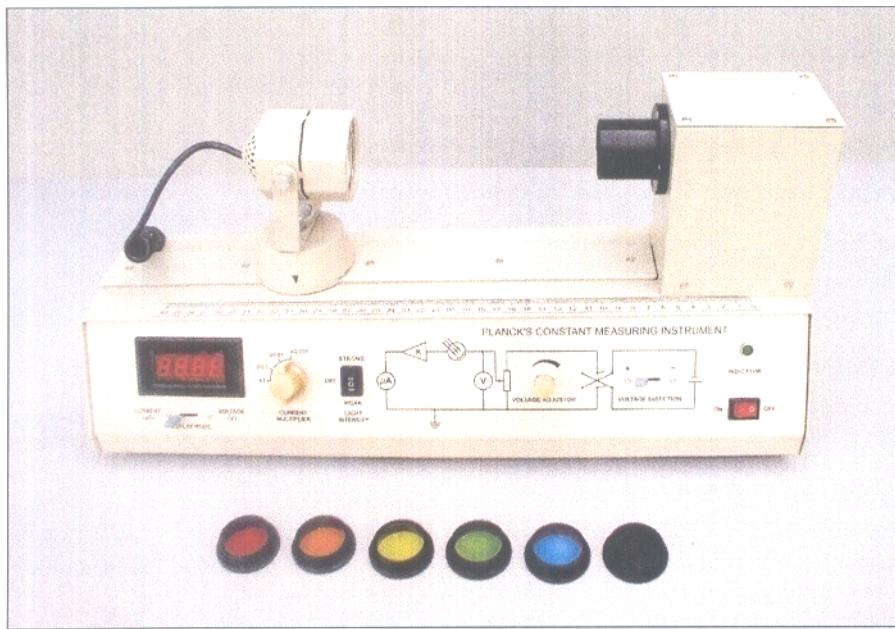
1. What is the significance of Frank Hertz experiment?
2. Why were the 4 electrodes used in the experiment?
3. Describe the behaviour of current recorded in the experiment?
4. Why was Argon gas used for performing the experiment? Can some other gas be used?
5. How is the excitation potential calculated using the dips observed in the VI graph?

## C2: PLANCK'S CONSTANT

**C2.1: Objective:** To find the value of Planck's constant and evaluate the work function of the cathode material by use of photoelectric cell.

### C2.2: Apparatus (Planck's Constant Set-up)

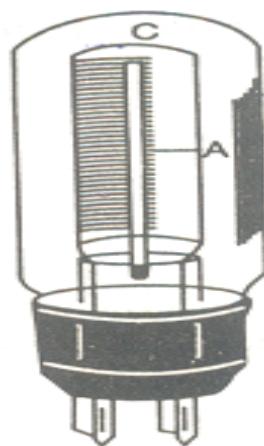
This set-up is a compact portable kit and a table top model. The schematic view of this kit is shown in the figure C2.1. This set-up comprises of the following components packed into it:



**Figure C2.1:** The Schematic view of Planck's constant set-up.

**Photoelectric Cell:** A photoelectric cell is an arrangement to convert light energy into electrical energy. According to their construction, the photoelectric cells are of three types:

(i) Photo-emissive cell (ii) Photo-conductive cell and (iii) Photo-voltaic cell. The photo-emissive cells can be of vacuum type as well as gas filled type. In this experiment the photo-emissive cell of vacuum type is used. It consists of a glass or quartz bulb depending upon its use with visible or ultraviolet light. This bulb encloses a cathode C in the form of a semi-cylindrical plate having large surface area and an anode A in the form of a straight wire. The cross-sectional view of this cell is shown in the figure C2.2.



**Figure C2.2:** The cross-sectional view of photo-emissive cell.

When the cell is to be used with visible light, a sensitive material like sodium, potassium or caesium is deposited on the cathode surface while for use with ultraviolet light cadmium is deposited. To have a large number of electrons per photon of light, composite materials such as caesium on silver oxide ( $\text{Ag-O-Cs}$ ) or antimony caesium alloy ( $\text{Sb-Cs}$ ) or a recently developed combination of bismuth, silver, oxygen and caesium ( $\text{Bi-Ag-O-Cs}$ ) are used as photo-cathode materials. The anode (A) is in the form of a straight wire made of

nickel or platinum. These electrodes are sealed in an evacuated glass or quartz bulb according to whether it is to be used with visible or ultraviolet light.

The current in a vacuum type cell is small which can be increased by filling the bulb with an inert gas like neon or argon at a pressure of few mm Hg. Such a cell is gas filled cell.

The phototube cell is placed at one extreme of an optical bench. It is well concealed to protect it from getting exposed to stray light. A combination of draw-tube and focusing lens fixed at the back is also provided to install color filters and focus the light onto the phototube.

**Light source** is in the form of a halogen tungsten lamp of rating 12V and 35W. This lamp is mounted on an optical bench to place and move the light source so as to adjust the distance between it and the photo-tube.

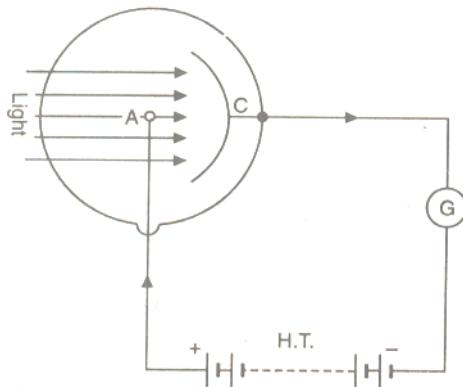
**Regulated Power Supply** applies voltage between two electrodes of the phototube. This voltage is continuously variable in the range up to  $\pm 15V$  through a multi-turn pot and is displayed through a 3.5 digit 7-segment LED.

**Digital nanometer** measures the current through the phototube. It is multi-ranged in which different ranges are selectable through a current multiplier switch to  $1000\mu A$ ,  $100 \mu A$ ,  $10 \mu A$  and  $1 \mu A$ . This nanometer has a resolution of  $1nA$  at  $1 \mu A$  range.

**Color filters** to provide light of wavelengths 460nm, 500nm, 540nm, 570nm and 635nm,

### C2.3: Theory

A simple circuit showing the working of a cell is shown in the figure C2.3 below. A light of suitable wavelength is allowed to fall on the cathode C which emits electrons.



**Figure C2.3:** The working of photocell with retarding voltage.

When the anode is maintained at a positive potential, the electrons are attracted towards it and a conventional current flows in the external circuit. If the anode A is maintained at a negative potential, the electrons are repelled by it and a smaller number of electrons are able to reach anode. The current, therefore, decreases and at a certain negative potential no photo-electron reaches anode. This potential is called stopping or cut-off potential. If  $V_s$  is the stopping potential and  $v_{\max}$  denotes the maximum velocity of emitted photo-electron, then we can write that:

$$eV_s = \frac{1}{2}mv_{\max}^2 \quad v_{\max} = \sqrt{\frac{2eV_s}{m}} \quad (C2.1)$$

The velocity of ejected electrons increases with increase in frequency of incident light. When the frequency is below a certain value called critical or threshold frequency, no electrons are ejected. If  $\nu_o$  is the threshold frequency the energy required to eject an electron just out of cathode surface, called the work function ( $W$ ), is

$$W = h\nu_o \quad (C2.2)$$

In photoelectric effect the whole of the energy of the incident photon is transferred to electrons in the metal. If light of frequency  $\nu > \nu_o$  falls on the photo-sensitive material, a part of the energy is used by the electron to come out of the metal surface and rest is imparted as kinetic energy to the electron.

$$h\nu = \frac{1}{2}mv_{\max}^2 + W = eV_s + h\nu_o$$

$$V_s = \frac{h}{e}\nu - \frac{h}{e}\nu_o \quad (C2.3)$$

Thus a graph between stopping potential and frequency of incident light is a straight line .

The slope of the straight line is given as:

$$\text{slope} = \tan \theta = \frac{h}{e}$$

$$h = e \tan \theta = e \left( \frac{\Delta V_s}{\Delta \nu} \right) \quad (C2.4)$$

The above equation helps in finding the value of Planck's constant. The intercept of the line with y-axis is given as

$$\text{Intercept} = \frac{h}{e}\nu_o = \frac{W}{e}$$

$$W = e(\text{Intercept}) \quad (C2.5)$$

The above equation defines the work function of the photo-sensitive material used in the photo-cathode.

## C2.4: Procedure

1. Insert the red filter ( $\lambda=636\text{nm}$ ), set the light intensity switch at strong light, voltage direction switch at negative and display mode switch at current display.

2. Adjust the de-accelerating voltage to 0 Volts and set the current multiplier at 0.001.
3. Increase the de-accelerating voltage to decrease the photocurrent to zero. Note the value of de-accelerating voltage ( $V_s$ ) corresponding to zero current for red wavelength of 636nm. This de-accelerating voltage is the stopping voltage at 636nm wavelength.
4. Repeat the experiment for other wavelengths corresponding to yellow, green and blue colors using different available filters in the same way.
5. Plot a graph between stopping potential ( $V_s$ ) on y-axis and frequency of light permitted by filter on x-axis. This plot will be a straight line.
6. The slope of this will be given by  $\frac{h}{e}$  which multiplied by charge of electron (e) will give the value of Planck's constant.
7. The intercept with the y-axis of  $V_s$  vs.  $v$  plot will yield work function of material of the photocathode using equation (C2.5).

## C2.5: Observations

**Table C2.1:** Tabulation of variation of stopping potential with frequency of incident light.

S.No.	Filter Colour	Frequency ( $v$ ) ( $10^{14}\text{Hz}$ )	Stopping Potential ( $V_s$ )
1.			
2.			
3.			
4.			

5.			
----	--	--	--

**Result:** Planck's constant =

Work Function =

## C2.6: Error

The percentage error in the deduced value of Planck's constant is given as:

$$\frac{\Delta h}{h} \times 100 = \left[ \frac{h_{th} - h_{exp}}{h_{th}} \right] \times 100$$

## C2.7: Precautions

1. The instrument along with its filters should be operated in dry, cool and indoor space.
2. Phototube should not be exposed to direct light.
3. The instrument should be kept in dustproof and moisture free environment.
4. Clean the filter and lens by absorbent cotton before use.
5. Switch off the power and cover the draw tube after finishing the experiment. This avoids undue ageing of phototube.

## C2.8: Some related activities

1. Verify laws of photoelectric emission.

## C2.9: Some suggested viva questions:

1. What is Photoelectric effect?
2. What is significance of Planck's constant?

3. What is Einstein's equation of Photoelectric effect? How is it used to calculate Planck's constant?
4. How does stopping potential vary with change in intensity and change in frequency of incident light?
5. How was the concept of quantisation proved using Photoelectric effect?
6. What type of Photoelectric cell is used in the experiment?

## C3: PHOTOVOLTAIC CELL

**C3.1: Objective:** To study (a) voltage-current characteristics (b) loading characteristics (c) power-resistance characteristics (d) inverse square law behavior of photocurrent with distance of source of light from the photovoltaic cell.

### C3.2: Apparatus (Solar Cell Set up)

This set is a table top portable model which comprises of a photovoltaic cell with mounting arrangement, lamp (12V, 21W) with mounting arrangement, DC voltmeter (0-1V), DC milliammeter (0-1mA), resistances (0, 100  $\Omega$ , 200  $\Omega$ , 300  $\Omega$ , 400  $\Omega$ , 500  $\Omega$ , 600 $\Omega$  and 1k $\Omega$ , 2k $\Omega$ , 3k $\Omega$  and infinity) selectable by a band switch.

### C3.3: Theory of Photovoltaic Cell

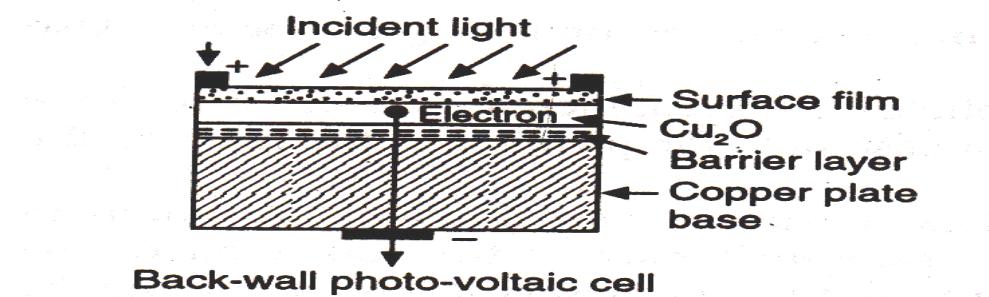
A photo-voltaic cell is the one which generates emf on irradiation of light due to photoelectron emission process. According to their construction, the photoelectric cells are of three types: (a) photo-emissive cell (b) photo-conductive cell (c) photovoltaic cell. The photovoltaic cell also called solar cell is capable of converting solar energy into electrical energy by a process called photo-voltaic conversion. The silicon solar array finds application as potent alternative source of power and has additional virtues in terms of its portability, economy and environmental friendliness. Solar cell arrays or panels are

nowadays being encouraged for usage to meet household power needs and these also find applications in industries, satellites and space-crafts and in many mobile installations.

### C3.3.1: Photovoltaic Cell

The photovoltaic cell is the only true cell which generates e.m.f after being irradiated with sunlight. It requires no external source of emf for its operation. The photovoltaic cell comprises of a thin silicon layer spread uniformly over a metallic surface, also called base of cell, by heat treatment. On the other side of the semiconductor layer, there is a thin translucent layer of silver, gold or platinum (also referred to as surface film) which allows the incident light to fall on the semiconductor layer. The photovoltaic cells exist in two different configurations which are back wall and front wall photovoltaic cell.

The back wall photovoltaic cell (see figure C3.1) has base plate of copper over which a transparent semiconductor layer of cuprous oxide is spread by means of heat treatment



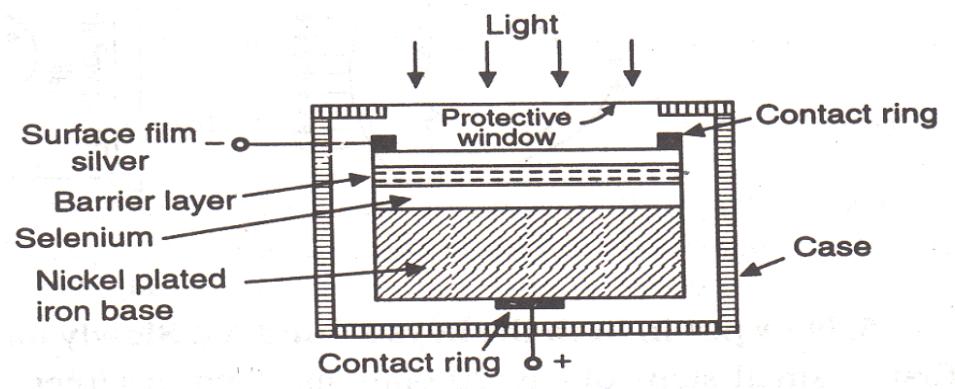
**Figure C3.1:** The schematic view of a back wall photovoltaic cell.

and the surface film is of gold. Here cuprous oxide is a p-type semiconductor with a band gap of 2.1eV. During the preparation of semiconducting oxide layer of Cu<sub>2</sub>O, a very thin insulating layer is formed between the base plate and semiconducting oxide layer, which is

commonly referred to as barrier layer. When the incident radiation illuminates the semiconductor layer of  $\text{Cu}_2\text{O}$ , it causes liberation of electrons, which overcome the barrier layer and flow out of the cell through the back plate. The construction of the back wall photovoltaic cell is as shown in the figure C3.1.

In the case of front wall photovoltaic cell, the barrier layer formed between the translucent metal film and semiconductor layer. This barrier layer is opaque in nature. In this type of cell, the electrons emitted as a result of irradiation of semiconductor layer, pass out through the front surface film instead of base plate.

One of the common and most used type of front wall cell is the selenium cell (see figure C3.2), which has a base plate of nickel plated iron over which a thin semiconductor layer of selenium is coated. The surface film is translucent layer of silver or platinum. A barrier layer is formed because the silver or platinum surface film is deposited through the process of



**Figure C3.2:** The schematic view of a selenium cell.

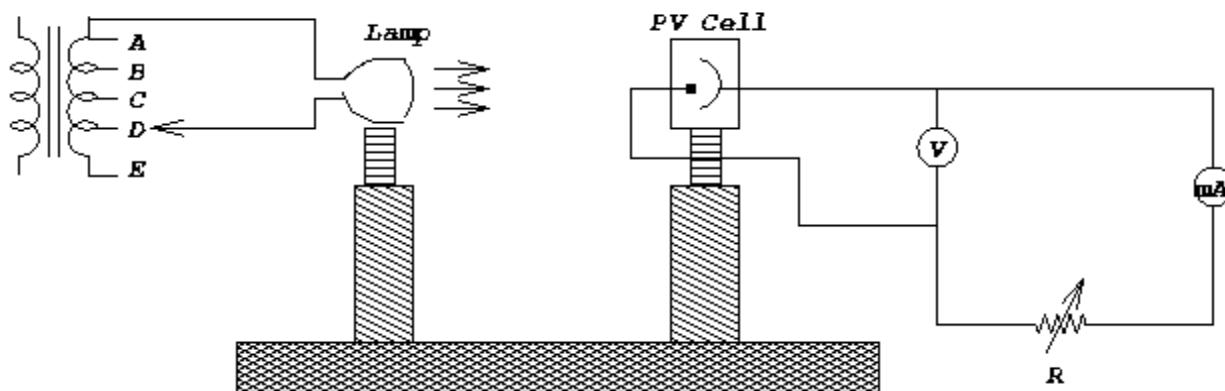
electron sputtering. The electrodes of this cell are copper rings, one of which is in contact with base plate and other the surface film. When the light illuminates the translucent

surface film, it pass through this layer and the barrier layer underneath before reaching the semiconductor layer. The electron emission, so caused in the semiconductor layer, pass through the surface film into the external circuit.

**C3.3.2: PN Junction Solar Cell:** The solar cell is a large PN junction diode having either n- or p-type portion much thinner than the other. This enables the light photon to reach the depletion layer and cause the thermal ion-pair generation. The diode is used in reverse bias mode so as to make the depletion layer thick and any ion-pairs formed by interaction of light in this layer, are swept across the junction which further raises the junction potential. In this way the light energy is converted into junction potential across the reverse biased solar cell. A single solar cell is capable of delivering a potential difference of 0.5V and thus a number of such cells are required to be connected in serial array to obtain the desired voltage. Their parallel combinations are also used when high output currents are desired.

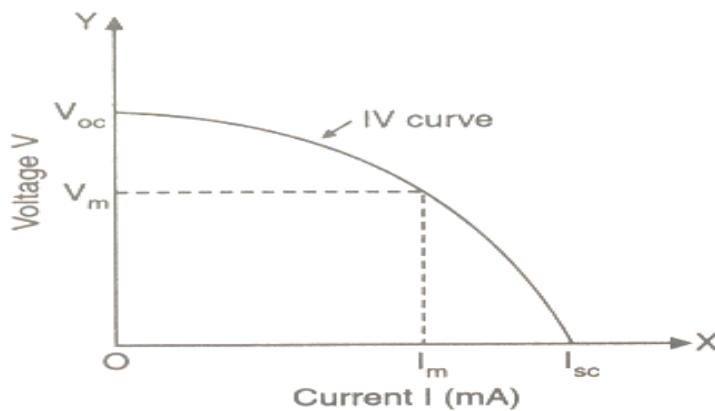
### C3.3.3: Characteristics of a Solar Cell

The circuit diagram for studying various characteristics of photovoltaic cell is shown in the figure C3.3. below.



**Figure C3.3:** The apparatus arrangement and circuit diagram for studying the characteristics of a solar cell.

**VI Characteristics of a Solar Cell:** In the photovoltaic cell the current flowing due to the generated photoelectric emf is proportional to the intensity of the incident radiations only when the electrodes are short circuited (or load resistance is zero). The emf generated is maximum when no current is drawn from the cell (i.e. infinite resistance is introduced in the external circuit) and is referred to as open circuit voltage. As we reduce the external resistance in the circuit, the voltage across the cell goes on decreasing and current increases continuously. The current in the circuit at zero load is maximum (*called short circuit current*). The V-I characteristics of the photovoltaic cell is shown in the figure C3.4 below



**Figure C3.4:** The VI characteristics of a solar cell.

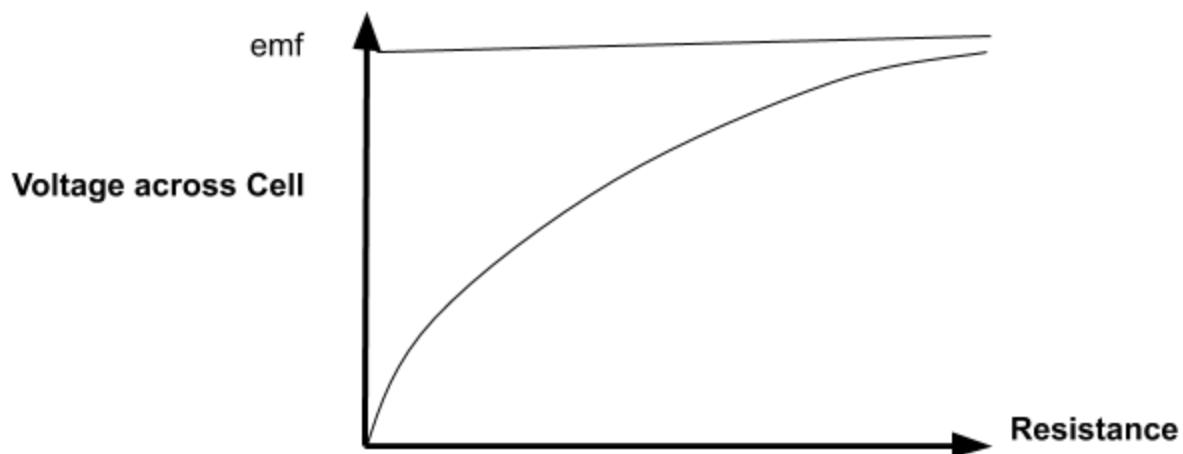
**Fill Factor of Solar Cell:** The ideal power ( $P_i$ ) is defined as the product of short circuit current and open-circuit voltage. The area of the largest rectangle that can be made in the

V-I curve gives the maximum useful power ( $P_{mu}$ ). The ratio of maximum useful power to the ideal power is called the fill factor (f). Thus we have

$$f = \frac{P_{mu}}{P_i} = \frac{V_m I_m}{V_{oc} I_{sc}} \quad (C3.1)$$

The fill factor gives the fraction of incident radiation energy that can be used as electrical energy.

**VR Characteristics of a Solar Cell:** Another characteristic being studied is the variation of voltage developed across a solar cell as a function of external load resistance included in the external circuit. It is observed that as external resistance is increased, the voltage across the solar cell increases and tends to saturate. This is shown in the figure C3.5 and this behavior is similar to an ordinary primary or secondary chemical cell.



**Figure C3.5:** The V-R characteristics of a solar cell.

As the external resistance increases, the current drawn from the cell decreases and the voltage across the cell approaches maximum value called emf of cell.

**Inverse Square Law of Light:** Everyday experience tells us that as we move further away from the light source, the source of light appear to be fainter. This is because of the fact that intensity of light ( $E$ ) falls off in inverse proportion to the square of distance ( $r$ ) of point of observation from the source of light. This is an inverse square law of light intensity and is expressed mathematically as:

$$E = \frac{k}{r^2} \quad (C3.2)$$

$E$  is the measured intensity of light,  $k$  is a constant for a given source and is related to its intrinsic brightness and  $r$  is the distance between the source and point of observation of light intensity. In this experiment, the photo-voltaic cell acts like a light detector. The light intensity is measured in terms of electric current produced in the photo-voltaic cell and is

measured by an ammeter. The graphical plot of  $E$  vs  $\frac{1}{r^2}$  gives a straight line, which verifies the inverse square law of light. This graphical plot tends to deviate from the linear

trend as  $\frac{1}{r^2}$  tends to be large. This is attributed to the fact that as

$\frac{1}{r^2} \rightarrow \text{large} \Rightarrow r \rightarrow \text{small}$  and the lamp, being of finite size, appears to be an aggregate of large number of point sources. Consequently the observed behavior of intensity is the result of superposition of effect produced by each point source and hence deviation from the linear trend is quite obvious.

### C3.4: Procedure

1. Keep the solar cell module in the sunlight for 5-10 minutes to activate it properly.
2. Take the set up to dark room to avoid the effect of stray light falling on the photovoltaic cell
3. Connect the solar cell to the voltmeter, milli-ammeter and a dial of variable resistors in the circuit arrangement as shown in the figure C3.3.

#### 4. For VI, VR and PR Characteristics of Solar Cell:

- (i) Place the tungsten lamp source at a suitable distance from the solar cell so that the voltage developed across the solar cell for the smallest introduced resistance is sufficiently observable.
- (ii) Disconnect the milli-ammeter by removing one of its terminals to obtain the open circuit condition. Note the voltmeter reading which is the open circuit voltage( $V_{oc}$ ).
- (iii) Connect the milliammeter again and make load resistance zero and note the value of short circuit current ( $I_{sc}$ ).
- (iv) Vary the load resistance and note the readings of voltmeter and ammeter. Draw the following graphs: (i) voltage vs current    (ii) voltage vs resistance    (iii) Power vs resistance.

#### 5. Verification of Inverse Square Law

- (i) Fix the position of the solar cell on the optical bench.
- (ii) Place the lamp on the optical bench such that the light incident on the solar cell is normal to its exposed surface.

- (iii) Adjust the brightness of the cell so that the current value is readable even at the greatest distance of separation between the lamp and solar cell.
- (iv) Fix the value of external load resistance by adjustment of variable resistance dial.
- (v) Keeping the cell and source far apart from each other, slowly reduce the distance in steps of 1cm to 2cm between them and note the reading of milli-ammeter at different distances between the source and the cell. The lamp and the solar cell should not be very close to each other as it violates the inverse square law assumptions.
- (vi) Plot I vs  $1/r^2$  graph to verify inverse square law.

### C3.5: Observations

**Table C3.1:** Tabulation for V-I, V-R and P-R characteristics.

Least count of ammeter =

Least count of voltmeter =

Open Circuit voltage =

Short circuit current =

S.No.	LOAD RESISTANCE	VOLTAGE (V) (in Volts)	CURRENT(I) (in mA))	POWER (P) (in mW)
1.				
2.				
3.				
4.				
5.				
6.				
7.				

8.				
9.				
10.				
11				
12				
13				
14				
15				

**Table C3.2:** Verification of Inverse Square law.

S.No .	Distance (r) between lamp and cell (in cm)	Current (in mA)	$\frac{1}{r^2}$ (in $cm^{-2}$ )
1.			
2.			
3.			
4.			
5.			
6.			
7.			
8.			

9.			
10.			

### C3.7: Results and Conclusion

Fill Factor(from the graph) =

### C3.6: Precautions

1. The solar cell should be properly exposed to the sunlight for 5 – 10 minutes.
2. Light should have normal incidence on the solar cell.
3. The current should be kept in safe limits through the load resistance.
4. The solar cell should not be kept in short circuit condition for a long time while noting the short circuit current.
5. Do not keep load resistance to zero when the instrument is not in use.

### C3.7: Suggested viva questions:

1. What is PV cell?
2. How does PV cell convert light energy to electrical energy?
3. How much potential does a single PV cell produces?
4. How does V vs R graph characterises the PV cell under study?
5. What is fill factor? What is its significance?
6. How does a PV cell act a reverse biased pn junction diode without being connected to an external voltage source?

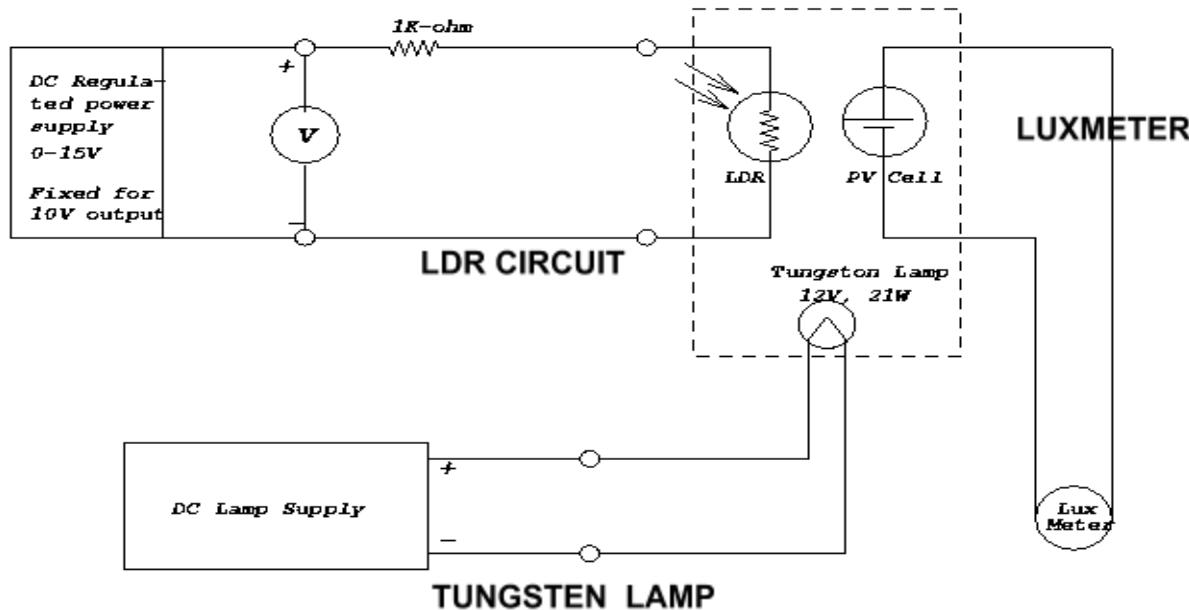
7. How does the band gap of materials used in construction of PV cell affect its efficiency?

## C4: PHOTORESISTOR

**C4.1: Objective:** To determine the response of a photoresistor to the varying intensity of light falling on it.

### C4.2: Apparatus (Photoresistor Set-up)

This set-up is a compact kit which is a portable table top model. The circuit including different components is shown in the figure C4.1. This set-up is comprised of the following components:



**Figure C4.1:** Circuit diagram for studying the characteristics of LDR.

DC regulated power supply fixed to 10V output, voltmeter, LDR, Tungsten lamp (12V, 21W), ammeter, Photovoltaic cell and lux-meter.

**LDR Circuit:** A fixed DC supply (0-15V) furnishes the applied potential to LDR and is usually fixed around 10V. The resistor of  $1\text{k}\Omega$  is connected in series to limit the current, which is measured by micro-ammeter. This circuit consists of a photoresistor or Light Dependent Resistor (LDR) which is an opto-electronic device. The LDR used in this experiment is CdS semiconductor whose resistance falls with increasing intensity of light influx. The variation of resistance of LDR is owing to greater thermal ion-pair generation caused by incidence of light flux.

**Lux meter:** A photovoltaic cell, in short circuit mode, is also fitted in close vicinity of LDR. The photocurrent generated in the short circuit mode, is proportional to the intensity of incident light flux. This current is passed through an operational amplifier based current to voltage converter, which deflects an analog meter calibrated in terms of light intensity. The lux meter is capable of measuring light intensity from close to zero to ~few klux.

**Tungsten Lamp:** This tungsten lamp of power rating of 12V and 21W is used to illuminate the LDR as well as photovoltaic cell. As can be seen from the figure C4.1, the lamp is placed in such a manner that it is equidistant from LDR as well as photovoltaic cell so that light flux falling on both is same and measurement of incident flux can be made through the photovoltaic cell. The voltage applied across the lamp can be varied to change the light intensity. The arrangement of tungsten lamp, LDR and photovoltaic cell are placed in light tight enclosure to avoid any exposure to stray light.

### C4.3: Theory

Semiconductors often have the ability to respond to various spectral regions of electromagnetic radiation. Silicon, Germanium, Gallium-Arsenide and Cadmium Sulphide

are materials which exhibit opto-electronic effects, thereby implying that their electrical properties are responsive to light. The conductivity of such materials increase as they are exposed to the increasing intensity of radiation. This is owing to the fact that valence electrons in these materials are excited to the conduction band on absorption of incident photons from radiation flux. The conductivity ( $\sigma$ ) of material is proportional to the concentration of charge carriers and is given as

$$\sigma = [n\mu_e + p\mu_h]e \quad (C4.1)$$

where  $n$  and  $p$  are free electron and hole densities,  $\mu_e$  and  $\mu_h$  are electron and hole mobilities. Radiant energy causes the dissociation of covalent bonds thereby leading to thermal ion-pair generation (i.e. electron-hole pairs), which results in increased free charge carriers and hence leading to fall in the overall resistance of material. This is the basic principle on which photo-resistor works. Such devices, working on this principle, are called photoconductive cells or Light Dependent Resistors (LDR). The carriers generated by photo-excitation move under the influence of applied electric field and if they survive electron-hole recombination and reach ohmic contacts, they constitute device current. The resistance of LDR constantly falls with the intensity of incident radiations while the current increases non-linearly with the intensity of incident radiation. Small amount of current flows in the LDR circuit even when no light is incident on it and is referred to as Dark Current. The origin of this current is owing to the thermal ion-pair always present in a semiconductor at room temperature. As the intensity of incident light is increased continuously, the resistance of the LDR decreases and approaches a low saturation value. This low saturation value of resistance is related to high saturation current flowing through

the LDR resulting from maximum possible thermal ion-pair generation caused by incident flux of radiation.

The response of a photoresistor with varying light intensity is non-linear and given by:

$$R_{LDR} \propto R_{dark} L^{-b} \quad (C4.2)$$

Where  $R_{LDR}$  is the resistance at light intensity  $L$  at the surface of LDR,  $R_{dark}$  is the LDR resistance without light and  $b$  is the material constant, often called spectral sensitivity of the photo-resistor and is a function of the wavelength of the incident radiation.

Cadmium sulphide (CdS) cells rely on the material's ability to vary its resistance according to the influx of light on the cell. Although not accurate, even a simple CdS cell can have a wide range of resistance from less than  $100 \Omega$  in bright light to in excess of  $10 M\Omega$  in darkness. Many commercially available CdS cells have a peak sensitivity in the region of 500nm - 600nm. The cells are also capable of responding to a broad range of frequencies, including [infrared \(IR\)](#), [visible light](#), and [ultraviolet \(UV\)](#). They find their application as automatic on/off switches in street light, sensors in heat seeking missiles and as sensors in many consumer items such as [camera light meters](#), clock radios, security alarms, [street lights](#) and outdoor clocks. They are also used in some dynamic compressors to control gain reduction. At the other end of the scale, [Ge:Cu](#) photoconductors are among the best far-infrared detectors available, and are used for [infrared astronomy](#) and [infrared spectroscopy](#).

#### C4.4: Procedure

1. Keep the lamp supply control to minimum. Switch ON the power. Note that the DC supply for 10V. Select  $\mu A$  range for LDR ammeter. If there is no deflection in

ammeter, it implies LDR exhibits high resistance in the dark. Otherwise note the value of dark current.

2. Select lowest step for lamp intensity. Increase the intensity by means of lamp voltage in steps of 0.2klux and note the corresponding flowing through the LDR current for a current deflection with readable intensity. Change the ammeter range accordingly.
3. Keep on changing the lamp voltage level to higher steps and further note the variation of LDR current with intensity.
4. Plot the current vs light intensity curve from the observations. The resistance of LDR is calculated by applying Ohm's law  $R = V/I$ , where  $I$  is LDR current and  $V$  is the applied DC potential. Keep in view that current limiting  $1\text{k}\Omega$  resistance must be subtracted from these results.
5. Also plot resistance vs light intensity curve.

#### C4.5: Observations

Least count of the voltmeter =

Least count of Lux-meter =

Least count of the ammeter =

Voltage applied across LDR =  $V =$

**Table C4.1:** Tabulation of current flowing in LDR at different intensities of incident light.

S.No.	Intensity (in klux)	Current in LDR (I) (in mA)	Total Resistance = $R' = \frac{V}{I}$ (in $\text{k}\Omega$ )	Resistance of LDR $R = \left(\frac{V}{I} - 1\right)$ (in $\text{k}\Omega$ )
1.				
2.				

3.				
4.				
5.				
6.				
7.				
8.				
9.				
10.				
11.				
12.				
13.				
14.				
15.				
16.				
17.				
18.				
19.				
20.				
21.				
22.				
23.				
24.				

25.				
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#### C4.6: Precautions and Sources of Error

1. A large resistance of  $1\text{k}\Omega$  should be connected with LDR to limit the current through it, thereby avoiding the risk of damage.
2. Do not keep the set up ON for long time as it heats the LDR circuit.
3. Equal amount of light of light must fall on LDR and photovoltaic cell so as to accurately measure incident light intensity.
4. Current at dark should be measured carefully and in perfect dark conditions.

#### C4.7: Some related activities.

1. To check the spectral sensitivity of the Photoresistor.

#### C4.8: Some suggested viva questions:

1. What type of material is used for construction of LDR? What is LDR used in experiment made of?
2. Why is PV cell used in LDR circuit?
3. What is the use of Photoresistors in electronics?
4. How does an LDR work?
5. Why does the current in the circuit becomes almost constant after certain value of light intensity?
6. What do you understand by dark current?

## **C5: Rutherford scattering experiment through a mechanical analogue.**

**C5.1: Aim:** To study Rutherford scattering experiment through a mechanical analogue to find dependence of scattering angle ‘θ’ on the kinetic energy and impact parameter ‘b’ of the incident beam of particles.

**C5.2: Apparatus:** Mechanical analogue set up of Rutherford’s experiment, electromagnet, dry cell or battery eliminator, connecting wires, tapping key, scale and steel balls.

### **C5.3: Theory:**

Rutherford scattering is the elastic scattering of charged particles by the Coulomb interaction. It is a physical phenomenon explained by Ernest Rutherford in 1911 that led to the development of the planetary Rutherford model of the atom and eventually the Bohr model. The experiment was performed by Geiger and Marsden where they targeted a very thin gold foil( thickness  $2.1 \times 10^{-7}$  m) with a stream of α-particles. The scattered particles were detected by a movable detector coated with ZnS. The results of the experiment lead to Rutherford’s nuclear model of atom.

Rutherford calculated the following relation between the impact parameter and the angle of scattering:

$$b = Z e^2 \cot(\theta/2) / 4 \pi \epsilon_0 (1/2 m v^2) \quad (\text{C5.1})$$

$$\text{or } b \propto \cot(\theta/2)$$

(C5.2)

In mechanical analogue

$$\frac{1}{2} m v^2 = mgh = Z e^2 \cot(\theta/2) / 4 \pi \epsilon_0 b$$

(C5.3)

## Mechanical Model

Alpha particle is very light, as compared to atom of metal but it is about 1000 times heavier than the electron. Thus electrons surrounding an atomic nucleus are pushed to either side as alpha particles goes speeding through and they have little effect upon the shape of trajectory.

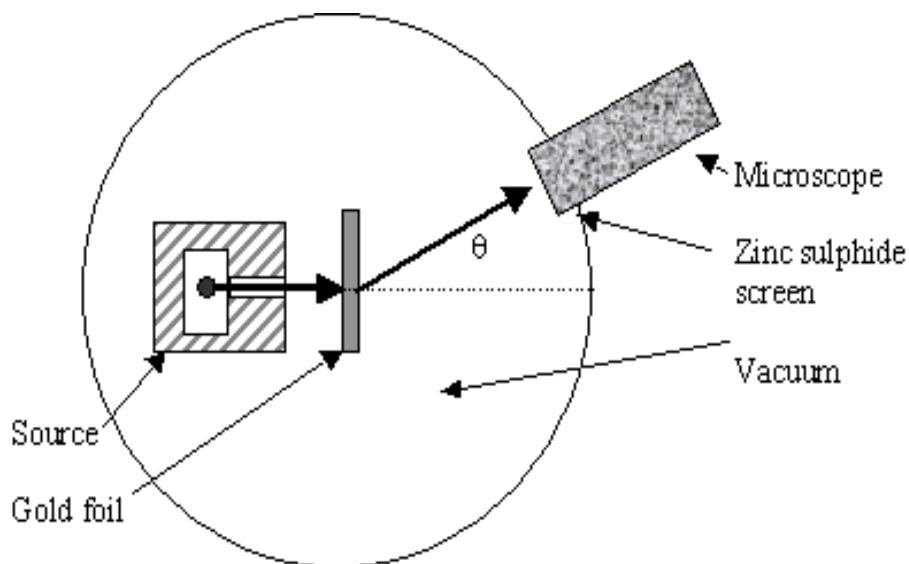


Figure C5.1: Rutherford scattering experiment setup

A mechanical model for demonstrating Rutherford scattering is derived from this equation.

$$\text{Potential Energy (U)} = (K q q') / r \quad (\text{C5.1})$$

The circulation peak represents the nucleus of an atom. Marbles represented by alpha particles roll down when they approach potential hill.

The electrostatic potential energy of alpha particles close to nucleus is analogous to gravitational potential energy of marbles along the hillside and electrostatic force of repulsion is analogous to component of the downward pull of gravity.

#### **C5.4: Assumptions:**

1. The initial and final kinetic energies of the alpha particles are approximately equal.

When an object strikes a massive object, the recoil speed of target is insignificant.

2. The target is considered to be thin so that only single scattering occurs.
3. The bombarding particle and target may be treated as a point mass and charges.

#### **C5.5: Procedure**

1. Connect the electromagnet to dry cell through tapping key.
2. Make arrangements such that ground guide for ball and sunmica board are at the same level.
3. Place the grooved guide at same position defining 'b'.
4. Fix the electromagnet at the some height h.
5. Hold the balls close to electromagnet pressing the key and note the angle of scattering.
6. Note the angles of scattering by varying the impact parameter 'b'.
7. Plot the function  $\cot(\theta/2)$  with respect to impact parameter.

8. Now fix the ground guide at some position corresponding to a fixed value of impact parameter 'b'.
9. Record the angles of scattering by varying the height of electromagnet.
10. Plot the function  $\cot(\theta/2)$  with respect to height 'h'.

### C5.6: Observations

Height 'h' =

**Table C5.1: Table for scattering angles for various values of impact parameter.**

S.No.	Impact Parameter 'b'(in cm)	Scattering angle		Mean $\theta$	Cot ( $\theta/2$ )
		$\theta_1$	$\theta_2$		
1					
2					
3					
4					
5					
6					

Impact parameter 'b' =

**Table C5.2: Table for scattering angles for different heights**

S.No.	Height 'h'(in cm)	Scattering angle		Mean $\theta$	Cot ( $\theta/2$ )
		$\theta_1$	$\theta_2$		
1					
2					
3					
4					
5					
6					

### C5.7: Precautions

1. The zero mark should be in line with the center of the nucleus.
2. Use steel balls only, as iron balls get magnetized very easily.
3. Grooved guide should be in level of sunmica, steel balls should have diameters less than 10 mm.

### C5.8: Sources of Error

1. The shape of the bump may not be perfect.
2. The marks of angles on board may not be accurate.

**C5.8: Suggested Viva questions:**

1. What were the observations made in Rutherford's alpha scattering experiment?
2. What conclusions were drawn from alpha scattering experiment?
3. What is impact parameter?
4. How does the scattering of steel balls vary with variation in the impact parameter?

How can it be explained conceptually?

## C6: Probability Theory and Statistics

**C6.1: Objective:** To understand the concept of probability using a set of two dice and calculate statistical parameters.

**C6.2: Apparatus:** A set of two dice.

**C6.3: Theory:**

Probability is an important mathematical concept which lie at the heart of Quantum Physics. The concept of probability is used when the outcome of an experiment is not certain. For example, it is not possible to tell with certainty whether we will get a head or a tail as an outcome on tossing a coin. Similarly the outcome of throwing of a dice cannot be predicted with certainty. In the context of Quantum Physics, every measurement has an uncertainty involved in it. According to Heisenberg uncertainty principle, position-momentum, energy-time and angular momentum-angle measurements have a minimum value of uncertainty in simultaneous precision measurements. Probability of an event can be defined as:

Probability of an event = Number of ways that event can occur / Total number of possible outcomes (C6.1)

With this formula, the probability of getting a head on tossing a coin is  $1/2$  and that of getting a six in single dice throw is  $1/6$ .

If a pair of dice is thrown, there are 36 possibilities to get total sum between 2 and 12. Various possible outcomes and their expected probabilities are listed in the Table C.6.1. If

an experiment is performed only once concept of probability do not make sense. For example, if a coin is tossed up and we get head, probability of head is 1 but that of tail is 0. Thus it is important to repeat the experiment a large number of times to get the sensible results according to the respective probabilities.

When we have a set of data, it is useful to calculate statistics parameters which are useful in interpreting the results. The parameters used for univariate data can be classified in two categories:

Measures of central tendency: A measure of central tendency is a single value that attempts to describe a set of data by identifying the central position within that set of data. The mean, median and mode are all valid measures of central tendency, but under different conditions, some measures of central tendency become more appropriate to use than others. In our experiment we are going to evaluate mean of the probability distribution.

Mean ( $\mu$ ): It gives the average value of the data and mathematical defined as:

$$\mu = \frac{\sum f_i x_i}{N} \quad (\text{C6.2})$$

Where  $f_i$  is the frequency of the  $i^{\text{th}}$  combination and  $x_i$  is the  $i^{\text{th}}$  combination.  $N$  is number of experiments performed.

Measures of Dispersion: In statistics, dispersion is the extent to which a distribution is spread or squeezed with respect to the central value. Variance and Standard deviation are

measure of dispersion of a data set. In our experiment, we will be evaluating the standard deviation.

**Standard Deviation ( $\sigma$ ):** Standard deviation is the measure of spread of data when the mean is used to calculate central tendency. Thus, it measures spread around the mean.

$$\sigma = \sqrt{\overline{\sum_{i=1}^N \frac{F_i(x_i - \bar{x})^2}{N}}} \quad (\text{C6.3})$$

Standard deviation is also useful when comparing the spread of two separate data sets that have approximately the same mean. The data set with the smaller standard deviation has a narrower spread of measurements around the mean and therefore usually has standard deviation is low has a better chance of being close to the mean than an item from a data set whose standard deviation is higher.

### Comparison of Bivariate distribution

**Chi-square statistics ( $\chi^2$ ):** The Chi-square test is intended to test how likely it is that an observed distribution is due to chance. It is also called a "goodness of fit" statistic, because it measures how well the observed distribution of data fits with the distribution that is expected if the variables are independent. It is mathematically represented as

$$\chi^2\text{-statistic} = \sum \frac{(O_i - E_i)^2}{E_i} \quad (\text{C6.4})$$

Where 'O<sub>i</sub>' is the observed frequency and 'E<sub>i</sub>' represents the expected frequency.

**Table C6.1: Summary of various combinations and their probabilities**

Sum/Combination	Possible combinations	Expected Probability
2	(1,1)	1/36
3	(1,2) (2,1)	2/36
4	(1,3) (3,1) (2,2)	3/36
5	(2,3) (3,2) (1,4) (4,1)	4/36
6	(2,4) (4,2) (1,5) (5,1) (3,3)	5/36
7	(1,6) (6,1) (4,3) (3,4) (2,5) (5,2)	6/36
8	(2,6) (6,2) (3,5) (5,3) (4,4)	5/36
9	(4,5) (5,4) (3,6) (6,3)	4/36
10	(5,5) (4,6) (6,4)	3/36
11	(5,6) (6,5)	2/36
12	(6,6)	1/36

**C6.4: Procedure:**

1. Take a pair of identical dice with six faces numbered from 1-6.
2. Throw dice and note down the sum of numbers appearing on the upper face of the dice.
3. You have to calculate the frequency of getting all possible value of sum from 2 to 12.

4. Repeat the throw many times until you get frequency valued at least 15 for all combinations.
  
5. Plot the observed and expected frequency as a function of the possible combination values.
  
6. Calculate Mean, Standard Deviation and  $\chi^2$ -statistic for the data table produced.
  
7. Calculate the observed and expected probability for getting combination giving even sum. Also calculate  $\chi^2$ - statistic
  
8. Calculate the observed and expected probability for getting combination giving sum between 8 and 11. Also calculate  $\chi^2$ -statistic.

## C6.5: Observations and Calculations

Total No. of trials =

**Table C6.2: Experimental observations**

Sr. No.	Sum/Combinati on $x_i$	Expected Frequency $(F_i)$	Observed Frequency $(O_i)$	$\frac{O_i(x_i - \mu)^2}{N}$	$\frac{(O_i - F_i)^2}{F_i}$
1	2				
2	3				
3	4				

4	5				
5	6				
6	7				
7	8				
8	9				
9	10				
10	11				
11	12				

Draw a plot between possible combinations and the observed and expected frequencies.

Mean ( $\mu$ ) = .....

Standard Deviation ( $\sigma$ ) = .....

$\chi^2$  - statistic = .....

- 1) Calculate the observed and expected probability for getting combination giving even sum. Also calculate  $\chi^2$  - statistic.

Observed Probability = .....  $\pm$  .....

Expected Probability = .....

$\chi^2$  - statistic = .....

2) Calculate the observed and expected probability for getting combination giving sum between 8 and 11. Also calculate  $\chi^2$ -statistic.

Observed Probability = ..... ± .....

Expected Probability = .....

$\chi^2$  - statistic = .....

### C6.6: Precautions

1. This is a random probability experiment, therefore do not try to manipulate frequencies.
2. By keeping the number of throws to a large value, the observed values will approach the expected results.

## C7: Study of probability Theory using coins

**C7.1: Objective:** To verify certain laws of probability distributions (ii) To verify laws of probability by throwing one, two and 10 coins.

**C7.2: Apparatus:** A few identical coins of same denomination, a plastic smooth mug, or a container or tumbler and a wooden tray.

**C7.3: Theory:** Probability: Suppose we toss a coin. Either the head can come up or the tail, i.e., the event can happen in a total number of two equally likely ways. The number of ways in which the head can come up is only one. Therefore, the probability that the head may come up is  $\frac{1}{2}$ . Similarly the probability that the tail may come up is also  $\frac{1}{2}$ .

Let us consider another example of a dice. A dice is a homogeneous cube with six faces marked respectively with number of dots from 1 to 6. It is supposed that the dice cannot fall on its edges. When the dice is thrown it falls with one of its faces upward. The face can have any number of dots from 1 to 6. Thus there are six equally likely outcomes in a single throw of dice. In other words, the probability of any one face (say with number 2) to come up is  $1/6$ . The probability of dice coming up with an even number is  $3/6 = \frac{1}{2}$  as there are three faces with an even number of dots.

From the above examples we find that in the case of throwing a coin the event can happen in two mutually exclusive and equally likely ways, out of which only one is favorable to the head coming up. In the case of throwing a dice, the event can happen in six mutually exclusive and equally likely ways, out of which only one is favorable to a particular case(say with number 2)coming up and the number of ways favorable to even numbered face coming up is 3. Therefore, probability of obtaining a favorable outcome can be defined as follows:

If an event A can occur in a total number of N exhaustive, mutually exclusive and equally likely ways and if a certain number 'm' of these are considered favorable to the event A, the ratio  $m/N$  is called its probability of success.

$$\text{Probability of an event } P = \frac{\text{No. of ways which favor the event}}{\text{Total No. of equally likely possible ways}} = \frac{m}{N} \quad (\text{C7.1})$$

Cumulative Probability: The probability that a variable is less than or equal to some value. This is called the cumulative probability because to find the answer, we simply add probabilities for all values qualifying as "less than or equal" to the specified value.

Example: Suppose we want to know the probability that the number of heads in four flips is 1 or less. The qualifying values are 0 and 1, so we add probabilities for those two possibilities.

$$P(\text{number of heads} = 2) = P(\text{number of heads} = 0) + P(\text{number of heads} = 1) = \\ (1/16) + (4/16) = 5/16.$$

The cumulative distribution is a listing of all possible values along with the cumulative probability for each value.

Zero probability: In case of throwing of a dice, suppose we want the probability of dice coming up with a face marked with number 8. As the dice has only 6 faces marked 1 to 6, the probability of such an event is 0.

Unit probability: Suppose in case of dice, we want the probability of dice coming up with any number less than 7. Then the probability of such an event is 1.

Total probability: If an event can occur in m ways and fail to happen in n ways, then the probability of happening of event is  $m/(m+n)$  and the probability of not happening of the event is  $n/(m+n)$ . As the event must either happen or not happen, the sum of two probabilities =  $m/(m+n) + n/(m+n) = 1$ .

#### Mean or Expected value of a Discrete variable:

The phrase expected value is a synonym for mean value in the long run (meaning for many repeats or a large sample size). For a discrete random variable, the calculation is Sum of (value  $\times$  probability) where we sum over all values (after separately calculating value  $\times$  probability for each value), expressed as:

$$\mu = E(x) = \sum_{i=1}^N x_i p_i \text{ meaning we take each observed 'x' value and multiply it by its}$$

respective probability. We then add these products to reach our expected value labeled  $E(x)$ .

Example: Suppose we want to know the mean of number of heads that turns up in four flips.

Heads	0	1	2	3	4
Probability	1/16	4/16	6/16	4/16	1/16

$$\text{Mean value of number of heads} = 0 \times 1/16 + 1 \times 4/16 + 2 \times 6/16 + 3 \times 4/16 + 4 \times 1/16 = 2.$$

Standard deviation: Knowing the expected value is not the only important characteristic one may want to know about a set of discrete numbers: one may also need to know the spread, or variability, of the data.

To calculate the standard deviation we first calculate the variance. From the variance, we take the square root and this provides us the standard deviation. Variance is given by

$$\sigma^2 = \sum_{i=1}^N (x_i - \mu)^2 \times p_i \quad (\text{C7.2})$$

and the formula for standard deviation is given by :

$$\sigma = \sqrt{\sum_{i=1}^N (x_i - \mu)^2 \times p_i} \quad (\text{C7.3})$$

In this expression we substitute our expression for  $\mu$ . However, an easier formula to use and remember for calculating the standard deviation is the following:

$$\sigma = \sqrt{\sum_{i=1}^N x_i^2 p_i} - \mu \quad (\text{C7.4})$$

Notice in the summation part of this equation that we only square each observed 'x' value and not the respective probability.

Principle of equal a priori probability: When we toss a coin we are clear in our mind that the coin will fall either with its head up or tail up. Similarly if we throw a dice, we are sure that the dice will fall with one of its six faces upwards. In the above cases, we assume an equal probability of all possible events in case the coin and dice are not biased.

This principle of assuming equal probability for events which are equally likely is known as principle of equal a priori probability.

(ii) The coins are similar or indistinguishable.

(I) Coins distinguishable: Two coins can be distinguishable means one coin is of copper and other of silver. Let  $a_1$  and  $b_1$  represent the head and tail respectively of the first coin and  $a_2$  and  $b_2$  represent the head and tail of second coin respectively.

No.	Event	Symbol	Probability
1.	Heads of two face up	$a_1a_2$	$\frac{1}{4}$
2.	Head of 1 <sup>st</sup> and tail of 2 <sup>nd</sup> up	$a_1b_2$	$\frac{1}{4}$
3.	Tail of 1 <sup>st</sup> and head of 2 <sup>nd</sup> up	$b_1a_2$	$\frac{1}{4}$
4.	Tails of two face up	$b_1b_2$	$\frac{1}{4}$

Thus the total number of ways in which the event is possible  $n = 4 = 2^2$

Number of favorable ways for any one of the event  $m = 1$ .

Probability of any one of the event to happen  $p = m/n = 1/4$

(ii) Coins indistinguishable: When the coins are similar

$$a_1 = a_2 = a \quad b_1 = b_2 = b$$

In this case the second and third event become identical and we are left with only three cases.

Now the probability of getting the head of one up and tail of second up is  $2 \times (1/4) =$

1/2. In other words the weight of this event is 2. This can be shown in tabular form as under.

No.	Event	Symbol	Weight	Probability

1	Heads of both coins up	aa	1	$\frac{1}{4}$
2	Head of one and tail of other up	ab	2	$\frac{1}{2}$
3	Tails of two face up	bb	1	$\frac{1}{4}$

Macrostate: Suppose we have 4 distinguishable particles and we wish to distribute them into two exactly similar compartments in an open box. Since the particles are distinguishable we denote them as a, b, c and d. Any particle thrown into the box must fall into one of the compartments. Since both the compartments are exactly alike, the particles have the same a priori probability of going into either of them. In other words, the probability of a particle falling into any compartment is  $\frac{1}{2}$ . The possible ways in which 4 particles can be distributed in two compartments is as under.

Compartment	Number of particles				
	Dist 1	Dist 2	Dist 3	Dist 4	Dist 5
1	0	1	2	3	4
2	4	3	2	1	0

That is there are five different distributions. Each compartment wise distribution of a system of particles is known as a macrostate.

For a system of  $n$  particles distributed among two compartments the various macrostates are  $(0,n)$ ,  $(1, n-1)$ ,  $(2, n-2)$ , ..... $(n-1,1),(n,0)$ . Thus the total number of macro states of  $n$  particles is  $(n+1)$ .

Micro-state: In the above example, since the particles are distinguishable each compartment wise distribution can have a number of different arrangements as shown below.

Macro state	Possible arrangement		No. of microstates
	Compartment 1	Compartment 2	
0,4	0	abcd	1
1,3	a	bcd	4
	b	acd	
	c	abd	
	d	abc	
2,2	ab	cd	6
	ac	bd	
	ad	bc	
	bc	ad	
	bd	ac	
	cd	ab	
3,1	bcd	a	4
	acd	b	

	abd abc	c d	
4,0	abcd	0	1

Each distinct arrangement is known as the microstate of the system.

The distribution (0,4) and (4, 0) can have only one arrangement, distributions (1,3) and (3, 1) can have 4 arrangement each and distribution (2,2) can have 6 arrangement respectively. In the above case of 4 particles divided in 2 compartments, the total number of microstates =  $2^4 = 16$

Total number of microstates for n particles to be divided in 2 compartments =  $2^n$ .

Total number of microstates for n particles to be divided in N compartments =  $N^n$

The number of microstates corresponding to a given macrostate is known as its frequency. The postulate of equal a priori probability may also be stated as: All accessible microstates corresponding to a possible macrostates are equally probable.

According to this postulate the probability of occurrence of a given macrostate is proportional to the number of microstates that correspond to the given macrostate.

Probability of distribution of n identical particles in two compartments.

If there are n identical particles to be distributed in two compartments, then the total number of ways in which the event is possible or the total number of microstates is  $2^n$ . If it is proposed to have r particles in one compartment and the remaining (n-r) in the other compartment, then the number of meaningful arrangements of the number of microstates in the macrostate (r, n-r) are

$${}^nC_r = \frac{n!}{r!(n-r)!} \quad (\text{C7.5})$$

Therefore, the probability of macrostate  $(r, n-r)$  is given by

$P(r) = \text{Number of microstates in the given macrostate} / \text{Total number of microstates of the system.}$

$$P(r) = \frac{n!}{r!(n-r)!} \times \frac{1}{2^n} \quad (\text{C7.6})$$

In the above discussion in terms of compartment, we mean a particular situation. Thus if there are  $n$  coins which are thrown  $N$  number of times, then one compartment means heads up and the other compartment means tails up.

Theoretical and observed frequency: Suppose we have  $n$  identical coins. If we throw them in such a way that  $r$  coins fall with their heads up and the remaining  $(n-r)$  with their tails up then the probability of the event

$$P(r) = \frac{n!}{r!(n-r)!} \times \frac{1}{2^n} \quad (\text{C7.7})$$

If the throw is repeated  $N$  times,  $N$  being very large, then for this event theoretical expected frequency  $F(r) = N \times P(r)$ . The actual observed frequency of  $r$  heads up is denoted by  $f(r)$ .

For very large value of  $N$  i.e. for very large number of trials of tossing of two coins, the theoretical expected frequency  $F(r)$  will almost be the same as actual observed frequency  $f(r)$  and the deviation if any will be very small.

**C7.4: Procedure:** 1. Place the tray on the table. See that it is clean and there are no dust particles in it. Take a single coin and put it in the given container. Shake it well and

throw the given coin in the tray. Note whether the head of the coin is up (a) or the tail is up (b). Repeat the process at least 100 times.

2. Now take two identical coins and put them in the cylindrical container. Shake the container well and throw the coins in the tray. Note down each time the combination is produced. If the heads of both coins are up record it as (a, a), if the head of one and the tail of the other is up record it as (a, b) and if the tails of both are up record it as (b, b). Repeat atleast 100 times.

3. Now take 10 identical coins and put them together in the cylindrical container. Shake them well and again throw them in the tray. Record the number of coins which show with heads up and also the number of coins which show with tails up. Repeat the process 100 times.

4. Draw a graph between expected frequency  $F(r)$  and  $r$  and between the observed frequency  $f(r)$  and  $r$ .

5. Evaluate the  $\chi^2$  statistic for the data collected.

## C7.5: Observations

### 1. With one coin

Number of throws,  $N = 100$

Number of coins,  $n = 1$ .

**Table C7.1: Record of observations using a single coin.**

Number	1	2	3	4	5	.....	98	99	100
Heads up	a	a	X	a	X	.....	..	X	a

Tails up	X	X	b	X	b	.....	..	b	X
----------	---	---	---	---	---	-------	----	---	---

Number of times head is up = A =

Theoretical probability of getting head up =  $\frac{1}{2}$

Therefore, Number of times head should be up

Theoretically (for 100 trials) =  $\frac{1}{2} \times 100 = 50$

Percentage variation =  $(50 - A)/50 \times 100 =$

## 2. With two coins

Number of trials N = 100

Number of coins n = 2

Record of observations as shown in the table below:

**Table C7.2: Record of observations using 2 coins**

No.	1	2	3	4	5	6	.....	99	100
Heads of two together	aa	X	X	X	aa	X	.....	aa	x
Tails of two together	X	bb	X	X	X	bb	.....	X	bb
Heads of one and tail	X	X	ab	ab	X	X	.....	X	X

of other									
----------	--	--	--	--	--	--	--	--	--

Experimental count of head-head combination = A =

Experimental count of tail – tail combination = B =

Experimental count of head – tail combination = C =

Theoretical probability of head – head or tail – tail combination =  $\frac{1}{4}$

Theoretical probability of head-head or tail- tail combination =  $100 \times \frac{1}{4} = 25$

Percentage deviation of head – head combination =  $\frac{25 - A}{25} \times 100 = \frac{25 - A}{25} \times 100 =$

Percentage deviation of tail – tail combination =  $\frac{25 - B}{25} \times 100 = \frac{25 - B}{25} \times 100 =$

Theoretical probability of Head – tail combination =  $\frac{1}{2}$

Theoretical count of head – tail combination =  $\frac{1}{2} \times 100 =$

Percentage deviation of head – tail combination =  $\frac{50 - C}{50} \times 100 = \frac{50 - C}{50} \times 100 =$

### 3. With 10 coins

Number of trials, N = 100

Number of coins, n = 10

**Table C7.3: Record the observations using 10 coins:**

No.	1	2	3	4	5	6	.....	99	100
No. of heads	4	7	8	6	5	4	....	7	9

up									
No. of tails up	6	3	2	4	5	6	.....	3	1

**Table C7.4: Observation Table**

No. of Trials =

Combinations (r)      (n-r) (Head)    (Tails)	Probability $P(r) = \frac{n!}{r!(n-r)!} \times \frac{1}{2^r}$	Expected frequency $F(r) = N \times P(r)$	Observed frequency $f(r)$	Deviation $f(r) - F(r)$	$\frac{(f(r) - F(r))^2}{F(r)}$
0,10					
1,9					
2,8					
3,7					
4,6					
5,5					
6,4					
7,3					
8,2					

9,1					
10,0					

$\chi^2$  statistic =

**C7.6: Precautions:** 1. The tray should be clean and smooth.

2. The mug or tumbler should be perfectly smooth from inside.
3. The coin should be thin and identical with perfectly smooth edges.
4. The tumbler and coins should be shaken well before throwing the coins.
5. The total number of trials should be as large as possible.

## C8: Michelson Interferometer

**C8.1: Objective:** To measure the wavelength of diode laser using Michelson interferometer.

**C8.2: Apparatus:** Michelson interferometer, laser source, aluminium plate with pinhole.

### C8.3: Theory:

#### (a) General Introduction

The Michelson-Morley experiment using the interferometer, was of crucial importance in the development of the subject of relativity by demonstrating the absence of the so-called 'ether drift'.

In the Michelson Interferometer, two beams obtained by amplitude division are sent in two directions at  $90^\circ$  and are reflected from plane mirrors so that they recombine to form interference fringes. The intensity pattern is proportional to  $\cos^2\left(\frac{2\pi d}{\lambda}\right)$  where  $d$  is the path difference between rays reflected from the two mirrors and  $\lambda$  is the wavelength.

The arrangement of the interferometer and the optical paths are shown schematically in Figure (C8.1). Two highly polished plane mirrors,  $M_1$  and  $M_2$ , and two plane-parallel glass plates form the main optical parts. The rear side of the first plate is partially silvered so that the light coming from the source  $S$  is divided into (1) reflected and (2) transmitted beams of equal intensity (amplitude division). Light reflected normally from mirror  $M_1$  passes through this first glass plate a third time and reaches the eye as shown. Light reflected from the mirror  $M_2$  passes through a second glass plate twice, once on the outward

journey and once after reflection from  $M_2$ . This second glass plate is known as the compensating plate, and its function is to make the path of the two rays in glass equal. In the interferometer, the two glass plates are combined into a single compensator unit. The mirror  $M_1$ , is mounted on a carriage and because during the experiment it is moved only multiples of the wavelength of light, these distances must be amplified by a mechanical lever connected to the micrometer screw gauge as shown in Figure (C8.1). To obtain any kind of fringes with the interferometer, the mirrors,  $M_1$  and  $M_2$  must be made perpendicular to each other by means of screws shown on mirror  $M_2$ , in Figure (C8.1). This alignment is a crucial part of the experiment.

**Fig. 1. The Michelson interferometer**

- A: half-silvered mirror
- B: compensating plate
- $M_1$ : mirror w. displacement control
- $M_2$ : mirror w. tilt control
- $M_2'$ : virtual image of  $M_2$
- P: observer
- S: source
- d: distance between images

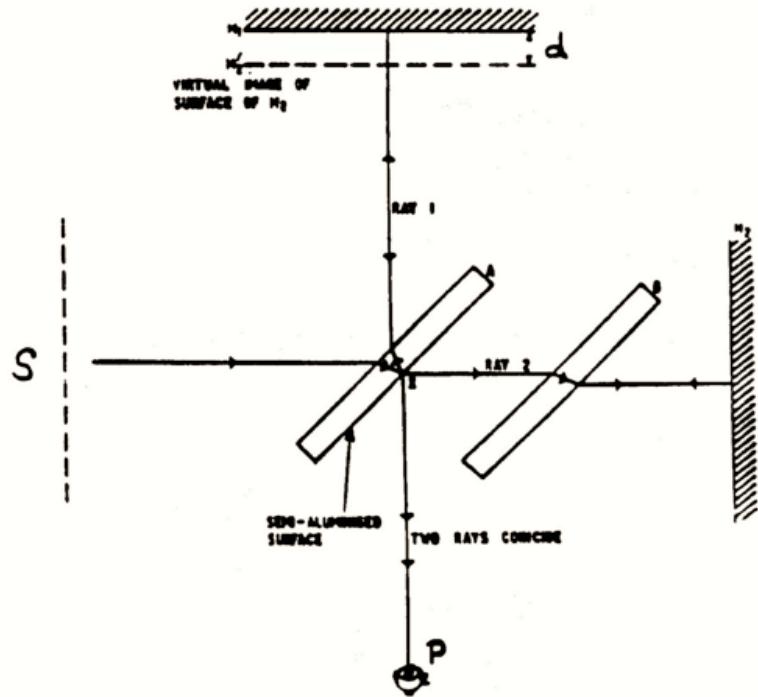


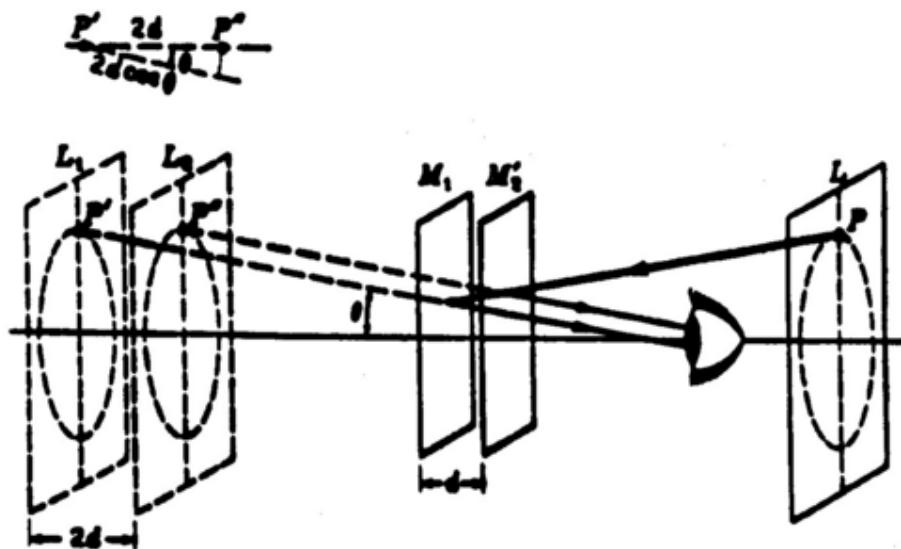
Fig C8.1: The Michelson Interferometer

### (c) Circular and localised fringe systems

Two different systems of fringes can be observed for different settings of the interferometer.

Circular fringes are produced with monochromatic light when the mirrors are in exact adjustment as shown in Figure (C8.2). The real mirror,  $M_2$ , has been replaced by its virtual image,  $M_2'$ , formed by reflection in  $G_1$ . Mirror  $M_2'$  is then parallel to  $M_1$ . These virtual sources are coherent because the phases of corresponding points in the two sources are exactly the same at all instants. If  $d$  is the separation  $M_1M_2'$ , the virtual sources will be separated by  $2d$ . When  $d$  is exactly an integral number of half wavelengths, all light reflected normally to the mirrors will be in phase. The path difference between the two rays coming to the eye from corresponding points  $P'$  and  $P''$  is  $2d\cos\theta = n\lambda$ , as shown in Figure (C8.2). Hence, when the eye is focused to receive parallel rays, the rays will reinforce each other to produce maxima for those angles  $\theta$  which satisfy the equation,

$$2d\cos\theta = n\lambda \quad (\text{C8.1})$$



**Fig C8.2:** Formation of circular fringes for a divergent beam.

Since, for a given  $n$ ,  $\lambda$ ,  $d$  and the angle  $\theta$  is constant, the maxima will lie in the form of circles centered on the perpendicular from the eye to the mirrors. Fringes of this type are called fringes of equal inclination.

If on the other hand, the mirrors  $M_1$  and  $M_2$  are not exactly parallel, fringes will still be seen with monochromatic light for path differences which do not exceed a few millimeters. In this case the space between the mirrors is wedge-shaped. These localized fringes are almost straight because the path difference across the field of view is due primarily to the variation of the thickness of the "air film" between the mirrors.

#### C8.4: Procedure

(a) Adjustment of the interferometer Michelson Interferometer is said to be in normal adjustment when the partially silvered beam splitter surface  $G_1$ , accurately bisects the angle between the reflecting mirrors  $M_1$  and  $M_2$  and normal to the fringes could be concentric circles. This can be done as follows:

1. Put the interferometer on a rigid table and level the instrument with three leveling screws provided at the base.
2. Now put the diode laser, about 50 to 60 cm away from the instrument such that its beam passes through the pinhole fitted in front of the instrument. Make sure that the laser beam falls at the middle of the Mirrors  $M_1$  and  $M_2$  after getting split from beam splitter plate.
3. Adjust the distances of  $M_1$  and  $M_2$  such that they are almost equal from  $G_1$ .
4. The beams after reflection from mirrors  $M_1$  and  $M_2$  will result in four spots on the wall or on a screen. One pair of spots 1 and 2 is formed due to partial reflection at the silvered surface of  $G_1$  and reflections from  $M_1$  and  $M_2$  respectively. While the other pair of spots is

formed due to partial reflections at  $M_1$  and  $M_2$  respectively. Out of these one pair is brighter than the other.

5. Now mirrors  $M_1$  and  $M_2$  are tilted carefully such as the two brighter images coincide.
6. Now the instrument is aligned. Introduce the lens in between the laser and pin hole and remove the pinhole out of the path of the beam. See that expanded beam falls on the two mirrors  $M_1$  and  $M_2$ . You will see the fringes on the wall or screen.
7. By using the fine tilt of fixed mirror and motion of the mirror  $M_1$ , you can adjust the center of the circular fringes.
8. Do not use the telescope and do not see directly into the laser beam.

#### **Determination of wavelength:**

First note the value on the screw gauge for the position of mirror corresponding to fringe in the center. Now move the mirror through some distance and count the number of bright fringes which appear to move in the outward direction from the center. The distance traveled by the mirror must be read off on the micrometer screw. One can then obtain the wavelength in the following way.

$$2d_1 = (2m_1 - 1)\lambda/2 ,$$

$$2d_2 = (2m_2 - 1)\lambda/2 \quad (C8.2)$$

$$2(d_2 - d_1) = 2(m_2 - m_1)\lambda/2 \quad (C8.3)$$

$$2 \Delta d = \Delta m \lambda$$

$$\text{Or } \lambda = \Delta d / N \quad (C8.4)$$

#### **C8.5: Observation and Results:**

Initial position of the screw gauge ( $d_0$ ) = \_\_\_\_\_ mm

**Table C8.1:**

S. No.	$d_i$ (in mm)	$\Delta d = d_{i-1} - d_i$	N	$\lambda_i = \frac{\Delta d}{N}$
1				
2				
3				
4				
5				
6				
7				
8				
9				
10				

$$\text{Then } \lambda = \sum_{i=1}^{10} = \frac{\lambda_i}{N}$$

**Result:** The wavelength of diode laser is =

**C8.6: Precautions:** 1. While assembling and operating the interferometer it is important to remember that laser beams can cause severe eye damage. Do not look

directly into the laser beams. Keep your head well above the horizontal plane of the laser beams at all

times. Use white index cards to locate beam spots along the various optical paths.

2. After turning on the laser allow 5 minutes to pass before attempting to make any adjustments.

3. If you turn off the laser for some reason, wait for 5 minutes before turning it on to allow it to cool down.

## C9: Millikan oil Drop Experiment

**C9.1: Aim:** To determine the size of charge on an electron.

**C9.2: Apparatus:** Vacuum pump, oil, microscope, Display, oil drop apparatus.

**C9.3: Theory:** In year 1897, J. J. Thomson demonstrated in a series of experiments that cathode rays were made up of small negatively charged particles, which were soon named electrons. The electron was the first subatomic particle ever discovered.

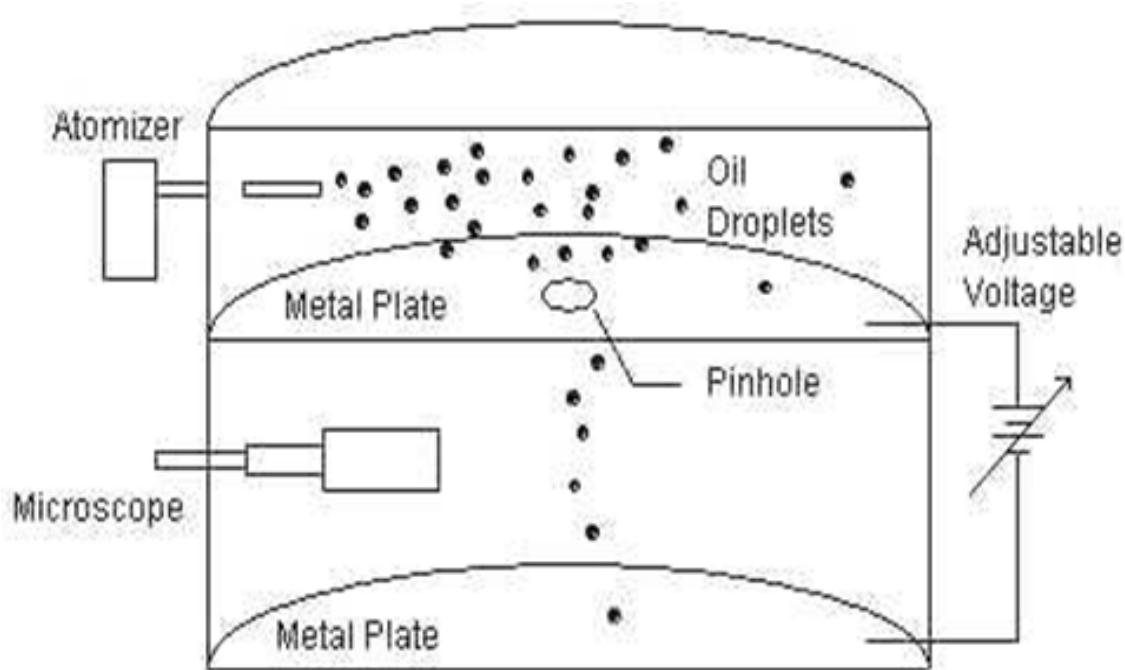
Millikan's oil-drop experiment was performed by Robert Millikan and Harvey Fletcher in 1909. It determined a precise value for the electric charge of the electron,  $e$ . The electron's charge is the fundamental unit of electric charge, because all electric charges are made up of groups (or the absence of groups) of electrons. This discretisation of charge is also elegantly demonstrated by Millikan's experiment.

The unit of electric charge is a fundamental physical constant and crucial to calculations within electromagnetism. Hence, an accurate determination of its value was a big achievement, recognised by the 1923 Nobel prize for physics.

**C9.4: Millikan's Apparatus:** Millikan's experiment is based around observing charged oil droplets in free fall and in the presence of an electric field. A fine mist of oil is sprayed across the top of a cylinder with a small 'chimney' that leads down to the cell (if the cell valve is open). The act of spraying will charge some of the released oil droplets through friction with the nozzle of the sprayer. The cell is the area enclosed between two metal plates that are connected to a power supply. Hence an electric field can be generated within the cell and its strength varied by adjusting the power supply. A light is used to illuminate the cell and the experimenter can observe within the cell by looking through a microscope.

## Terminal velocity

As an object falls through a fluid, such as air or water, the force of gravity will accelerate the object and speed it up. As a consequence of this increasing speed, the drag force acting on the object, that resists the falling, also increases. Eventually these forces will balance (along with a buoyancy force) and therefore the object no longer accelerates. At this point the object is falling at a constant speed, which is called the terminal velocity. The terminal velocity is the maximum speed the object will obtain while free falling through the fluid.



**Fig 9.1: Millikan's Oil Drop experiment Apparatus**

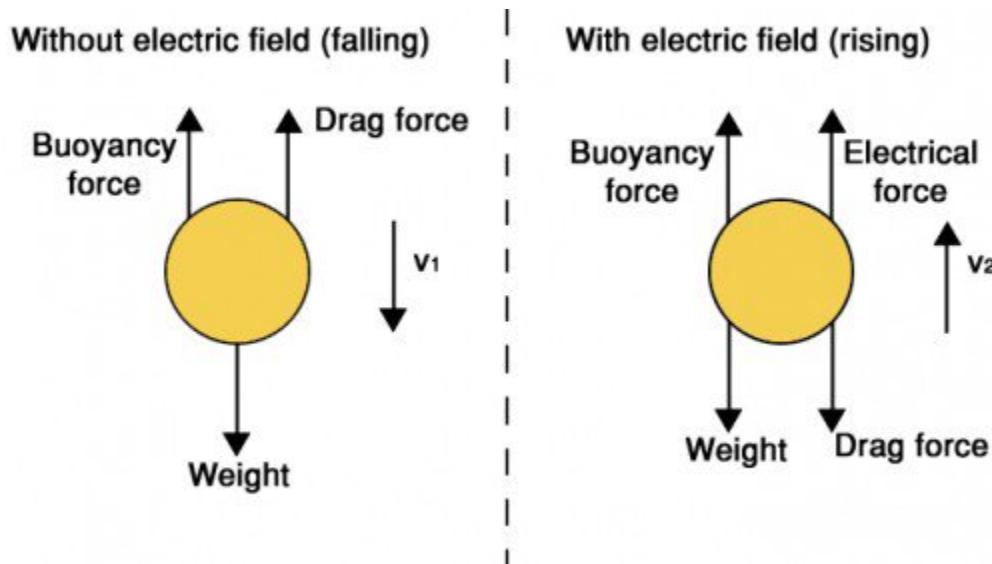
## Theory

Millikan's experiment focuses on the motion of individual charged oil droplets within the cell.

To understand this motion the forces acting on an individual oil droplet need to be considered.

As the droplets are very small, the droplets are reasonably assumed to be spherical in shape.

The diagram below shows the forces and their directions that act on a droplet in two scenarios: when the droplet free falls and when an electric field causes the droplet to rise.



The different forces acting on an oil drop falling through air (left)

and rising through air due to an applied electric field (right).

The most obvious force is the gravitational pull of the Earth on the droplet, also known as the weight of the droplet. Weight is given by the droplet volume multiplied by the density of the oil ( $\rho_{oil}$ ) multiplied by the gravitational acceleration ( $g$ ). Earth's gravitational acceleration

is known to be  $9.81 \text{ m/s}^2$  and the density of the oil is usually also known (or could be determined in another experiment). However, the radius of the droplet ( $r$ ) is unknown and extremely hard to measure.

$$F_{\text{weight}} = mg = \frac{4}{3}\pi r^3 \rho_{\text{oil}} g \quad (\text{C 9.1})$$

As the droplet is immersed in air (a fluid) it will experience an upward buoyancy force. Archimedes' principle states that this buoyancy force is equal to the weight of fluid displaced by the submerged object. Therefore, the buoyancy force acting on the droplet is an identical expression to the weight except the density of air is used ( $\rho_{\text{air}}$ ). The density of air is a known value.

$$F_{\text{buoyancy}} = \frac{4}{3}\pi r^3 \rho_{\text{air}} g \quad (\text{C 9.2})$$

The droplet also experiences a drag force that opposes its motion. This is also called air resistance and occurs as a consequence of friction between the droplet and the surrounding air molecules. Drag is described by Stoke's law, which says that the force depends on the droplet radius, coefficient of viscosity of air ( $\eta$ ) and the terminal velocity of the droplet during the fall ( $v_f$ ). The constant drag force is given by

$$F_{\text{drag}} = 6\pi r \eta v_f \quad (\text{C 9.3})$$

The gravitational and buoyancy forces acting on the droplet are balanced by the drag force

$$\frac{4}{3}\pi r^3 \rho_{oil} g - \frac{4}{3}\pi r^3 \rho_{air} g = 6\pi r \eta v_f \quad (\text{C 9.4})$$

The falling velocity  $v_f$  is thus given by

$$v_f = \frac{2gr^2}{9\eta} (\rho_{oil} - \rho_{air}) \quad (\text{C 9.5})$$

If the droplet carries a charge  $ne$  and is moving upward with terminal velocity  $v_r$  under the influence of the applied electric field  $V/d$  between the parallel plate electrodes separated by the distance  $d$  and potential difference  $V$ , the force balance equation is,

$$\frac{4}{3}\pi r^3 \rho_{oil} g - \frac{4}{3}\pi r^3 \rho_{air} g - 6\pi r \eta v_f = \frac{Vne}{d} \quad (\text{C 9.6})$$

Subtracting (C 10.4) from ( C10.6 ) and solving for 'ne', we get

$$ne = \frac{6\pi\eta rd}{V} (v_f + v_r) \quad (\text{C 9.7})$$

Dividing (7) by (5), we get

$$ne = \frac{4\pi gd}{3V} (\rho_{oil} - \rho_{air}) r^3 (1 + \frac{v_r}{v_f}) \quad (\text{C 9.8})$$

The Stokes law used in obtaining eqn. (4) - (8) assumes that the droplets are moving slowly, that there is no slipping of the medium over the surface of the droplet, that the medium is of quite large extent compared to the size of the droplet and that inhomogeneities in the medium are of a size small compared to the size of the droplets. In the present case all the assumptions except the last one are reasonably valid. The radii of

the droplets are of the order of one micron and therefore not greater than the mean free path of the air molecules. The droplets will tend to fall more quickly in the free space between the air molecules. The expression for the falling velocity  $v_f$  corrected for this effect on the basis of kinetic theory is

$$v_f = \frac{2}{9} \frac{g r^2}{\eta} (\rho_{oil} - \rho_{air})(1 + \frac{c}{P_r}) \quad (C9.9)$$

where  $c = 6.17 \times 10^{-8} m \text{ of Hg} - m$  is a correction factor and  $P$  (in mm of Hg) is the atmospheric pressure. Writing

$$\xi = \frac{9\eta}{2g} v_f / (\rho_{oil} - \rho_{air}) \quad (C9.10)$$

$$\zeta = \frac{c}{2P} \quad (C9.11)$$

Equation (C10.9) now reduces to

$$r^2 + 2\zeta r - \xi = 0 \quad (C9.12)$$

The radius of the droplet is given by the positive root of this equation

$$r = -\zeta + \sqrt{\zeta^2 + \xi} \quad (C9.13)$$

The charge  $ne$  may be obtained by first calculating  $\xi$  and  $\zeta$  from eqn. (10) and (11), then calculating the radius  $r$  from eqn. (12) and finally  $ne$  from eqn. (7)

### C9.5: Procedure:

1. Note the atmospheric pressure and room temperature.
2. Switch on the power supply.
3. Fix two horizontal lines on the monitor. These are the present lines. The rise and fall times of the droplets to move from one line to the other are to be recorded. A good choice is to choose the second line from the top and the last but one line from the bottom. This will leave some room at the top and at the bottom for maneuvering and preventing the droplet from getting lost.
4. Press the 'clear key' to make the time meter read zero.
5. Spray droplets of oil from the atomiser. One or two sprays are sufficient.
6. As droplets drift down, some of them pass through the hole in the upper parallel plate and reach the region in-between the plates where they are illuminated.
7. These are viewed by the microscope as unresolved points of diffracted light and the images are transmitted to the monitor screen.
8. These droplets drift down slowly under gravitational force.
9. Under the influence of the electric field between the plates, the motion of the droplets will get affected if they are charged. If a droplet moves downward more slowly under the influence of the electric force (corresponding to the upper plate at a positive potential), the drop is negatively charged. Its motion can be arrested or even reversed by an increase in potential. If the downward drifting of the droplet increases under the action of electric force, the droplet is positively charged. Such

droplets are ignored in the present set up. After a short interval of time, only negatively charged droplets will be left.

10. Suitable negatively charged droplet is selected. Suitable means that it drifts downward slowly (about 10 - 15 s free fall time) under gravitational field indicating that the droplet is not too heavy. Its mass should also not be too small, otherwise it will bounce around due to random collisions with air molecules, Brownian motion, and it will be difficult to estimate when the droplet actually crosses the line. It should also be possible to make it rise by applying a voltage of about 500 V indicating that there are only a few charge quanta on the droplet.
11. Measure the free fall time with voltage off and the rise time with some suitable voltage on the droplet to move between the two lines chosen on the monitor. Repeat these measurements of free fall time and the rise time several times. Take average free fall and rise times. Keep the voltage fixed.
12. Repeat step 11 using different droplets and applying different voltage to each.
13. Evaluate the value of  $ne$ . Identify the minimum value of the charge  $ne$  on the droplets.
14. Divide the value of the charge  $ne$  on all the droplets by this minimum value.
15. The result will be numbers close to integers for all the droplets.
16. Now extrapolate all the numbers or the earlier ones if no multiplication has been done to the nearest integer.
17. Divide the value of the charge  $ne$  on all the droplets by the respective integers.

18. This gives the charge on an electron by different droplets. Take the average of these values.

### C9.6: Observations and Calculations:

1. Distance between the parallel parallel plates 'd' =  $5 \times 10^{-3}$  m.
2. Distance L between the chosen top and bottom lines on monitor =  $1 \times 10^{-3}$  m
3. Relative density of oil with respect to air =  $\rho_{oil} - \rho_{air} = 928 \text{ kg/m}^3$
4. Room temperature T =
5. Atmospheric pressure P = mm of Hg
6. Coefficient of viscosity of air  $\eta = 1.842 \times 10^{-5} \text{ Kg/msec.}$

### Observation Table C9.1:

Droplet No.	S.No.	Free fall time	Rise Time	Mean free fall time $t_f$ (sec)	Mean rise time $t_r$ (sec)	Mean free fall velocity $v_f$	Voltage V (in volt)
1	1						
	2						
	3						
	4						
	5						

2	1					
	2					
	3					
	4					
	5					
3	1					
	2					
	3					
	4					
	5					
4	1					
	2					
	3					
	4					
	5					
5	1					
	2					

	3						
	4						
	5						

### Calculations of common constants:

1.  $C = 4\pi d g (\rho_{oil} - \rho_{air}) / 3 = 190.13$
2.  $D = 9\eta / 2g (\rho_{oil} - \rho_{air}) = 9.04 \times 10^{-9}$
3.  $\zeta = c/2P = 4.06 \times 10^{-8}$ ,  $c = 6.17 \times 10^{-8}$

### Calculation from droplet data:

Observation Table C9.2

Droplet No.	$\xi = D v_f$	$r = \sqrt{-\zeta + \sqrt{\zeta^2 + \xi}}$	$r^3$	$T = 1 + \frac{t_f}{t_r}$	$ne = CT r^3 / V$
1					
2					
3					
4					
5					

**Observation Table C9.3**

S.No.	ne	ne divided by the lowest value of ne	Nearest integer $n_{eff}$	$ne/n_{eff}$
1				
2				
3				
4				
5				

**C9.7: Result:** Mean value of  $ne/n_{eff} =$

Actual value =  $1.6 \times 10^{-19} C$

Percentage Error =

**C9.8: Suggested Viva questions:**

1. What did Millikan's oil drop experiment reveal about the nature of electric charge?
2. Is oil droplets positively or negatively charged? How did they acquire charge?

3. Why did Millikan use oil instead of water?
4. Why do some oil droplets move up and some down?
5. Which are the forces acting on the droplet?

## C10: Thomson's e/m Ratio

**C:10.1: Aim:** To measure charge to mass ratio of an electron.

**C:10.2 Apparatus:** Thomson's e/m ratio experimental setup. It consists of an electron gun with a filament, cathode, anode as its sub-parts, deflecting plates , Helmholtz coils and a glass bulb containing the apparatus.

The arrangement for measuring e/m ratio, that is, charge to mass ratio of the electron, is a simple sep-up. It is based on Thomson's method. The e/m tube is bulb-like and contains a filament, a cathode, a grid, a pair of deflection plates and an anode. The filament heats the cathode which emits electrons. The electrons are accelerated through a known potential applied between the cathode and the anode. The grid and the anode have a hole through which electrons can pass. The tube is filled with Helium at very low pressure. Some of the electrons emitted by the cathode collide with Helium atoms which get excited and radiate visible light. The electron beam thus leaves a visible track in the tube and all manipulations on it can be seen. The tube is placed between a pair of fixed Helmholtz coils which produce a uniform and known magnetic field. The beam is deflected in a circular path of radius r depending on the accelerating potential V, the magnetic field B and the charge to mass ratio e/m. This circular path is visible and the diameter d can be measured and e/m ratio obtained from relation

$$\frac{e}{m} = \frac{8V}{B^2 d^2}$$

This setup can also be used to study the electron beam deflection for different directions of the magnetic field by varying the orientation of the e/m tube.

The deflecting plates play no role in the experiment. They are interesting for visual observation of how the electron beam gets deflected when a potential difference is applied between the deflecting plates.

### C.10.3: Theory:

#### **Relationship connecting e/m ratio to accelerating potential V, magnetic field B and radius r of the circular path.**

When the electrons are accelerated through the potential V, they gain kinetic energy equal to their charge times the accelerating potential. Therefore,

$$eV = \frac{1}{2}mv^2 \quad (\text{C 10.1})$$

Therefore, the final velocity (non-relativistic) of the electrons is given by

$$v = \sqrt{\frac{eV}{2m}} \quad (\text{C 10.2})$$

When these electrons pass through a region having magnetic field B, they are acted upon by a force, called the Lorentz force, given by  $ev \times B$ . If the electrons are initially moving along x-axis and the magnetic field is along the z-axis, the electrons describe a circular path in the xy-plane with the centripetal force balancing the Lorentz force,

$$\begin{aligned} evB &= \frac{mv^2}{r} \\ \text{or } v &= \frac{eBr}{m} \end{aligned} \quad (\text{C 10.3})$$

Eliminating v between Eq. (1) and (3), we get

$$\frac{eBr}{m} = \sqrt{\frac{2eV}{m}}$$

$$\text{or } \frac{e}{m} = \frac{2V}{B^2 r^2} = \frac{8V}{B^2 d^2} \quad (\text{C 10.4})$$

where 'd' is the diameter of the circular path. This result assumes that the magnetic field B is uniform. The field B is produced by a pair of Helmholtz coils(separated by a distance equal to their radius). If n is the number of turns in a coil and 'a' is its radius, then the magnetic field B, midway between the coils is given by

$$B = 2 \times \frac{\mu_0 I n}{2(5/4)^{3/4} a} = 2 \times \left( \frac{2\pi I n}{2(5/4)^{3/4} a} \times 10^{-7} \right) \text{Tesla} \quad (\text{C 10.5})$$

when the current of I amp is flowing in the coils.  $\mu_0$  is permeability of free space and is given by  $\mu_0 = 4\pi \times 10^{-7} \text{ N/A}^2$ . This field is uniform in the region where the electrons move. Putting the value of B in Eq (), we get

$$\frac{e}{m} = \left( \frac{125 a^2}{128 \pi^2 n^2} \times 10^{14} \right) \frac{V}{I^2 d^2} \quad (\text{C 10.6})$$

#### C 10.4. Procedure:

1. Before the power is switched on, make sure all the control knobs are at their minimum position.
2. Turn on the apparatus and wait a little for the cathode to heat up.
3. Turn the accelerating voltage adjust knob clockwise to increase the voltage.  
Rectilinear electron beam emerging from the cathode will be visible. Adjust the accelerator voltage at about 200 V.
4. The electrons themselves are not visible. The green glow observed in the tube is due to the Helium gas filled in the bulb. When electrons collide with the atoms of the gas, the gas atoms gets excited and emit green light on deexcitation.

5. Rotate the e/m - tube so that the electron beam is parallel to the plane of the Helmholtz coils. Do not take it out of the socket.
  6. Earth's magnetic field interfere with the measurements. However, this magnetic field is weak and we can ignore it in the calculations as a first approximation.
  7. Slowly turn the current adjust knob clockwise to increase the current for the Helmholtz coils. The electron beam will get curved. Increasing the current will increase the curvature of the beam.
  8. In case the electron beam does not make a complete closed circular path, rotate the socket of the tube until the path is closed circle.
  9. In order to measure the diameter of the electron beam, a hollow tube(fitted with cross wires at its both ends) is fixed on the slider of the scale. This tube fixes the line of sight during measurements.
10. Note the ammeter reading for the current to the Helmholtz coils and the voltmeter reading for the accelerating voltage.
11. Decrease the accelerating voltage by a small amount ( say 20 V) and measure the diameter of the electron beam path.

#### C.10.5: Observation Table:

Accelerating	Current to	Diameter of	(Diameter) <sup>2</sup>	$V / I^2d^2$
--------------	------------	-------------	-------------------------	--------------

Voltage (Volt)	Helmholtz coils (A)	the beam path (m)	(m <sup>2</sup> )	

**C.10.6: Graph:** For a given value of current I, draw a graph between V and d<sup>2</sup>. It should be a straight line.

$$\text{Value of } \frac{V}{I^2 d^2} \text{ from graph} =$$

$$\text{Value of } e/m \text{ from graph} =$$

#### **C.10.7. Precautions and Sources of Error:**

1. The main source of error in this experiment is the velocity of the electrons. There is a hole in the anode to allow the electrons to pass through it. This makes the velocity of the electrons non-uniform and slightly less than the theoretical value. Further the collisions of the electrons with the helium gas in the tube decrease their velocity a little bit. The effect of these errors can be minimized by measuring

the outside radius of the electron beam path and by not using low values of the accelerating voltage.

2. Other source of error is the measurement of the diameter of electron.

#### **C10.8: Suggested Activity:**

1. Evaluation of Magnetic field being applied to the apparatus. Henceforth compare its value with the value of Earth's magnetic field.

#### **C10.9: Suggested viva questions:**

1. Discuss how much difficulty the Earth's magnetic field will cause in your experiment.
2. In what direction should the apparatus be aligned to minimize the effects of the Earth's magnetic field? Explain.
3. What happens to the beam of electrons as the coil current is increased?
4. What do you observe as the accelerating voltage is changed while keeping the coil current constant? Explain why this occurs.
5. What is the significance of measuring charge to mass ratio?

## C11: HALL EFFECT

**C11.1: Objective:** To determine the Hall coefficient of a given semiconductor material and then evaluate charge carrier type, density and mobility of charge carriers.

### C11.2: Apparatus

The experimental set-up for studying Hall effect and obtaining the characteristic parameters of the semiconductor under investigation comprises of the following components:

**(a) Electromagnet Assembly:** It consists of a constant current power supply connected to assembly of two poles of the electromagnet. The two poles of an electromagnet are in the form of short cylindrical rods of iron tapered at ends which are inserted in coils of copper wires having large number of turns. These iron rods act as core of coils and get magnetized by the field produced when large current is passed through the coil. The pole pieces are connected to a mechanical rotor system which can cause the motion thereby resulting in variation of distance between two poles. This electromagnet can produce a magnetic field of intensity up to 7kG.

**(b) Hall Effect Set up:** It consists of a regulated power supply feeding constant current to the semiconductor crystal (Ge) specimen through a 9 pin connector at the front panel. On the front panel are fixed a digital

ammeter and a digital voltmeter. The ammeter reads the current in milliamperes which is fed to the semiconductor crystal while voltmeter reads the Hall voltage developed across the perpendicular faces of the crystal.

**(c) Hall Probe for experiment:** The Hall probe is in the form of a Germanium crystal specimen, under investigation, with four soldered contacts mounted on a PCB strip. The Hall element is mounted in a pen type case and a 4 core cable is provided for connection with measuring device and current source. This probe is placed between two poles of magnet and a current is passed using two cables connected to opposite soldered contacts. This results in potential difference between other two soldered contacts which are oriented in direction perpendicular to those used for sending current through the Ge crystal.

**(d) Hall Probe for Gauss-meter:** It is similar in construction to the Hall probe but the Ge-crystal is replaced by Indium-Arsenide (InAs) crystal. It has InAs crystal which has soldered to it 4 linearly arranged probes of which outer probes are used to send current while inner two are used to measure Hall voltage. A fixed current is sent through 2 probes and the Hall voltage developed on transverse pair of faces is measured by the remaining two probes. The Hall voltage developed is calibrated in units of magnetic field and read as digital display on the gaussmeter.

**(e) Digital Gauss-meter:** The Gaussmeter is employed for measuring DC

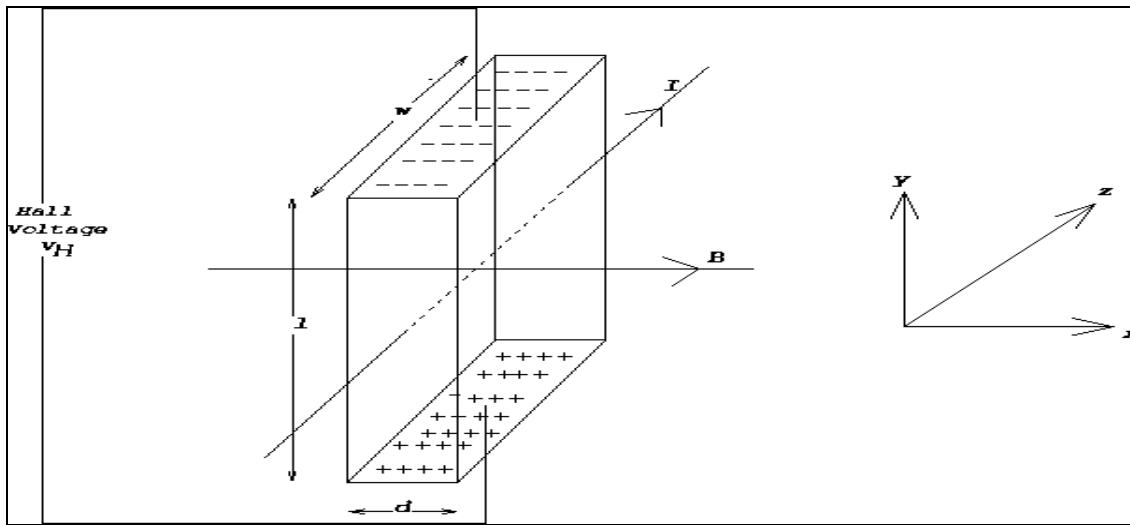
magnetic field upto 20kG. It gives the digital display of the magnetic field through the 3.5 digit, 7 segments LED. The display has two way switch positioning (i) X1 position capable of measuring magnetic field in the range 0 – 2kG with a resolution of 2G (ii) X10 position capable of measuring magnetic field in the range 0 – 20kG with a resolution of 10G. The accuracy of displayed reading is about 3%. This is operated through the AC supply of  $(220 \pm 10\%)$  volts, 50Hz. On the front panel of Gauss-meter is knob for zero adjustment of magnetic field and a nine pin connector to fix the hall probe. If the magnetic field intensity indicated by the Gauss-meter is positive (without sign), the pole facing the crystal of the Hall probe is north.

- (f) The Gauss-meter works on the principle of Hall effect in semiconductors. A semiconductor material carrying current develops an emf, when placed in an applied magnetic field in a direction perpendicular to the flow of electric current. The magnitude of emf developed (Hall voltage) is proportional to the intensity of magnetic field. This small emf is amplified through high stability amplifier and is read by a digital voltmeter (0 – 200mV) connected at the output of the amplifier. This amplifier is directly calibrated in the units of magnetic field.

### C11.3: Theory

The Hall effect was discovered by Edwin Hall in 1879. He observed that when the semiconductor crystal is placed in a magnetic field acting in the direction

perpendicular to the current flowing in it, then an electric field is induced which is perpendicular to both current and magnetic field. The potential difference, so observed, is called the Hall voltage. The figure C11.1 depicts the Hall effect pictorially.



**Figure C11.1:** The pictorial representation of Hall effect.

- (a) **Sign of Charge Carriers:** Suppose the current is due to the flow of positive charge carriers i.e., holes which drift along the positive x direction (say) with velocity  $v$ , the Lorentz force suffered by these carriers due to magnetic field acting along the z direction is

$$F_m = e(v \times B) \quad (C11.1)$$

which is directed along  $-y$  direction, as shown in figure C11.1. It results in deflection and consequent accumulation of positive charge carriers on the bottom surface of the crystal and an electric field will be induced in  $+y$  direction.

The equilibrium condition will be reached when further charge carriers will continue to move along straight line paths due to counterbalancing of magnetic force by the electric force given by

$$F_e = eE_H \quad (C11.2)$$

$$\begin{aligned} F_m &= F_e \\ evB &= eE_H \\ E_H &= vB \end{aligned} \quad (C11.3)$$

The drift velocity of the charge carriers is given as

$$v = \frac{I}{neA} \quad (C11.4)$$

where A is the area of cross-section of the crystal across which the charge flows and n is the number density of charge carriers. Using equation (C11.4) in (C11.3), we get

$$E_H = \frac{IB}{neA} \quad (C11.5)$$

where J is the current density. The Hall coefficient ( $R_H$ ) is defined as

$$R_H = \frac{E_H A}{IB} = \frac{1}{ne} \quad (C11.6)$$

If the charge carriers are negative then, the Hall coefficient is given as

$$R_H = \frac{E_H}{JB} = -\frac{1}{ne} \quad (C11.7)$$

Hence the sign of Hall coefficient gives the type of carrier constituting the current flow in the given semiconductor.

**(b) Measurement of Hall Coefficient:** The Hall coefficient can be determined by measuring the Hall voltage that generates the Hall field. If  $V_H$  is the Hall voltage developed across the crystal of thickness d and width w, then Hall coefficient is given as

$$R_H = \frac{V_H ld}{IB} = \frac{V_H d}{IB} \quad (C11.8)$$

(c) **Mobility of Charge Carriers:** The mobility ( $\mu$ ) of a charge carrier is defined as the ratio of its drift velocity to the applied electric field strength.

$$\mu = \frac{v}{E} \quad (C11.9)$$

Hence equation (C11.3) can be expressed as

$$\mu = \frac{R_H J}{E} \quad (C11.10)$$

According to the microscopic form of Ohm's law, we have  $J = \sigma E$ . Hence

$$\mu = R_H \sigma = \left(\frac{E_H}{E}\right)\left(\frac{1}{B}\right) \quad (C11.11)$$

The ratio  $\phi = E_H / E$  is called Hall angle. Hence mobility can be expressed as

$$\mu = \frac{\phi}{B} \quad (C11.12)$$

## C11.4: Procedure

**(a) Zero Setting of Gaussmeter:** Now place the Hall probe (InAs) attached with gauss-meter between the pole pieces to measure the magnetic field. Reduce the current in the electromagnet power supply to zero. If the gauss-meter depicts non-zero value of magnetic field, then it should be adjusted to zero using zero adjustment knob.

**(b) Zero Calibration:** It has to be done before starting the experiment. Some current is given to the semiconductor sample which is kept outside the magnetic field (magnetic field is switched OFF during this calibration). Some voltage will be developed. Actually it should be zero and for this purpose, four pins in contact with the semiconductor can be adjusted to give zero voltage. However the adjustment of these pins is fixed, the Hall voltage developed in this case will be called **zero field potential** which should be subtracted from the subsequent readings. Increase the current in the steps of 0.2mA and note the corresponding zero field potential ( $V_{H0}$ ) at different values of current.

**(c) Connections:** The connections in the Hall probe are such that widthwise terminals are for the voltage output and lengthwise are for the current input. Check the connections and switch ON the Hall effect set up and adjust the current to a few mA.

**(d) Placement of Hall Probe:** Place the Hall probe in the magnetic field by placing it in the space between the poles of an electromagnet by fitting it in the stand. This probe should be placed so the face of crystal is perpendicular to the magnetic field. Switch ON the electromagnet by connecting the pole pieces to the power supply.

**(e)  $V_H$  vs I Study:** Measure the Hall voltage ( $V_{HM}$ ) at different values of current flowing through the crystal keeping the magnetic field intensity constant. The actual value of Hall voltage will be given by subtracting zero field potential ( $V_{H0}$ ) value corresponding to the same current value i.e.  $V_H = V_{HM} - V_{H0}$ . The constant value of magnetic field should be kept as high as 2kG or higher so that stray fields do not have notable contribution.

**\*Note: The current should not be increased more than 2mA.**

**(f)  $V_H$  vs B Study:** Measure the Hall voltage as a function of magnetic field keeping a suitable constant value of current flowing through the semiconductor crystal. The actual value of Hall voltage will be given by subtracting zero field potential ( $V_{H0}$ ) value corresponding to constant current value chosen i.e.  $V_H = V_{HM} - V_{H0}$ . The magnetic field strength be increased in steps of 250-500G so as to have a large number of observations for graph plotting.

**(g) Determination of Hall Coefficient:** Plot the graph between Hall voltage ( $V_H$ ) and magnetic field (B). Also plot the graph between the Hall voltage ( $V_H$ ) and current (I). From the slopes  $V_H/I$  and  $V_H/B$ , calculate the hall coefficient in two different ways [using the equation (C11.8) ] and take their mean value.

**(h) Deductions of Charge Carrier Density:** Further the charge carrier density can be evaluated by using Hall coefficient value in equation

$$n = \frac{1}{eR_H}$$

**(i) Deduction of Mobility of Charge carriers:** The mobility of charge carriers can also be evaluated by using the equation

$$\mu = R_H \sigma = \frac{R_H}{\rho}$$

### C11.5: Observations

Least count of digital voltmeter of Hall effect set up =

Least count of digital ammeter of Hall effect set up =

Least count of digital gaussmeter =

**Table C11.1: Hall Voltage vs Current at constant magnetic field**

Intensity of magnetic field =

S.No.	Current(in mA)	$V_{HO}$ (in mV)	$V_{HM}$ (in mV)	$V_H = V_{HM} - V_{HO}$ (in mV)
1.				
2.				
3.				
4.				
5.				
6.				
7.				
8.				
9.				
10.				
11.				

12.				
13.				
14.				
15.				

**Table C11.2: Hall Voltage vs magnetic field at constant current**

Current flowing through Semiconductor =

Zero field potential =

S.No.	Magnetic Field (in G)	$V_{HM}$ (in mV)	$V_H = V_{HM} - V_{HO}$ (in mV)
1.			
2.			
3.			
4.			
5.			
6.			
7.			
8.			
9.			
10.			
11.			
12.			
13.			
14.			
15.			

## C11.7: Calculations and Results:

$$R_H(\text{from } VH \text{ vs } I \text{ graph}) = \frac{V_{HD}}{IB} =$$

$$R_H(\text{from } VH \text{ vs } B \text{ graph}) = \frac{V_{HD}}{IB} =$$

Average value of  $R_H$  =

$$\text{Charge carrier density, } n = \frac{1}{eR_H}$$

$$\text{Mobility of charge carriers} = \mu = R_H \sigma =$$

## C11.6: Precautions and Sources of Error

- (a) The Hall probes (both for measurement of field and that containing Ge crystal) must be handled carefully as InAs and Ge crystals are fragile. The mounting of these probes on stand must be carefully done without exerting any extra pressure.
- (b) Due to their high resistivity, the Ge crystal warms up even when a small current is passed through it. To avoid damage of Ge crystal, do not pass more than 2mA current.
- (c) The Hall probe is placed in the magnetic field in such a way that maximum Hall voltage is generated.
- (d) The Zero field Hall voltage must be measured outside the electromagnet as there might be some remnant magnetic field due to repeated hysteresis cycling of pole pieces through magnetic field.
- (e) Hall voltage developed must be measured very accurately.
- (f) The magnetic field must be varied gradually so that the coils of the electromagnet are not damaged.
- (g) The air gap between the two poles of the electromagnet must be less than 2cm.

(h) Make sure that no RF source (e.g. cellular phone) is in operation in the vicinity of the set up.

### C11.7: Suggested questions for Viva:

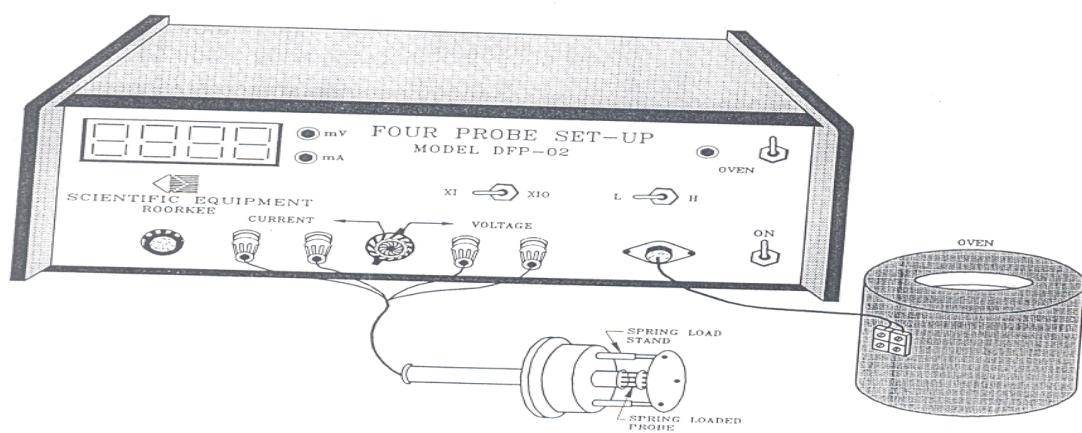
1. What is Hall effect?
2. What is the significance of Hall effect?
3. What is Hall potential? How is it developed in the sample used in the experiment?
4. What is Hall probe used for?
5. What is the sign of Hall Coefficient for an intrinsic semiconductor?
6. Is it possible to measure Hall's coefficient for metals?
7. Why is the Hall coefficient of semiconductors many orders of magnitudes greater than that of metals?
8. What is the effect of temperature on Hall coefficient of a lightly doped semiconductor?
9. What is Hall angle?

# C12: FOUR PROBE METHOD

**C12.1: Objective:** To study the temperature dependence of resistivity of a semiconductor using four probe method and determine the energy band gap of given semiconductor.

## C12.2:Apparatus (Four Probe Set-up)

The four probe set-up is used to measure the resistivity of thin semiconductor films where ohmic contacts and charge carrier injection contribute significantly so as to distort the measurement. The set-up used is a compact table top model shown in the figure C12.1. This set-up is composed of an oven to heat the semiconductor sample, four probe arrangement, closed box containing circuitry to measure the voltage developed when a constant current is supplied to the semiconductor specimen. The current and voltage can be read from the digital display.



**Figure C12.1:** The schematic view of four probe set-up.

**C12.2.1: Four Probes Arrangement:** It has four individually spring loaded probes which are collinear and equally spaced. The probes are mounted in a Teflon bush, which ensures a good electrical insulation between them. A Teflon spacer near the tips is also provided to keep the probes at equal distance. The whole arrangement is provided on a suitable stand and the leads are provided for the voltage measurement.

**C12.2.2: Samples:** The sample is in the form of a germanium crystal chip of 5mm x 10mm. It should be handled with care as it is highly brittle.

**C12.2.3: Oven:** It is a small oven for the variation of temperature of the crystal from the room temperature to about 200°C.

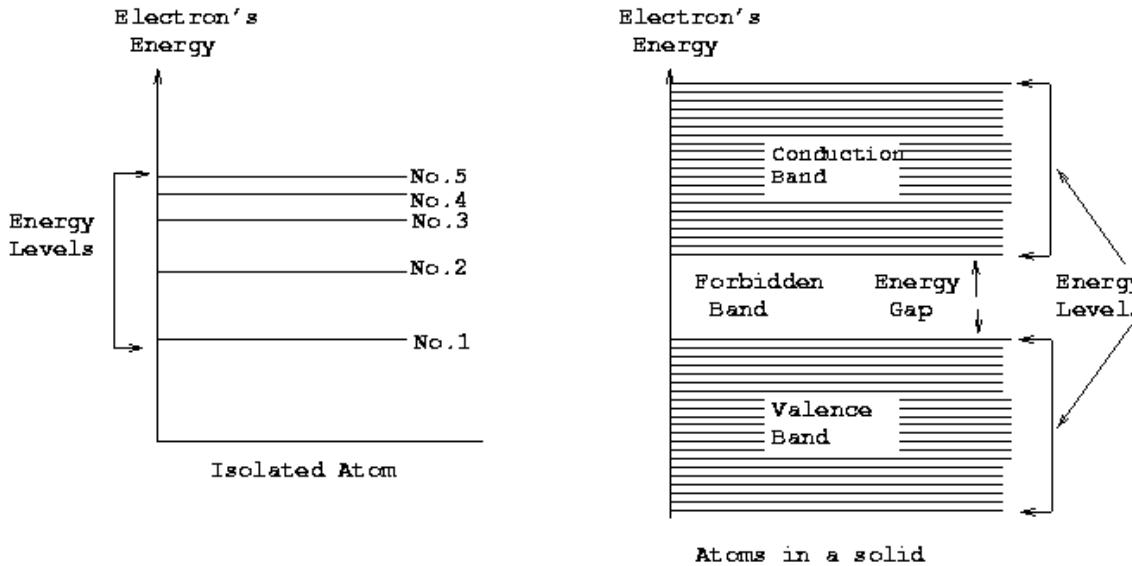
**C12.2.4: Four Probe Set-up:**

- (a) *Multi-range Digital Voltmeter:* The digital voltmeter has a range (0-200mV) at X1 scale and 0-2V at X10 scale.
- (b) *Constant Current Generator:* It is an IC regulated current generator to provide a constant current to the outer probes irrespective of the changing resistance of the sample as a result of temperature variation. The basic scheme is to use the feedback principle to limit the load current of the supply to the preset maximum value. The variation in the current is achieved by a potentiometer included for that purpose. The supply is highly regulated and practically ripple free DC source. The current is measured by the digital panel meter of range of 0-20mA.
- (c) *Oven Power Supply:* Suitable voltage for the oven is obtained through a step down transformer with a provision for low and high rates of heating. A glowing LED indicates that the oven power supply is ON.

**C12.3: Theory**

**C12.3.1: Classification of Solids:** Solid materials exhibit amazing range of electrical conductivity which extends over 27 orders of magnitude and probably no other physical property experiences this wide variation. One way to classify solids is based on their electrical conductivity wherein they fall into three groups : conductors, semiconductors and insulators. Metals are good conductors having conductivities of the order of  $10^7 (\Omega m)^{-1}$ . At the other extreme are insulators with low conductivities ranging between  $10^{-10}$ - $10^{-20} (\Omega m)^{-1}$ . Materials with intermediate conductivities are termed as semiconductors.

**C12.3.2: Energy Bands in Solids:** When atoms are spaced far enough, as in gas, they have little influence upon each other and are very much like isolated atoms. But the atoms in the solids have marked influence upon each other. The forces, binding the atoms together, greatly modify the behavior of electrons in other atoms. One consequence of this close proximity of atoms is to cause the perturbation of individual discrete energy levels of the atoms to form a group of closely spaced energy levels together called an energy band. The figure C12.2 shows the difference in the energy arrangement between an isolated atom and the atomic aggregate forming a solid. During the formation of solid by close and regular packing of atoms, the filled energy levels group together to form *valence band* while the empty energy levels form *conduction band*. The filled energy levels, forming the valence band, get depressed in energy while the levels comprising the conduction band get raised in energy thereby creating a *forbidden energy gap* between two bands where there are no allowed energy levels for electrons to occupy. The magnitude of this band gap determines whether a solid is conductor, semiconductor or an insulator.



**Figure C12.2:** The atomic levels in an isolated atom and a solid.

The band gap in case of conductor is either very small (less than an eV) or zero (valence and conduction bands are overlapping). In the case of insulators, the band gap is of the order of >5eV while in semiconductors it is intermediate. At zero Kelvin temperature, all the electrons are found in valence band and none in conduction band as a result of which all semiconductor materials behave as insulators and conduct no current. At higher temperatures, the excitations from valence to conduction bands are possible due to supply of thermal energy and conductivity of semiconductors increases with temperature.

### C12.3.3: Temperature Variation of Resistivity of a Semiconductor:

The conductivity of a semiconductor is given as:

$$\sigma = (n \mu_e + p \mu_p) |e| \quad (C12.1)$$

where  $n$  and  $p$  electron and hole densities respectively while  $\mu_e$  and  $\mu_h$  are electron and hole mobility respectively. The mobility is a measure of acceleration suffered by the charge carrier per unit strength of the applied external field before colliding with a thermally vibrating crystal atom or being scattered by the Coulomb forces due to dopant impurity. These collisions impede the mobility and cause resistive heating of the crystal.

For an intrinsic semiconductor  $n=p=n_i$  where  $n_i$  is the intrinsic charge carrier density. The conductivity of intrinsic semiconductor takes the form:

$$\sigma = (\mu_e + \mu_p) n_i |e| \quad (C12.2)$$

For a p-type semiconductor  $p \gg n$  therefore we get conductivity as:

$$\sigma = p \mu_p |e| \quad (C12.3)$$

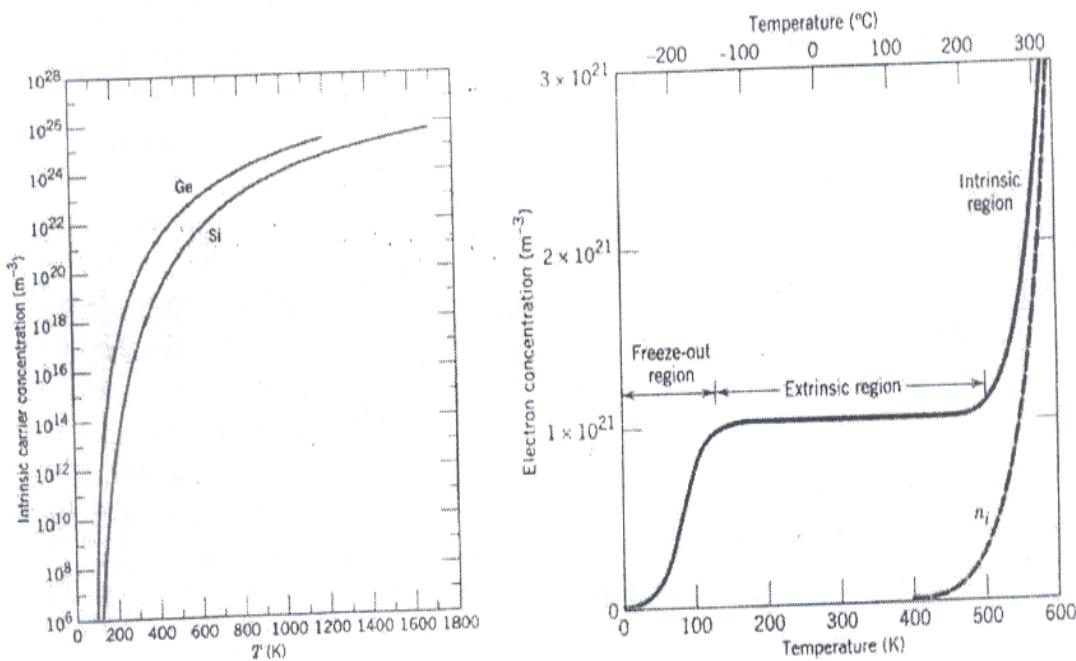
For an n-type semiconductor  $n \gg p$  therefore we get conductivity as:

$$\sigma = n \mu_e |e| \quad (C12.4)$$

Thus we find that the resistivity  $\left( \rho = \frac{1}{\sigma} \right)$  of a semiconductor depends upon two factors, the concentration and mobility of charge carriers, both of which are temperature dependent. The electron and hole mobilities decrease in magnitude with rise in temperature due to enhanced thermal scattering of the carriers at high temperature. The mobility is however weakly dependent on temperature and  $\mu_e > \mu_h$  at any temperature. While the mobility has a mild temperature effect on

conductivity, the carrier density variation has a strong effect. The n and p values can change by order of magnitudes.

The variation of carrier density with temperature for Si and Ge semiconductors is shown in the part (a) of the figure C12.3 below. The carrier concentration increases with rise in temperature because of enhanced thermal excitations of electrons from valence to conduction band. At any temperature  $n_i$  for Si is smaller than that for Ge because of smaller band gap of Ge.



**Figure C12.3:** The variation of intrinsic charge carrier concentration with temperature.

On the other hand, the variation of charge carrier concentration with temperature for an extrinsic semiconductor is much different. The electron concentration vs temperature for Si doped with  $10^{21}/\text{m}^3$  phosphorous atoms is shown in part (b) of the figure C12.4. Note that extrinsic curve has three regions which are discussed below:

At very low temperature below 100K, electron concentration is very low and almost zero at 0K. Over this range, the thermal energy is just not sufficient to excite electrons from the donor level into

conduction band. This is known as *freeze out region* as charge carriers are frozen to the position of the dopant atoms. However the electron concentration rises with temperature which implies that increasing number of donor atoms release their electron into the conduction band.

At intermediate temperatures between 150 – 450K, the material is n type and electron concentration remains constant. This is known as *extrinsic temperature region*. Electrons are excited from the donor phosphorous level to the conduction band. In this temperature range all the phosphorous atoms have donated their extra electrons and electron concentration is same as that of phosphorous dopant concentration. Also the intrinsic excitations in this region are insignificant in comparison to the extrinsic donor excitations.

Finally at higher temperatures, the electron concentration increases above the phosphorous dopant concentration and approaches the intrinsic curve with rise in temperature. This is known as the *intrinsic temperature region* since at these higher temperatures the semiconductor behaves as intrinsic i.e. charge carrier concentration resulting from electron excitations across the band gap becomes equal to and then completely overwhelm the donor carriers. Thus including the temperature dependence of mobility and carrier concentration, in total the temperature dependence of resistivity is expressed as:

$$\rho = A e^{\frac{E_g}{2KT}} \quad (C12.5)$$

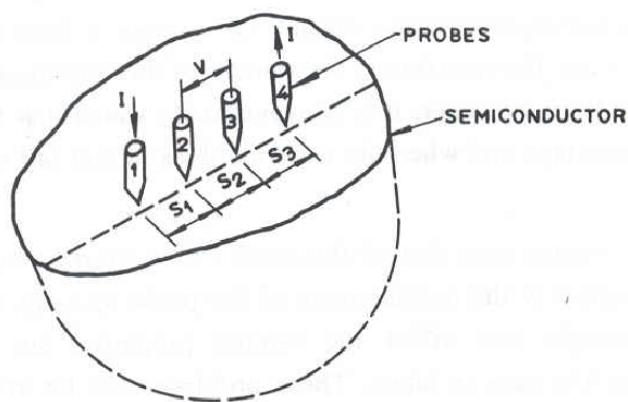
where A is a constant.

**C12.3.4: Four Probe Method:** Many conventional methods of resistivity measurements are unsatisfactory for semiconductors because of the following reasons:

- (a) Metal-semiconductor contacts are rectifying in nature.

- (i) There is general minority carrier injection by one of current carrying contacts. An excess concentration of minority carriers will affect the potential of other contacts and modulate the resistance of material.

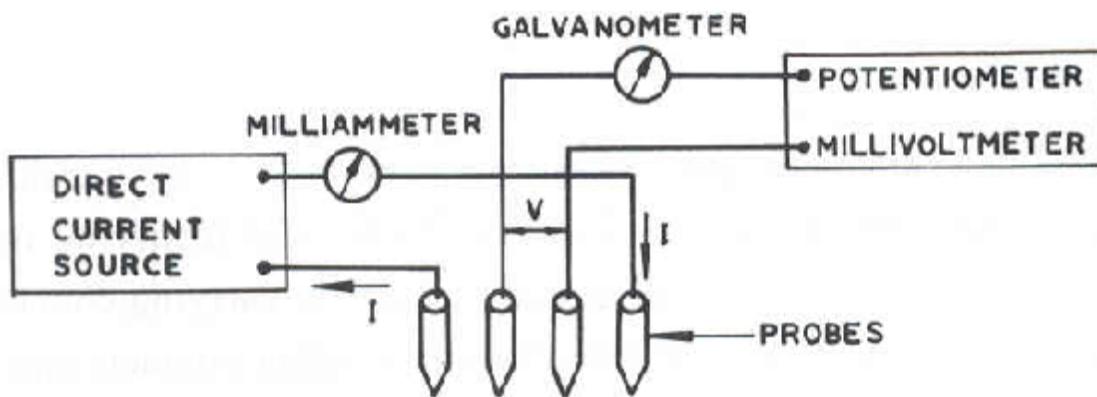
The method described here overcomes the above-mentioned difficulties and offers many advantages. It permits measurement of resistivity of a variety of shapes of sample including the resistivity of small volumes within a bigger piece of semiconductor. In this manner the resistivity of both sides of pn junction can be determined with good accuracy before the material is cut into bars for making devices. This method of measurement is also applicable to Silicon and other semiconductor materials. The basic model for all these measurements is indicated in the figure C12.4 shown below.



**Figure C12.4:** The four probe arrangement placed on the surface of a semiconductor.

Four sharp probes are placed on flat surface of the material of the sample, current is passed through the two outer probes and the floating potential is measured across the inner pair of probes. If the flat surface, on which the probes rest, is large and the crystal is big in size, it may be considered to be semi-infinite volume. To prevent the minority carrier injection and make good contacts, the surface on which probes rest may be mechanically lapped. The experimental circuit used for measurements is shown in the figure C12.5. A nominal value of probe spacing which has

been found to be satisfactory is an equal distance of 2.0mm between adjacent probes. This permits measurements of resistivity for n-type or p-type semiconductor from 0.001 to 50 $\Omega$ cm.



**Figure C12.5:** The circuit diagram for the resistivity measurements by four probe method.

In order to use four probe method in semiconductor crystals or slides, it is necessary to assume that:

- (a) The resistivity of the material is uniform over the area of measurement.
- (b) If there is a minority carrier injection into the semiconductor by the current carrying electrode most of the carriers recombine near the electrodes so that their effect on the conductivity is negligible. This implies that measurements should be made on the surface where recombination rate is very high, such as a lapped surface.
- (c) The surface on which probes rest is flat with no surface leakage.
- (d) The four probes used for resistivity measurements make contact with the surface at points which lie on a straight line.
- (e) The diameter of the contact between the metallic probes and the semiconductor must be much smaller than the distance between adjacent probes.
- (f) The boundary between the current carrying electrodes and the bulk material is hemispherical and small in diameter.

- (g) The surface of the semiconductor crystal may be either conducting or non-conducting. A conducting boundary is one on which the material of lower resistivity than semiconductor is plated. A non-conducting boundary is produced when the surface of the crystal is in contact with an insulator.

In this experiment, the bottom surface is non-conducting and also a given sample can be assumed to be of infinite area. Therefore assuming  $W \ll S$ , the resistivity is given as:

$$\rho = \rho_0 = \left(\frac{V}{I}\right) \frac{\pi W}{\ln 2} \quad (C12.6)$$

#### C12.4: Procedure

1. Put the sample on the base plate of the four probe arrangement. Unscrew the pipe constituting four probes and let the four probes rest at the middle of the sample. Apply a very gentle pressure on the probes and tighten the pipe in this position. Check the continuity between the probes for proper electrical contacts.
2. Connect the outer pair of probes leads to the constant current power supply and the inner pair of the probe to voltage terminals.
3. Place the four probe arrangement in the oven and fix the thermometer in the oven through the hole provided.
4. Switch on the mains of the four probe set-up and put the digital panel meter in the current measuring mode through the selector switch. In this position the LED facing mA would glow. Adjust the current to the desired value.
5. Now put the digital panel meter in the voltage measuring mode. In this position LED facing mV would glow and the meter would read the voltage between the probes.

6. Connect the oven power supply. Rate of heating may be selected with the help of switch LOW to HIGH as desired. Switch ON the power to the oven. The glowing LED indicates the power to the oven is ON.
7. Heat the crystal upto a temperature of  $75^{\circ}\text{C}$  and note down the voltage across the probes at regular intervals of temperature (say  $5^{\circ}\text{C}$ ) till the temperature falls to room temperature.

### C12.5: Observations and Calculations

Least count of digital ammeter =

Least count of digital voltmeter =

Least count of thermometer =

Constant current =

Distance between the probes (S) =

Thickness of crystal =

**Table C12.1:** Variation of resistivity of Ge crystal with temperature.

S.No.	Temperature (T) (in K)	$1/T (\text{K}^{-1})$	Voltage (in V)	Resistivity ( $\rho$ ) (in $\Omega\text{m}$ )	$\ln(\rho)$
1.					
2.					
3.					
4.					
5.					
6.					
7.					
8.					

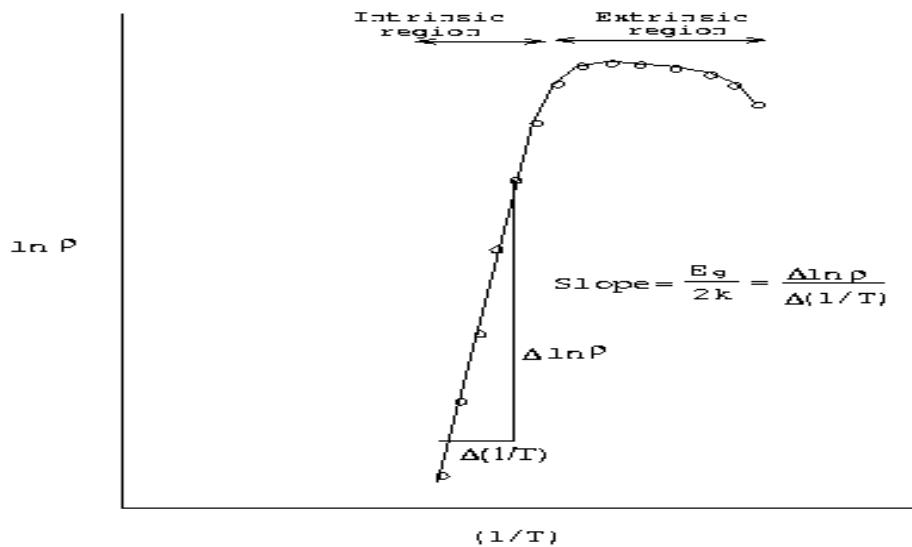
9.					
10.					
11.					
12.					
13.					
14.					
15.					

Using equation (C12.6) we can write

$$\ln \rho = \ln A + \frac{E_g}{2KT} \quad (\text{C12.7})$$

which is of the form  $y=mx+c$  and hence represents a straight line. Thus the graph is plotted

between  $\ln \rho$  and  $\frac{1}{T}$ . Its slope is given as:



**Figure C12.6:** The plot of  $\ln \rho$  vs  $1/T$  showing different regions of conductivity of a semiconductor.

$$\text{slope} = \frac{E_g}{2k}$$

$$E_g = \text{slope} \times 2k \quad (\text{C12.8})$$

It should be noted that equation (C12.6) is valid only in the intrinsic region and hence slope should be determined in this region.

## C12.6: Precautions and Sources of error

1. The Ge crystal is very brittle. Therefore use only the minimum pressure required for maintained proper electrical contacts.
2. The current through the probes 1 and 4 should be large enough so as to obtain a reasonable value of voltage across the probes 2 and 3.
3. The surface of Ge crystal may be cleaned with alcohol to remove any contamination if present.

## C12.8: Questions for Viva:

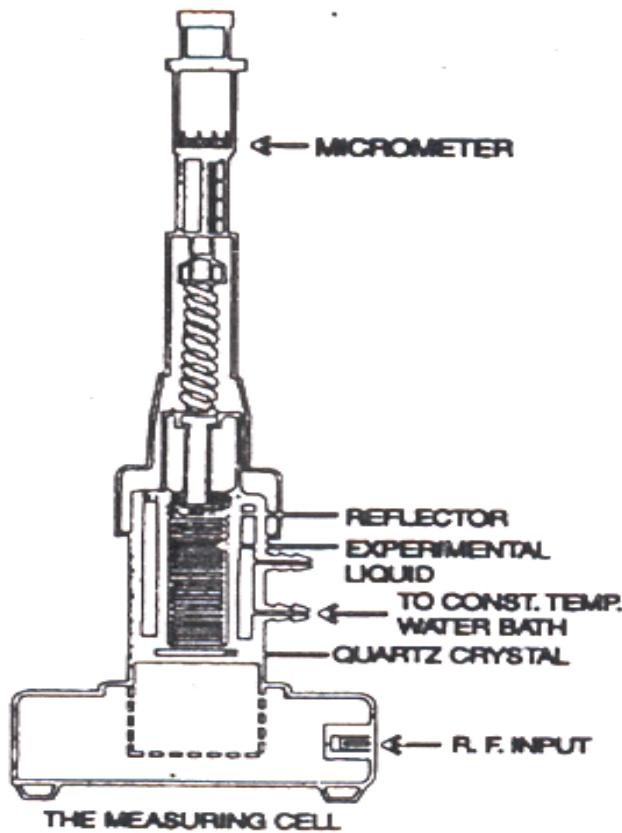
1. What is Four probe method used for?
2. How is four probe method better than 2 probe method?
3. Why are 4 probes used in four probe method?
4. Why is the current kept constant for measuring the resistivity of a semiconductor using four probe at different temperatures?
5. How do the conductivities of metals and semiconductor depend on temperature?
6. How does the band gap of a semiconductor vary with temperature?

## C13: ULTRASONIC INTERFEROMETER

**C13.1: Objective:** To determine the velocity of ultrasonic waves in different liquids using ultrasonic interferometer.

### C13.2: Apparatus

**C13.2.1.: Ultrasonic Interferometer:** The schematic picture of the ultrasonic interferometer is shown in the figure B9.1. It consists of the following parts:

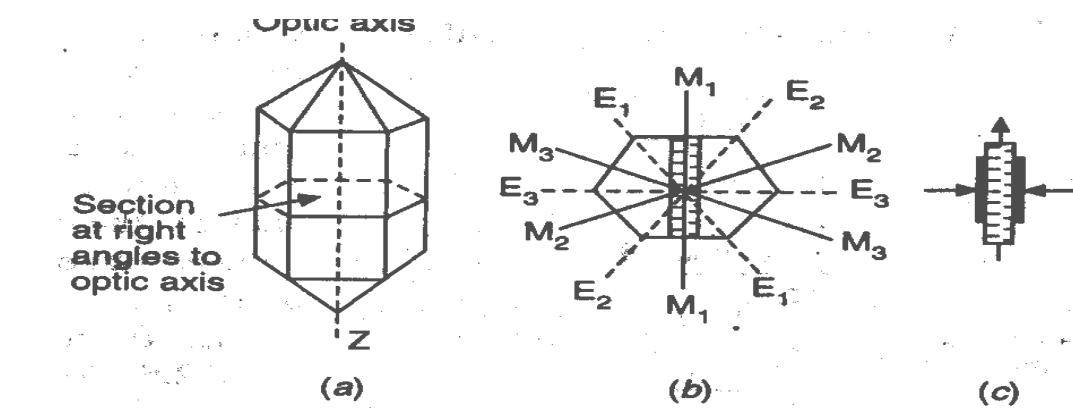


**Figure C13.1:** The schematic picture of an ultrasonic interferometer.

**High Frequency Generator** is designed to excite mechanical vibrations in the annular shaped quartz crystal which is fixed at the bottom of the measuring cell. A micro-ammeter observes changes in current and two controls for the purpose of gain and initial adjustment of micro-ammeter are provided on the front panel of the high frequency generator. This high frequency generators are available in two types depending upon the capability of producing alternating voltage of fixed frequency or a variable one.

**Measuring Cell** is double walled for maintaining the temperature of the liquid constant during the experiment. A fine micrometer screw has been provided at the top, which can lower or raise the metallic reflector plate in the liquid through a known distance. It has a annular shaped quartz crystal fixed at its bottom.

**Quartz Crystal** is found in nature in the form of hexagonal prisms capped with hexagonal pyramid at both ends as shown in the figure C13.2(a). The straight line joining their pointed ends is called the optic axis of the crystal. The section of the crystal cut perpendicular to the optic axis is a hexagon and three lines passing the opposite vertices of the hexagon form the electric axes (see figure 13.2(b)). Three axes perpendicular to these electric axes and passing through the centre of opposite faces form the mechanical axes.



**Figure C13.2:** The description of quartz crystal cured to a form suitable for use in

The quartz crystal is cut in the form of a slab such that its length is parallel to one of the mechanical axes, breadth is parallel to optic axis and thickness is parallel to electric axis perpendicular to the considered mechanical axis. In this experiment, the quartz crystal is used for generating ultrasonic waves through its excitation by radiofrequency signal. This crystal is in shape of a circular annular ring with central annulus as one terminal and outer periphery as the second. These vibrations in the quartz crystal are driven by alternating electric field generated by high frequency generator and the two systems (quartz crystals as driven system and high frequency generator as driver system) resonate thereby leading to maximum power transfer from generator to crystal at resonant frequency. The mechanical vibrations of the quartz crystal causes alternate compression and expansion in the liquid column placed vertically above thereby generating ultrasonic waves in the liquid.

**C13.3: Theory of Ultrasonic Waves:****C13.3.1: Ultrasonic Waves**

Mechanical waves of frequency greater than 20kHz are called ultrasonic waves. Human ear cannot respond to these waves. The upper frequency limit of these waves is around 500MHz and is decided by the generator of these waves. Due to small wavelength and noiseless performance, these waves are preferred over audible range of frequencies in numerous applications.

**C13.3.2: Production of Ultrasonic Waves**

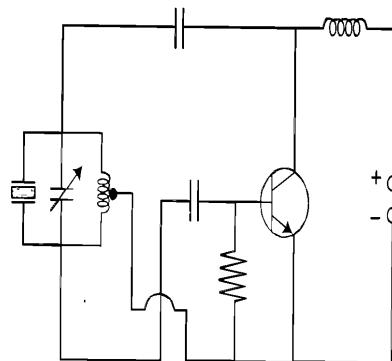
There are many methods of producing ultrasonic waves e.g., whistles, sirens, mechanical vibrators, electromagnetic transducers etc. Most commonly methods used are those involving piezoelectric

and magnetostriiction techniques. This experiment uses the piezoelectric method, which is discussed below:

**Piezoelectric Effect:** Certain crystals, when subjected to stress, develop electrostatic potential difference across a pair of opposite faces. The electrostatic potential difference is created across a pair of faces which lie perpendicular to the pair of faces subjected to mechanical stress. This phenomenon is referred to as piezoelectric effect and such crystals are piezoelectric in nature (for example quartz, rochelle salt, barium titanate). The polarity of potential generated due to compressive stress is opposite in polarity to that due to tensile stress.

In the converse phenomenon, crystal expands and contracts along one axis when an appropriate electric field is applied along an axis perpendicular to it. In a piezoelectric crystal there are three specific axes or directions; electric axis (x-axis), mechanical axis (y-axis) and optic axis (z-axis). If the electrical and mechanical stresses are applied along their respective electric and mechanical axes, then the piezoelectric effect is maximum.

**Generation of Ultrasonic Waves:** For the generation of ultrasonic waves, the electric field of definite frequency is applied to a crystal along its proper axis which causes the crystal to expand and contract alternatively resulting in generation of ultrasonic waves. The frequency of ultrasonic waves can be controlled by the alternating frequency of imposed electric field. The alternating electric field is generated by an electronic oscillator shown in the figure C13.3 below:



**Figure C13.3:** The circuit diagram of radiofrequency oscillator for applying alternating electric field on the quartz crystal.

The crystal is cut with the major surface along the electric axis and is mounted between the plates as shown in the figure C13.2. The frequency of the oscillator is controlled by LC combination.

### **Working of an Ultrasonic Interferometer**

If the alternating potential difference is applied to the crystal slab, rapid alteration of compression and extension takes place in two perpendicular directions resulting in forced vibrations of the crystal. When the frequency of the forced vibrations becomes equal to the natural frequency of the crystal, resonance occurs and amplitude of vibrations becomes sufficiently large. When the alternating voltage is applied across its thickness, there are alternating changes in its thickness and length. The resonance frequency of crystal vibrations is

$$v = \frac{1}{2d} \sqrt{\frac{E}{\rho}} \quad (C13.1)$$

The ultrasonic waves are generated by the quartz crystal of annular ring shape on which is applied an alternating electric field of certain radiofrequency. The potential difference is applied between inner and outer surface of annulus and this potential alternates its polarity at rate double that of applied electric field frequency. The ultrasonic waves propagate through the liquid filled in the measuring cell and gets reflected by the metal plate reflector. Stationary waves are set in the sample due to superposition of incident and reflected waves. In this experiment the acoustic path length is varied by movement of metal reflector till the resonance takes place. The reflector is attached to a micrometer so that the motion can be controlled precisely to obtain the position of resonance. For different positions of the reflector, the peaks observed indicate the resonance. The peaks decrease in amplitude for increase in distance from the source. They are separated by a distance of half-wavelength. For measurements in solids it is not possible to vary acoustic path

length and hence frequency is varied. The resonance peaks in this experiments are revealed by the current drawn from the radiofrequency generator which attains maximum value. The current achieves maximum value as maximum power is derived by the quartz crystal from the radiofrequency generator at resonance.

The relation between the wavelength  $\lambda$  of the stationary wave excited in the liquid column and the

distance between two antinodes (d) is  $d = \frac{\lambda}{2}$ . The velocity of ultrasonic waves in the medium is

given as

$$v = \sqrt{\frac{E}{\rho}} \quad (C13.2)$$

The frequency will be

$$f = \frac{v}{2t} \text{ Hz} \quad (C13.3)$$

The experimental value of frequency is slightly different from the theoretical value. It is because the vibrations in other direction are not considered.

### C13.3.3: Detection of Ultrasonic Waves

The ultrasonic waves can't be detected directly but some animals like bat can do so. These waves are detected by the following techniques:

**Piezo-electric Detector:** When one pair of faces of a quartz crystal is subjected to ultrasonic waves, on the other perpendicular faces, varying electric charges are produced. Though these charges are small but can be amplified and detected.

**Kundt's Tube:** When ultrasonic waves of relatively large wavelength are passed through Kundt's tube, then lycopodium powder sprinkled in the tube, collects in the form of heaps at nodal points and is blown off from anti-nodal points. This helps in detection of ultrasonic waves and measuring its wavelength.

**Acoustic Diffraction Method:** When the ultrasonic waves are propagated in a liquid, the density varies from layer to layer due to periodic variation of pressure. If under this condition, monochromatic light is passed through the liquid at right angles to the waves, then liquid behaves like a diffraction grating. The diffraction pattern of light, so obtained with this grating, can be used to find the wavelength of ultrasonic waves.

#### C13.4: Procedure

1. **Measurement of Least Count:** Study the micrometer screw arrangement carefully. Observe that it has a linear screw graduated in 0.5cm and a circular scale of 50 equally marked divisions. Adjust the circular scale's zero division to coincide with reference line of linear scale and note the reading on linear scale. Give a fixed number of complete rotations to the screw and note the reading on the linear scale. The least count of the micrometer screw will be given as:

$$LC = \frac{\text{Distance moved on linear scale}}{\text{Number of complete rotations} \times \text{Number of circular scale divisions}} \quad (\text{C13.4})$$

2. The measuring cell is connected to the output terminal of the high frequency generator through a shielded cable. The cell is filled with the experimental liquid before switching on the generator.
3. The radiofrequency generator is switched on. Wait for 5 minutes to give enough time for circuit to get sufficiently warmed up before measurements are started.
4. The ultrasonic waves propagate normal to the quartz crystal and reflected by the movable metallic reflector plate. The standing waves are formed in the liquid between reflector plate and the quartz crystal.
5. Rotate the micrometer screw slowly and observe the range of micro-ammeter needle. Using **Gain** and **Adjustment** knobs, obtain the range of deflection to be 40-80 $\mu$ A. The adjustments be

made in such a manner that the needle should neither reach zero nor maxima position so as to avoid its sticking at either of the ends

6. Rotate the micrometer slowly and note the positions where the current in the radiofrequency generator shows maxima. The maxima in the current is as a result of sudden increase in power absorbed by the quartz crystal at resonance frequency.
7. A number of maxima readings of the anode current are passed and position of each one of them is noted through the micrometer screw reading. The total distance ( $d$ ) moved by the micrometer is related to the wavelength ( $\lambda$ ) of the ultrasonic waves as:

$$d = \frac{n\lambda}{2} \quad (\text{C13.5})$$

The velocity of ultrasonic waves generated is given by the relation:

$$v = v\lambda \quad (\text{C13.6})$$

where  $v = 2\text{MHz}$  is the frequency of vibration of the quartz crystal.

8. Repeat the same exercise with the tap water and calculate velocity of ultrasonic waves in the tap water.
9. Conclude the variation of the velocity of ultrasonic waves with density of liquids.

### C13.5: Observations

**Table C13.1:** Tabulation of position of current maxima.

Least count of micrometer screw =

Least count of micro-ammeter =

Liquid used =

No. of Current Maxima (N)	Position of Current Maxima ( $x_N$ ) (in mm)	$d_n = x_N - x_{N-1}$

1		
2		
3		
4		
5		
6		
7		
8		
9		
N <sub>max</sub> =10		

$$d = \frac{\sum_{i=1}^{N_{max}-1}}{N_{max}-1} = \frac{\lambda}{2} \quad (C13.7)$$

The velocity of ultrasonic waves is given as:

$$v = (2 \times 10^6 \times \lambda) \text{ m/s} \quad (C13.8)$$

Velocity of ultrasonic waves in tap water = ..... m/s

Velocity of ultrasonic waves in alcohol = ..... m/s

### C13.6: Precautions

1. Do not switch on the generator without filling the experimental liquid in the cell.
2. Give generator 5 minutes time before starting to take observations.

3. Remove the experimental liquid after use. Keep it clean and dry.
4. Keep micrometer open at position marked with 25mm after use.
5. While cleaning the cell, let it dry out by itself and do not wipe with cloth or any other thing inside.

### C13.7: Some Related Activities

1. To study the dependence of the velocity of ultrasonic waves on density of different organic liquids using ultrasonic interferometer.
2. To study the dependence of the velocity of ultrasonic waves on temperature of liquid using the ultrasonic interferometer.
3. To measure the bulk modulus of elasticity of a liquid by measurement of velocity of ultrasonic waves by using ultrasonic interferometer.

### C13.8: Questions for Viva:

1. What do you understand by ultrasonic waves? How does it differ from the sound waves?
2. List various methods to generate and detect ultrasonic waves.
3. What do you understand by piezoelectric effect? Explain its role in the production of ultrasonic waves.
4. Explain the unique features of quartz crystal which are responsible for its piezoelectric behavior.
5. Explain the variation in the speed of ultrasonic waves with change in medium properties.
6. What do you understand by standing waves? Explain their formation in the ultrasonic interferometer.
7. What do you understand by interference? Explain its role in the ultrasonic interferometer.
8. Explain the reason of fluctuations (between maxima and minima) in the current displayed on the ultrasonic interferometer.
9. List various applications of ultrasonic waves.

