Sudan University of Science and Technology College of Graduate Studies

Reachability and Controllability for Time - Invariant Control System with physical Applications
قابلية الوصول والتحكمية لزمن انظمة التحكم اللا متغير مع التطبيقات الفيزيائية

# A Thesis submitted in Fulfillment Requirements for the Degree of Ph.D in Mathematics 

By:

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## Dedication

To My parents.
Who always support
To My Brothers and My Sister
Who always encourage me
To my best friends
Who always stand beside and help me
To Classmates (My batch),
Who don't find similar to it

## Acknowledgments

First, I thank Allah for guiding me and taking care of me all the time. My life is so blessed because of hi
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#### Abstract

The controller is an element which accepts the error in some form and decides the proper corrective action. The block diagrams is shorthand pictorial representation of control system of the cause-and-effect relationship between input and output of physical system or dynamic system is time invariant, is shifting the input of the time axis leads to an equivalent of shifting of the output along the time with no other changes. The aim of this research is study systems that the time invariant system and controllable. The controllability conditions are exactly the same in terms of the matrices for the discrete and continuous case. So we are justified in making multiple inputs, multiple outputs. Control system is a method for using any models of no solution for transfer functions and has extended to multiple inputs, multiple outputs gained selection stability and process. It has been shown that the properties of zeros mean error and single input; single output case can be also achieved. Bounded variance amplification that was seen for the multiple inputs, multiple outputs case. A design procedure is presented to achieved ideal multiple inputs, multiple outputs control case into more sensitive modeling error which are inherent in practical process. Our application is the first solution follows from a general result on the global stabilization of controllable linear system with delay in the input by bounded control laws with a distributed term.


## الخلاصة

المتحكم هو عنصر يقبل الخطأ في شكل ما ويقرر الإجراء التصحيحي المناسب. الهخططات الصندوقية هي تمثيل مختصر لنظام التحكم لعلاقة السبب والنتيجة بين مدخلات ومخرجات النظام الديناميكي او النظام الفيزيائي، و هو الزمن الثابت، وهو توني ما يكافئ تحول اللخرجات على طول الوقت دون أي تغييرات أخرى. الهـف من هذا البحث هو دراسة الانظمة التي تعمل بنظام الزمن الثابت والسيطرة عليها. شروط التحكم هي نفسها تماما من حيث المصفوفات للحالة المنفصلة والمستمرة. لذلك لابد من التحقق في تققيم انظمة متعدية الدذلات، متعددة المخرجات. نظام التحكم هو طريقة لاستخدام أي نمازج لايمكن حلها بالدوال
 أن خصائص الأصفار تعني الخطأ والمدخل الفردي، اللخرج الفردي، ويوكن ايضا تحقيق حالة خرج واحدة. تضخيم التباين المهند الذي شو هد للعديد من المدخلات، حالة الهخرجات المتعددة. تم تقديم إجر اء التصميم لتحقيق المدخلات المتعدة المثالية، وحالة التحكم في النواتج المتعددة في خطأ النمذجة الأكثر حساسية الكامنة في معاجة العملية. إن تطبيقنا هو الحل الأول من نتيجة عامة على الاستقرار الثشامل لنظام خطي خاضع للتحكم مع تأخير في الدخلات من خلال قوانين المر اقبة الكحدة مع توزيع الفترة.

## Introduction

Control theory is an important branch of mathematics that has several applications of distinct area technology, Engineering, Economics, sociology, among others. Control theory has many applications in our life's style Control theories commonly used today are classical control theory (also called conventional .Control theory), modern control theory, and robust control theory. This book, presents comprehensive treatments of the analysis and design of control systems based, on the classical control theory and modern control theory. Automatic control is essential in any field of engineering and science automatic. Control is an important and integral part of space-vehicle systems, robotic systems, modern, manufacturing systems, and any industrial operations involving control of temperature, pressure, humidity, flow, etc. It is desirable that most engineers and scientists are. Familiar with theory and practice of automatic control all necessary background materials are included. Mathematical, background materials related to Laplace transforms and vector-matrix analysis, are presented separately in appendixes. The first significant work in automatic control was James Watt's centrifugal governor, for the speed control of a steam engine in the eighteenth century other, significant works in the early stages of development of control theory were due to Minorsky Hazen, and Nyquist among many others .Minorsky worked on, automatic controllers for steering ships and showed how stability could be determined, from the differential equations describing the system, Nyquist, developed a relatively simple procedure for determining the stability of closed-loop, systems on the basis of open-loop response to steady-state sinusoidal inputs .Hazen, who introduced the term servomechanisms for position control systems, discussed the design of relay servomechanisms capable of closely following a changing, input. Modern control theory is based on time-domain analysis of differential equation, systems. Modern control theory made the design of control systems simpler because. The theory is based on a model of an actual control system. However, the system's, stability is sensitive to the error between the actual system and its model. This means that when the designed controller based on a model is applied to the actual. System, the system may not be stable. To avoid this situation, we design the control. System by first
setting up the range of possible errors and then designing the controller.in such a way that, if the error of the system stays within the assumed, range, the designed control system will stay stable. The design method based on this. Principle is called robust control theory. This theory incorporates both the frequency response, approach and the time-domain approach. The theory is mathematically very, complex.

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## Chapter one

## Introduction

## 1. Background

The concept of a control system is to sense deviation of the control of the output from the desired value and correct it till the desired output is achieved. The deviation of the actual output from its desired value is called. An error the measurement of error is possible because of feedback. The feedback allows us to compare the actual output with its desired value to generate the error. The error is denoted by as e (t) the desired value of output is also called reference input or a set point the error obtained is required to be analyzed to take the proper corrective action. The controller is an element which accepts the error in some from and decides the proper corrective action. The output of the controller is then applied to the process or final controller element this brings the output back to its desired set point value. The controller is the beart of a control system. The accuracy of the entire system depends on how sensitive is the controller has its own logic to handle the error. Now it is manipulating such an error the controllers such as microprocessors microcontrollers, computers are used such controllers execute certain algorithm to calculate the manipulating sign

## Definition of input 1.1

The input is the stimulus excitation or command Applied to a control system typically from an external energy source.

## Definition of output 1.2

The output is the actual response obtained from a control system .It may or may not be equal to the specified response implied by the .Inputs and output can have many different forms inputs for examples may be physical variables or more abstract quantities such as reference, set point or desired values for the output of the control system.

## Definition of open loop control 1.3

An open loop control system is one in which the control action is independent of the output.

## Definition of closed loop control 1.4

A closed loop control system is one in which the control action is somehow dependent on the output closed loop control systems are more commonly called feedback control system.

## Definition of feedback control 1.5

Is that property of a closed loop system which permits the output or some other controlled variable to be compared which the input to the system.

## Characteristics of feedback 1.6

The presence of feedback typically imparts the following to a system.

A- Increased accuracy.
B- Tendency toward oscillation or in stability
C- Reduced sensitive of the ratio of output to input to variation in system parameters and other Characteristics.

E- Reduced effects of external disturbances or noise
H- Increased band width
The bandwidth of a system is a frequency response measure of how well the system responds to (or filters) variation or frequencies the input signal.

### 1.1 Control system models or representations

To solve a control systems problem, we must put the specification or description of the system configuration and its components into a form amenable to analysis or design. Three basic representations (models) of components and systems are used extensively in the study of control systems.

1- Mathematical models in the form of differential equations difference equations and or other mathematical relation for example Laplace and Z-transforms.

2- Block diagrams.
3- Signal flow graphs.

### 1.1.1 Block diagrams:Fundamentals

A block diagram is a shorthand pictorial representation of the cause and effect relationship between the input and output of a physical system .The simplest form of the block diagram is single block with one input and one output


Figure (1.1) block diagram single input and single output

The operation of addition and subtraction have a special representation the block becomes a small circle called a summing point with the appropriate plus or minus sign associated with the arrows entering the circle the output is the algebraic sum of the inputs any number enter a summing point for examples


Figure (1.2) addition




Figure (1.5) some authors put a cross in the circle


### 1.1.2 Error

The error detectors compare the feedback signal $b(t)$ with the reference input $r(t)$ to generate an error.

$$
\therefore e(t)=b(t)
$$

$$
\begin{equation*}
-r(t) \tag{1.1}
\end{equation*}
$$

This gives absolute indication of an error the range of the measured variable $b(t)$

Thus span $=b_{\text {max }}-b_{\text {min }}$
The Hence error can be expressed as

$$
\begin{array}{r}
e_{P}=\frac{r-b}{b_{\max }-b_{\min }} \\
\times 100 \tag{1.2}
\end{array}
$$

Where $e_{P} \equiv$ error as $\%$ of span

## 1.Example1

The range of measured variable for ascertains control system is 2 mv to 12 mv and a set point 7 mv . Find the error as percent of span when the measured variable is 6.5 mv .

$$
\begin{aligned}
& b_{\max }=12 \mathrm{mv} \quad, b_{\min }=2 \mathrm{mv} \quad, b=6.5 \mathrm{mv} \quad r=7 \mathrm{mv} \\
& e_{P}=\frac{r-b}{b_{\max }-b_{\min }} \times 100=\frac{7-6.5}{12-2} \times 100=5
\end{aligned}
$$

### 1.1.3 Variable Range

In practical systems, the controlled variable has a range of values within.
Which the control is required to be maintained. This range specified as the maximum and minimum values allowed for the controlled variable. It can be specified as some nominal values and plus minus tolerance allowed about this value such range is important for the design of controllers

### 1.1.4 Controller Output Range

Similar to the controller variable a range is associated with a controlled output variable and minimum values. But often the controller output is expressed as a percentage where minimum controller output is $0 \%$ and maximum controller out is $100 \%$ but $0 \%$ controller output does not mean zero output. For example it is necessary of the system that a steam flow
corresponding to $\left(\frac{1}{4}\right)^{t h}$ opening of the values should be minimum Thus $0 \%$ controller output in such case corresponds to the $\left(\frac{1}{4}\right)^{t h}$ opening of the value. The controller output as a percent of full scale when the output changes within the specified range is expressed as
$p=\frac{U-U_{\min }}{U_{\max }-U_{\text {min }}} \times 100$
Where
$p \equiv$ Controller output as a percent of full scale
$U \equiv$ Value of the output
$U_{\max } \equiv$ Maximum value of controlling variable
$U_{\text {min }} \equiv$ Minimum value of controlling variable

### 1.1.5 Control Lag

The control system can have a large associate with it, the control lag is the time required by the process and controller loop to make the necessary changes to obtain the output at its set point the control lag must be compared with the process lag while designing the controllers for example. In a process value is required to be open or closed for corresponding the output variable physically the of opening. Or closing of the value is very slow and is the part of the process lag. In such a case there is no point in designing a fast controller than the process lag.

### 1.1.6 Dead Zone

Many a times a dead zone is associated with a process control loop the time corresponding to dead zone is called dead time. The elapsed between the instant when error occurs and instant when first corrective action occurs is called dead time .Nothing happens the error occurs this part is also called dead hand the effect of such dead time must be considered while the design of the controllers.

## 1.2 classifications of the controllers

The classification of the controllers is based on the response of the controllers and mode of response of the controller

### 1.2.1 Discontinuous controller Mode

The discontinuous mode controllers are further classified as ON, OFF controllers and multiposition controllers. For example in a simple temperature control of a room the heater is to be controlled it should be switched on or off by the controller when temperature crosses its set point. Such an operation and the mode of operation is called discontinuous mode of controller but in some process control systems simple on/off decision is not sufficient for example controlling the steam slow by opening or closing the value in such case a smooth opening or closing of value is necessary. The controller in such a case is said to be operating in a continuous mode thus the controllers are basically classified discontinuous controllers

### 1.3.1 Continuous Controller Mode

The continuous mode controllers are further classified as derivative controllers. Some continuous mode controllers can be combined to obtain
composite controller mode. For example of such composite controllers are PI, PD and PID controllers. The most of the controllers are placed in the forward path of control system but in some cases input to the controller is controlled though a feedback path. The example of such a controller is rate feedback controller but in the continuous controller output smoothly proportional of the error or proportional to some form of the error. Depending upon which form of the error is used as the input to the controller to product the continuous controller output these controllers are classified as. Proportional control mode, Integral control mode and Derivative control mode

### 1.3.1.1 Proportional Control Mode

In this control the output of control is simple proportional to the error $\mathbf{e}(\mathbf{t})$ the relation between the error $\mathbf{e}(\mathbf{t})$ and controller output P is determined by constant called proportional gain constant denoted by as $K_{p}$, The output of the controller is a linear function of the error e ( t ). Thus each value of the error has error has a unique value of the controller output. The range of the error which covers $0 \%$ to $100 \%$ controller output is called proportional hand. Now though there exists linear relation between controller output and the error for a zero error the controller output should not be zero otherwise the process will come to halt. Hence there exists some controller output $P_{0}$ for the zero error. Hence mathematically the proportional control mode is expressed as
$P(t)=k_{p} e(t)+P_{0}$
Where $\quad k_{p}=$ Proportional gain constant

$$
p_{0}=\text { controller output with zero error }
$$



Figure (1.7) the error may be positive or negative the proportional hand

The error may be positive or negative the proportional hand is mathematically defined by
$P_{B}=\frac{100}{k_{p}}$

### 1.3.1.2 Characteristic of proportional control mode

The various characteristics of the proportional mode are
1-When the error of zero, the controller output is constant equal to $P_{0}$
2-If the error occurs, then for every $1 \%$ of the error the correction of the $K_{p}$ $\%$ is achieved if the error is positive $K_{p}$ correction gets added to $P_{0}$ and if error is negative $K_{p} \%$ correction gets subtracted from $P_{0}$

3-The band of error exists for which the output of the controller is the between $0 \%$ to $100 \%$ without saturation

4-The gain $P_{0}$ and error band $P_{B}$ are inversely proportional to each other

### 1.3.2 Offset

The major disadvantage of the proportional control mode is that it produces an offset error in the output When the load changes the output deviates from the set point such a derivation is called offset error or steady state error such an offset error is the offset error depends upon the relation rate of the controller slow reaction rate produces small offset error while fast reaction rate produces large offset error. The dead time or Transfer lag present in the system further worsens the result it produces not only the large offset at the output but the time required to achieve steady state is also large. The offset error can be minimized by the large proportional gain $K_{p}$ which reduce the proportional hand. If $k_{p}$ is made very large the proportional band comes so small that .It acts as an on, off controller producing oscillations about the set point instead of an offset error. The proportional controller can be suitable where

A- Manual reset of the operating point is possible
B- Load changes are small
C- The dead time exists in the system is small


Figure (1.8) offset error in proportional mode

### 1.3.1.3 Integral Control Mode

In the proportional control mode, error reduces but cannot go to zero. It finally produces an offset an offset error it cannot adapt with the changing load conditions. To avoid this control mode is oftenly used in the control systems which is based on the history of the errors. This mode is called integral mode or reset action controller the value of the controller output $\mathrm{p}(\mathrm{t})$ is changed at a rate which is proportional to the actuating error signal $\mathrm{e}(\mathrm{t})$ mathematically it is expressed as
$\frac{d p(t)}{d t}=k_{i} e(t)$
Where $k_{i} \equiv$ constant relating error and rate the constant $k_{i}$ also integral constant integrating the above equation actual controller output at any time $t$ can be obtained as
$P(t)=k_{i} \int e(t) d t+p(0)$
$\mathrm{P}(0) \equiv$ controller output when integral action starts at $\mathrm{t}=0$

(a) Error

(b) Controller output

## Figure 1.9 Integral mode

The scale factor or constant $k_{i}$ expresses the scalling between error and the controller output thus a large value of $k_{i}$ mean that a small error produces a large rate of change of $p(t)$ and viceversa If there is positive error this is shown in the figure 1.8 controller output begins to ramp up. The input error step it can be seen that when error is positive the output $\mathrm{p}(\mathrm{t})$ ramp up


Figure (1.10) the step response of integral control mode

The step response of integral control mode is shown in figure 1.10
The integration time constant is the time taken for the out to the change by an Amount equal to the input error step this is shown in figure 1.11 it can be seen that when error is positive the output ramps up, For zero error there is
no change in the output and when error is negative the output $\mathrm{p}(\mathrm{t})$ ramps down


Figure (1.11) step response

### 1.3.1.4 Characteristics the integral mode

The integrating controller is relatively slow controller its output at a rate which is dependent on the integrating time constant until the error signal is can celled compared to the proportional control. The integral control requires time to build an appreciable output. However it continuous to act till the error signal disappears this corrects the problem of the offset error in the proportional controller. Thus for an integral mode

A- If error is zero, the output remains at a fixed value to what is was when the error became zero.

B- If the error is not zero, then the output begins to increase at a rate $k_{i} \%$ per second for every $\pm 1 \%$ for error in some case the inverse of $k_{i}$ called integral time is specified denoted as $T_{i}$
$T_{i}=\frac{1}{k_{i}}=$ integral time
It is expressed in minutes instead of second. The comparison of proportional and integral mode behavior at the time of occurrence of an error signal is tabulated below.
$\left.\left.\begin{array}{|l|l|l|}\hline \text { Controller } & \text { Initial behavior } & \text { Steady state behavior } \\ \hline \mathrm{P} & \begin{array}{l}\text { Acts Immediately } \\ k_{p}\end{array} & \begin{array}{l}\text { Offset error always } \\ \text { present larger the }\end{array} \\ k_{p} \text { smaller the error }\end{array}\right\} \begin{array}{l}\text { Acts slowly it is } \\ \text { The time integral of } \\ \text { the error signal }\end{array} \quad \begin{array}{l}\text { Error signal always } \\ \text { becomes zero }\end{array}\right\}$

It can be seen that proportional mode is more favorable at the start while the integral is better for steady state response in pure integral mode, error can oscillate about zero and can be cyclic ,Hence is practice integral mode is never used alone but combined with the proportional mode to enjoy the advantages of both the modes.

### 1.3.1.5 Derivative control Mode

In practice the error is function of time and at a particular instant it can be zero but it may not remain zero for ever after that instant. Hence some action is required corresponding to the rate at which the error is changing. Such a controller is called derivative controller. In this mode the output of the controller depends on the time rate of change of the actual errors. Hence it is also called rate action mode anticipatory action mode. The mathematical equation for the mode is
$p(t)=k_{d} \frac{d e(t)}{d t}$
Where $k_{d} \equiv$ Derivative gain constant indicates by now much $\%$ the controller output must change for every \% per sec of the change of the error. Generally $k_{d}$ is expressed in minutes. The important feature of this type of control mode signal there is a unique value of the controller output. The advantage of the derivative control action is that it responds to the rate of change of error and produce of significant correction before the magnitude of the actuating error becomes too large. Derivative control thus anticipates the actuating error initiates an early corrective action and tends to increase stability of the system improving the transient response

### 1.3.1.6 Characteristics of Derivative Control Mode

The figure 1.12 shows how derivative mode change the controller output for the various rates of the change of the error


Figure (1.12) the derivative mode changes the controller output for various rates of change of the error.

The derivative mode changes the controller output for various rates of change of the error. The controller output is $50 \%$ for the zero error when error starts increasing the controller output suddenly jumps to the higher value it further jumps to higher value for higher rate of increase of error then error becomes constant the output returns to $50 \%$ when error is decreasing Having negative slope controller output decrease suddenly to a lower value. The Various Characteristics of the derivative mode are

A- For a given rate of change of error signal there is a unique value of the controller output.

B- When the error is zero the controller output is zero.
C- When the error is constant - i-e rate of change of error is zero the controller output is zero.
D- When error is changing the controller output changes $k_{d} \%$

For even $1 \%$ per second rate change of error

## Note:

Hence it is never used along its gain should be small because faster rate of change of error can cause very large sudden change of controller output this may lead to the instability of the system.

### 1.3.1.7 Composite Control Modes

As mentionable earlier due to offset error proportional mode is not used alone similarly integral and derivative modes are also not used individually in practice. Thus to take the advantages of various modes together the composite control modes are used the various composite controller modes are.
i. Proportional integral mode (PI).
ii. Proportional Derivative mode (PD).
iii.Proportional + integral + Derivative mode (PID)

Let us see the characteristics of these three modes

## Proportional Integral Mode (PI Control Mode)

This is a composite control mode obtained by combining the Proportional mode and the integral mode. The mathematical expression for such a composite control is
$P(t)=k_{p} e(t)+k_{p} k_{i} \int_{0}^{t} e(t) d t+P(0)$
Where $P(0)=$ initial value of the output at $\mathrm{t}=0$. The important advantage of this control is that one to one correspondence of proportional mode is
available while the offset gets eliminated due to integral mode the integral part of such a composite control provides a reset of the zero error output after load change occurs.


Figure (1.13) Behavior of PI controllers

The composite PI mode completely removes the offset problems of proportional mode such a mode can be used in the system with the frequent or large load changes. But the process must have relatively slow changes in the load to prevent the oscillations.

Proportional + Derivative Mode (PD control mode)
The series combination of proportional and derivative control mode gives proportional plus derivative control mode. The mathematical expression for the PD composite control is

$$
\begin{equation*}
P(t)=k_{p} e(t)+k_{p} k_{d} \frac{d e(t)}{d t}+P(0) \tag{1.11}
\end{equation*}
$$

The behavior of such a PD control to a ramp type of the input in the figure


Figure (1.14) Behavior of PD controllers

Rise in signal due to proportional action between time $t_{1}$ and $t_{2}$ Immediate rise due derivative action

### 1.3.1.8 Behavior of PD Controller

The ramp function of error occurs at $t=t_{1}$ the derivative mode cause a step $V_{d}$ at $t_{2}$ and proportional mode cause arise of $V_{d}$ at $t_{2}$. This is for direct action PD control Proportional derivative Type of Controller. A controller in the forward path which changes the controller output corresponding to proportional plus derivative of error signal is PD controller
I.e. output of controller $=k e(t)+T_{d} \frac{d e(t)}{d t}$

$$
\begin{equation*}
k E(s)+s T_{d} E(s)=E(s)\left[k+s T_{d}\right] \tag{1.13}
\end{equation*}
$$ function of such controller is $\left[k+s T_{d}\right]$ this can be realized as shown in the

Figure (1.15) controller output of PD controller


Assuming $k=1$
We can write
$G(s)=\frac{\left(1+s T_{d}\right) w_{n}^{2}}{s\left(s+2 \zeta w_{n}\right)}$
And
$\frac{C(s)}{R(s)}=\frac{\left(1+s T_{d}\right) w_{n}^{2}}{s^{2}+S\left[2 \zeta w_{n}+w_{n}^{2} T_{d}\right]+w_{n}^{2}}$
Comparing denominator with standard form $w_{n}$ is same as in the previous p type controller

And
$2 \dot{\zeta} w_{n}=2 \zeta w_{n}+w_{n}^{2} T_{d}$ and $\dot{\zeta}=$
$\zeta+\frac{w_{n} T_{d}}{2}$

Because of this controller damping ratio increases by factor $\frac{w_{n} T_{d}}{2}$.I.e. system becomes type in nature. Now as order increase by one system relatively becomes less stable as $k_{i}$ must be designed in such a way system will remain in stable condition second order system is always stable. Hence transient response gets affected badly if controller is not designed properly. While
$k_{p}=\lim _{s \rightarrow 0} G(s) H(s)=\infty \quad e_{s s}=$
0
$k_{v}=\lim _{s \rightarrow 0} s G(s) H(s)=\infty \quad \cdot e_{s s}=0$
Hence as type is increased by one error becomes zero for ramp type of inputs steady state of system gets improved and system becomes more accurate in nature. Hence PI Controller has following Effect
i. it increase order of the system
ii. it increase type of the system
iii. Design of $k_{1}$ must be proper to maintain stability of system so It makes system relatively state error reduce tremendously for same type of inputs. In general this controller improves steady state part affecting the transient part

### 1.3.1.9 PID Type of Controller

As PD improves transient and PI improves steady combination of two may be used to improve overall time response of the system this can be realized as shown in the Figure (16)


Figure (1.16) the design of such controller is complicated in practice
The design of such controller is complicated in practice, Rate Feedback Controller (output Derivative Controller). This is achieved by feeding back the derivative of output signal internally using a tachogenerator and comparing with the signal proportional to error as shown this is called minor loop feedback compensation output of controller

$$
\begin{equation*}
k E(t)-k_{t} \frac{d e(t)}{d t} \tag{1.18}
\end{equation*}
$$

Take Laplace
$\therefore$ Output controller $k E(s)-s k_{1} c(s)$
This can be realized as shown in the lower Figure (17)


Figure (1.17) controller output PID controller

Let us study its effect on same system which is considered earlier

$$
G(s)=\frac{w_{n}^{2}}{s\left(s+2 \zeta w_{n}\right)}
$$

Time constant ( T ) is the time required by the system output to reach $63.2 \%$ of its final value during the first attempt. The equation for the actual response
$c(t)$ is

$$
\begin{gather*}
C(t)=1-\left\{\frac{e^{-\zeta w_{n} t}}{\sqrt{1-\zeta^{2}}} \sin \left(w_{d} t+\theta\right)\right\} \text { and } w_{d} \\
=w_{n} \sqrt{1-\zeta^{2}} \tag{1.21}
\end{gather*}
$$

= Damped Frequency of Oscillations and

$$
\begin{equation*}
\theta=\tan ^{-1}\left\{\frac{\sqrt{1-\zeta^{2}}}{\zeta}\right\} \tag{1.22}
\end{equation*}
$$

### 1.3.6 Steady State Error

Consider a simple closed loop system using negative feedback as shown in the Figure (18)


Figure (1.18) Steady State Error
Where
$\mathrm{E}(\mathrm{s})=$ Error signal and
$\mathrm{B}(\mathrm{s})=$ Feedback signal
$E(s)=R(s)-B(s)$ But $B(s)=C(s) H(s)$ and $C(s)=$
$E(s) G(s)(1.23)$
$E(s)=R(s)-C(s) . H(s)$
$E(s)=R(s)-E(s) G(s) H(s)$
$\therefore E(s)=\frac{R(s)}{1+G(s) \cdot H(s)}$
For non-unity Feedback
$E(s)=\frac{R(s)}{1+G(s)}$
For unity Feedback

This $\mathrm{E}(\mathrm{s})$ is error in Laplace domain and is expression in ' $\mathrm{s}^{\prime}$. We want to calculate the error value in time domain. Corresponding error will be e(t) now steady state of the system is that state which remains as $t \rightarrow \infty$

Steady state error
$e_{s s}=\lim _{t \rightarrow \infty} e(t)$
Now we can relate this in Laplace domain by using final value theorem which states that
$\left.\lim _{t \rightarrow \infty} F(t)=\lim _{t \rightarrow \infty} s F(s) \mathrm{L} \sum f(t)\right\}$

Therefore
$e_{s s}=\lim _{t \rightarrow \infty} e(t)=\lim _{s \rightarrow 0} s E(s)$
Where $\mathrm{E}(\mathrm{s})$ is $\mathrm{L}\{e(t)\}$
$e_{s s}=\lim _{s \rightarrow 0} \frac{s R(s)}{1+G(s) H(s)}$
For negative feedback system use positive sign in dominator while use negative sign in denominator if system uses positive feedback From the above expression it can be concluded that stead state error depends on, i. $R(s)$ i.e. Reference input its type and magnitude
ii. $\mathrm{G}(\mathrm{s}) . \mathrm{H}(\mathrm{s})$ i.e. open loop transfer function
iii. Dominant non linarites present if any

## 2.Example 1

The figure (1.19) shown PD controller used for the system .Determine the value of $T_{d}$ so that system will be critically damped calculate its, settling time


Figure (1.19) PD controller

$$
\begin{aligned}
& G(s)=\frac{\left(1+s T_{d}\right)^{4}}{s(s+1.6)} \quad \mathrm{H}(\mathrm{~s})=1 \\
& \therefore \frac{C(s)}{R(s)}=\frac{\frac{\left(1+s T_{d}\right)^{4}}{s(s+1.6)}}{1+\frac{\left(1+s T_{d}\right)^{4}}{s(s+1.6)}}=\frac{\left(1+s T_{d}\right) 4}{s^{2}+1.65+4 T_{d} s+4}
\end{aligned}
$$

Comparing denominator with standard form
$w_{n}^{2}=4, \quad w_{n}=2$ And
$2 \zeta w_{n}=1.6+4 T_{d}$

$$
\therefore \zeta=\frac{1.6+4 T_{d}}{4}
$$

Now system required is critically damped i.e. $\zeta=1$

$$
\therefore 1=\frac{1.6+4 T_{d}}{4} \Rightarrow 4=1,6+4 T_{d}
$$

$$
\begin{gathered}
\therefore 4 T_{d}=4-1.6=2.4 \quad \therefore=T_{d}=\frac{2.4}{4}=0.6 \\
\therefore=T_{s}=\frac{4}{2 \times 1}=2 \mathrm{sec} \text { And settling time } \frac{4}{\zeta w_{n}}
\end{gathered}
$$

## 3 .Example 1

APID controller has $k_{p}=2.0, k_{i}=2,2 \mathrm{sec}^{-1}$
$k_{D}=2 \sec \quad$ And $\quad P_{1}(0)=40 \%$
Draw the plot of controller output for error


Figure (1.20) PID controller
$k_{p}=2, \quad k_{1}=2.2 \sec ^{-1} \quad k_{D}=2 \sec$
$P_{1}(0)=40 \%$
For 0-2 $\quad E_{p}=m_{1} t$ where $m_{1}=$ slope
$E_{p}=t$ For 0-2 sec $\frac{2-0}{2-0}=1$ for 2-4 $E_{p}=m_{2} t+c_{2}$
Two point on the line are $(2,2) c_{n d}(4,-3)$
$m_{2}=\frac{y_{2}-y_{1}}{x_{2}-x_{1}}=\frac{-3-2}{4-2}=-2.5$
$\operatorname{At}(2,2)=-2.5 \times 2+c_{2}=c_{2}=7$
$\therefore E_{p}=-2.5 t+7$ For $2-4 \mathrm{sec}$
For 4-6 sec
Two points on the line are $(4,-3)$
$\therefore m_{3}=\frac{0-(-3)}{6-4}=\frac{3}{2}=1.5$
$\therefore E_{p}=1.5 t+c_{3}$
$4+c_{3} \times \operatorname{At}(4,-3)=-3=1.6$
$\therefore c_{3}=-9$
$\therefore E_{p}=+1.5 t-9$ For $4-6 \mathrm{sec}$
The mode Equation for PID controller
$P=k_{P} E_{P}+K_{P} K_{i} \int_{0}^{t} E_{P} d t+K_{P} K_{D} \frac{d E_{p}}{d t}+P_{1}(0)$
$\therefore P=2 E_{P}+4.4 \int_{0}^{t} E_{P} d t+4 \frac{d E_{p}}{d t}+40$
For 0-2
$P_{1}=2 t+4.4 \int_{0}^{t} t d t+4 \frac{d}{d t}(t)+40=2 \mathrm{t}+2.2 t^{2}+4+40$
This plotted for 0-2 sec
At the end of 2 sec
The integral term has accumulated to
$P_{1}(2)=4.4 \int_{0}^{t} t d t+40=4.4 \times\left[\frac{t^{2}}{2}\right]+40=48.8 \%$
For 2-4 sec

$$
\begin{gathered}
P_{2}=2(-2.5 t+7)+4.4 \int_{2}^{t}(-2.5 t+7) d t+4 \frac{d}{d t}(-2.5 t+7)+48.8 \\
-5 t+14+4.4\left\{-1.25\left(t^{2}-4\right)+7(t-2)\right\}-10+48.8 \\
=-5.5 t^{2}+25.8 t+13.2
\end{gathered}
$$

This is plotted for 2-4 sec
At the end of 2 sec the integral term has accumulated to

$$
\left.P_{1}(4)=44\left(-1.25\left(t^{2}-4\right)+7\right)+7\left(t^{2}-4\right)\right\}\left.\right|_{t=4}+48.8=44.4 \%
$$

For 4-6 sec
$P_{3}=2(1.5 t-9)+4.4 \int_{4}^{t}(1.5 t-a) d t+4 \frac{d}{d t}(1.5 t-a)+44.4=3 t-$
$18+4.4\left[0.75 t^{2}-\mathrm{a} \mathrm{t}\right]_{1}^{T}+4 \times 1.5+44.4=3 t-18+44\left\{0.75\left(t^{2}-16\right)-\right.$
$9(t-4)\}+6+44.4=3.3 t^{2}-36.6 t+138$
This is plotted for $4-6 \mathrm{sec}$
After 6 sec error is zero hence the out will simply be the accumulated integral response providing a constant out put
$\therefore P_{1}(6)=44\left[075\left(t^{2}-16\right)-\left.a(t-4)\right|_{t=1}+44.4=31.2 \%\right.$

The complete graph of controller output is shown in the figure (21)


Figure (1.21) graphical solution of PID controller

## 4.Example 1

Temperature control system has the block diagram given in figure
The input signal is a voltage and represents the desired temperatures $\theta_{r}$ is a unit step the integral term and
(i) $\mathrm{D}(\mathrm{s})=1$
(ii) $\mathrm{D}(\mathrm{s})=1+\frac{0.1}{s}$
(iii) $\mathrm{D}(\mathrm{s})=1+0.3 \mathrm{~s}$

What is effect of the integral term in the PI controller and derivative term in PD controller on the steady state error


For given system

$$
G(s)=\frac{200 D(s)}{(s+1)(s+2)}
$$

$H(s)=0.02$
$R(s)=\frac{0.02}{s}$
$e_{s s}=\lim _{s \rightarrow 0} s F(s)=\lim _{s \rightarrow 0} \frac{s R(s)}{1+G(s) H(s)}$
$\mathrm{D}(\mathrm{s})=1$
$\therefore e_{S S}=\lim _{s \rightarrow 0} \frac{\frac{s \times 0.02}{s}}{1+\frac{200}{(s+1)(s+2)} \times 0.02}=\frac{0.02}{1+\frac{200 \times 0.02}{2}}=6.66 \times 10^{-3}$
(ii) $\mathrm{D}(\mathrm{s})=1+\frac{0.1}{s}$
$\therefore e_{s S}=\lim _{s \rightarrow 0} \frac{s \times \frac{0.02}{s}}{1+\left[\frac{200(1+0.3 s)}{(s+1)(s+2)} \times 0.02\right]}=6.66 \times 10^{-3}$
Due to PI controller the steady state error reduces drastically while PD controller has no effect on the stead state error

## 5.Example 1

An integral controller is used for temperature control within arrange 40-60c ${ }^{\circ}$ the set point is $48 c^{o}$, the controller output is initially $12 \%$. When error is zero the integral constant $k_{1}=-0.2 \%$ controller output per second percentage error if the temperature increase $54 c^{\circ}$, Calculate the controller output After 2 sec for a constant error

For integral controller

$$
P(t)=k_{i} \int E_{p} d t+P(0)
$$

Controller output initially denoted by $P(0)=12 \%$ the integral constant denoted by $k_{1}=-0.29 \% \sec \%$ error
$E_{P}=$ Error $=$ constant $=\frac{r-b}{b_{\max }-b_{\min }} \times 100$
Now r $=$ set point $=48 c^{\circ}$
$\mathrm{b}=$ Actual temperature $=54 c^{\circ}$
$b_{\text {max }}=60 c^{0} \cdot b_{m-x}=40 c^{o}$

$$
E_{P}=\frac{48-54}{60-40} \times 100=-30
$$

$\int E_{P d t=E_{p} t}$ As error is constant
$\mathrm{P}(\mathrm{t})=(-0.2)(-30)(\mathrm{t})+12$
At $t=2=(0.2)(30 \times 2)+12=24 \%$
This controller output after 2 sec

## Example 1.5

A proportional controller is employed for the cont. of temperature in the range $50 c^{o}-130 c^{o}$ with a set point of $73.5 c^{o}$ the zero error controller output is $60 \%$ what will be the offset error resulting from a change in the controller output to $55 \%$ the proportional gain is $2 \%$ find the offset in $c^{o}$ For proportional control mode
$P=k_{P} E_{P}+P(0)$
$\mathrm{P}(0)=$ controller output with no error $=50 \%$
$k_{P}=$ Proportional gain $=2 \%$ per second

$$
E_{P}=\frac{P-P_{0}}{k_{P}}=\frac{55-50}{2}=2.5
$$

For the design of a control system it's important to understand how the system of interest behave and how it respond to deferent control design the Laplace transform as discussed in the Laplace transform module is a valuable tool that can be used to solve differential equations and obtain the dynamic

A Transfer function $\mathrm{G}(\mathrm{s})$ is defined as the following relation between the output of the system $\mathrm{Y}(\mathrm{s})$ and the input to the system $\mathrm{U}(\mathrm{s})$

$$
\begin{equation*}
\mathrm{G}(\mathrm{~s})=\frac{Y(s)}{U(s)} \tag{1.33}
\end{equation*}
$$

The roots of polynomial $\mathrm{U}(\mathrm{s})$ poles of the system and roots of $\mathrm{Y}(\mathrm{S})$ is called the zeros of the system If the input of the system is a unit impulse $(U(S)=1)$

### 1.4 Blok Diagrams and Transfer functions

The combination of Blok Diagrams and Transfer functions is a powerful way to represent the control systems. Transfer function relating different signals in the system. Derivation of transfer functions from state model.

Consider a standard state model derived for linear time invariant system as
$\dot{X}(t)=A X(t)+B U(t)$
$Y(t)=C X(t)+D U(t)$
And taking the Laplace transfer of both sides
$[S X(s)-X(0)]=A X(s)+B U(s)$ and $Y(s)=C X(s)+D U(s)$
Note that as the system is time invariant the coefficient of matrices A, B, C and D are constant while the definition of the transfer function is based on the assumption of zero initial condition
$X(0=0)$
$\therefore S X(s)=A X(s)+B U(s)$ and $S X(s)-A X(s)=B U(s)$

Now S is an operator while A is matrix of order $(\mathrm{n} \times \mathrm{n})$, hence to match the orders of two terms on lift hand side multiply "S" by identity matrix I of the same order
$\therefore S I X(s) A X(s)=B U(S)$ and $[S I-A] X(s)=B U(s)$

Multiplying both sides by $[S I-A]^{-1}$

$$
\begin{align*}
& {[S I-A]^{-1}[S I-A] X(s)=[S I-A]^{-1} B U(s)}  \tag{1.39}\\
& \therefore X(s)=[S I-A]^{-1} B U(s) \tag{1.40}
\end{align*}
$$

Substituting in the equation (1.36) we get

$$
\begin{align*}
& Y(s)=C[S I-A]^{-1} B U(s)+D U(s) \text { and } Y(s) \\
&=\left[C\left([S I-A]^{-1} B+D\right) U(s)\right]  \tag{1.41}\\
& \therefore \mathrm{T}(\mathrm{~s})=\frac{Y(s)}{U(s)}=C[S I-A]^{-1} B+D \quad \text { But }[S I-A]^{-1}=\frac{A d j[S I-A]}{|S I-A|} \tag{1.42}
\end{align*}
$$

The state model of a system is not unique but the transfer function of obtained from any state model is unique. It is independent of the method used to express the system in state model form.

### 1.4.1 Characteristic equation

It is seen from the excretion of transfer function that the denominator is $[S I-A]$ the equation obtained by equating denominator of transfer function to zero is called characteristic equation the root of this equation are the closed loop poles of the system thus the characteristic equation of the system is $[S I-A]=0$ The stability of the system depends on the roots, the roots of the equation $[S I-A]=0$ are called Eigen values of matrix A and this are generally denoted by $\lambda$

## 7 .Example1

Consider a system having state model
$\left[\begin{array}{l}x_{1}{ }_{1} \\ x_{2}\end{array}\right]=\left[\begin{array}{cc}-2 & -3 \\ 4 & 2\end{array}\right]\left[\begin{array}{l}X_{1} \\ X_{2}\end{array}\right]+\left[\begin{array}{l}3 \\ 5\end{array}\right] U$ And
$\mathrm{Y}=\left[\begin{array}{ll}1 & 1\end{array}\right]\left[\begin{array}{l}X_{1} \\ X_{2}\end{array}\right]$
With $\mathrm{D}=0$ obtain
$[S I-A]^{-1}=\frac{\operatorname{Adj}[S I-A]}{|S I-A|}$
$\lceil S I-A\rceil=\left[\begin{array}{cc}S+2 & 3 \\ -4 & S-2\end{array}\right]$
$\operatorname{Adj}=\left[\begin{array}{ll}C_{11} & C_{12} \\ C_{21} & C_{22}\end{array}\right]^{T}=\left[\begin{array}{cc}S-2 & 4 \\ -3 & S+2\end{array}\right]^{T}=\left[\begin{array}{cc}S-2 & -3 \\ 4 & S+2\end{array}\right]$
$|s I-A|=(S+2)(S-2)+12=S^{2}-4 S+12=S^{2}+8$

$$
[s I-A]^{-1}=\frac{\left[\begin{array}{cc}
S-2 & -3 \\
4 & S+2
\end{array}\right]}{s^{2}+8}
$$

$\mathrm{T} . \mathrm{F}=\frac{\left[\begin{array}{cc}1 & 1\end{array}\right]\left[\begin{array}{cc}S-2 & -3 \\ 4 & S+2\end{array}\right]\left[\begin{array}{l}3 \\ 5\end{array}\right]}{S^{2}+8}=\frac{[8 S+1]}{S^{2}+8}$.


Figure (1.23) product of the transfer functions.

The transfer function of the system is $G=G_{1} G_{2}$ ie the product of the transfer functions.


Figure (1.24) a parallel connection of systems

Consider a parallel connection of systems with the transfer functions $G_{1}$ and $G_{2}$. Let $u=e^{s t}$ be the input to the system the pure exponential output of first system $y_{1}=G_{1} u$ and the output of the second system is $y_{2}=G_{2} u$

The pure exponential output of the parallel connection is thus
$y=G_{1} u+G_{2} u=\left(G_{1}+G_{2}\right) u$
The transfer function of the parallel connection is thus

$$
G=G_{1}+G_{2}
$$



Figure (1.25) feedback connection of systems
Consider a feedback connection of systems with the transfer functions $G_{1}$ and $G_{2}$. Let $r=e^{s t}$ be the input to the system $y$ the pure exponential output and be the pure exponential part of the error. Writing the relations for the different blocks and a summation unit we find

$$
y=G_{1} e \quad \text { and } e=r-G_{2} y
$$

Elimination of $e$ gives $y=G_{1}\left(r-G_{1} y\right.$.Hence $\left(1+G_{1} G_{2}\right) \mathrm{y}=G_{1} r$
Which implies $y=\frac{G_{1}}{1+G_{1} G_{2}} \cdot r$.The transfer function of the feed Feedback connection is thus

$$
G(s)=\frac{G_{1}(s)}{G_{1}(s)+G_{2}(s)}
$$



Figure (1.26) Closed loop transfer function feed back

$$
\begin{gather*}
y_{c l}=G_{p} G_{c}\left(R-y_{c l}\right), \quad T_{(s)}=\frac{y_{c l}(s)}{R(s)} \\
R G_{p} G_{c}=y_{c l}\left(G_{p} G_{c}+1\right) \quad, y_{c l}=R \frac{G_{p} G_{c}}{1+G_{p} G_{c}} \\
\therefore T(s)=\frac{G_{p} G_{c}}{1+G_{p} G_{c}} \tag{1.43}
\end{gather*}
$$

Figure (1.27) Closed loop transfer function with a senor transfer function

$$
y=G_{p} G_{c}\left(R-y(s) H(s) \Rightarrow y=G_{p} G_{c} R-G_{p} G_{c} H y\right.
$$

$\left(1+G_{p} G_{c} H\right) y=G_{p} G_{c} R$
$\therefore T(s)=\frac{y(s)}{R(s)}=\frac{G_{p} G_{c}}{1+G_{p} G_{c} H}$

### 1.5 Signal flow graphs

A signal flow graph is pictorial representation of the simultaneous equations describing a system. It graphically displays the transmission of the signal through the system as does the block diagram. But it is easier draw and therefore easier to manipulate than the block diagram. The properties of signal flow graphs are represented

### 1.5.1 Fundamental of signal flow

Let us first consider the simplest equation

$$
\begin{equation*}
x_{i}=A_{i j} x_{j} \tag{1.45}
\end{equation*}
$$

The variables $x_{i}$ and $x_{j}$ can be functions of time complex frequency or other quantity they many even constants which are variables in mathematical sense.

For signal flow graphs $A_{i j}$ is a mathematical operator mapping $x_{i}$ into $x_{j}$ and is called the transmission function. For example $A_{i j}$ may be a constant in which case $x_{i}$ is a constant times $x_{j}$ in equation (1.45) if $x_{i}$ and $x_{j}$ are functions of s or z $A_{i j}$ may be transfer function $A_{i j}$ s and $A_{i j} z$ the signal flow graph for equation (1.45)is given in figure 1this simplest form of a signal flow graph .Note that the variables $x_{i}$ and $x_{j}$ are represented by a small dot called anode and the transmission function $A_{i j}$ is represented by a line with arrow called a branch


Figure (1.28)
Every variable in a signal flow graph is designated by anode and every transmission function by a branch. Branches are always unidirectional. The arrow denotes the direction of signal flow graph

## Example 1.8

Ohm,s law state that $E=R I$ where E is a voltage I a current and R a resistance the signal flow graph for this equation is given by


Figure (1.29)

## The addition rule

The value of the variable designated by anode is equal to the sum of the all signals entering the node in the other words the equation represented by

$$
\begin{equation*}
x_{i}=\sum_{j=1}^{n} A_{i j} x_{i} \tag{1.46}
\end{equation*}
$$



## The Transmission Rule

The value of the variable designated by anode is transmitted on every branch leaving that node in other words the equation

$$
\begin{array}{rl}
x_{i}=A_{i k} x_{k} & i \\
& =1,2, \cdots, n, k \text { fixed } \tag{1.47}
\end{array}
$$

Is represented by


## Example 1.9

The signal flow graph of simultaneous equations $Y=3 X, Z=-4 x$ is given by figure (1.32)


## The Multiplication Rule

A cascaded (series) connection of $n-1$ branches with transmission function equal to product of the old ones that is $x_{n}=A_{21} \cdot A_{32} \cdot A_{43}$.
$\cdots A_{n(n-1)} \cdot x_{1}$
The signal flow graph equivalence represented by figure


Figure (1.33)

## Example 1.10

The signal flow graph of the simultaneous equation $Y=10 x, Z=-20 x$ is the given by the by figure

which reduces to


Figure (1.34)

## Definitions 1.8

The following terminology is frequently used in signal flow graph theory the example associated with each definition refer to figure


## Definition 1.8.1

A path is a continuous, unidirectional succession of branches along which co node is passed more than once

For example $x_{1}$ to $x_{2}$ to $x_{3}$ to $x_{4}, x_{2}$ to $x_{3}$ and back to $x_{2}$ to $x_{1}$ to $x_{2}$ to $x_{4}$ are paths

## Definition 1.8.2

An input node or source is anode with only outgoing branches .For example $x_{i}$ is an input node

## Definition 1.8.3

An output node or sink is anode with only incoming branches For example $x_{4}$ is an output node

## Definition 1.8.4

A forward path is path from the input node to the output. For example $x_{1}$ to $x_{2}$ to $x_{3}$ to $x_{4}$ are forward path

## Definition1.8.5

A feedback path or feedback loop is a path which originates on the same node

For example $x_{2}$ to $x_{3}$ and back to $x_{2}$ is a feedback path

## Definition 1.8.6

A self-loop is a feedback loop consisting of a single branch .For example $A_{33}$ is a self-loop

## Definition 1.8.7

The gain of branch is transmission function of that branch gain encountered in traversing a path .For example the path gain of the forward from $x_{1}$ to $x_{2}$ to $x_{3}$ to $x_{4}$ is $A_{21} A_{32} A_{43}$

## Definition1.8.8

The loop gain is product of the branch gains of the loop .For example the loop gain of the feedback loop form $x_{2}$ to $x_{3}$ and back $x_{2}$ is $A_{32} A_{23}$

### 1.5.2 Construction of signal flow graph

The signal flow graph of linear feedback control system whose components are specified non interacting transfer functions can be constructed by direct reference to block diagram of the system. Each variable of the block diagram becomes anode and each block becomes branch



### 1.5.3 Block diagram Reduction using flow graphs and the General input-output gain formula

Often the easiest way to determine the control ratio of complicated block diagram into a signal flow graph

## Example 1.11

Let us determine the control ratio $\frac{C}{R}$ and the canonical block diagram of the feedback control system


Figure (1.36)
The signal flow graph is gives in figure (1.36) there are forward path

$$
P_{1}=G_{1} G_{2} G_{4}, P_{2}=G_{1} G_{3} G_{4}
$$



There are three feedback loops

$$
p_{11}=G_{1} G_{2} H_{1} \quad, P_{21}=-G_{1} G_{2} G_{4} H_{2}, P_{31}=-G_{1} G_{3} G_{4} H_{2}
$$

There are no nontouching loops and all loops touch forward paths than $\Delta_{1}=1, \Delta_{2}=1$
Therefore the control ratio is

$$
\begin{aligned}
T=\frac{C}{R}=\frac{P_{1} \Delta_{1}+P_{2} \Delta_{2}}{\Delta} & =\frac{G_{1} G_{2} G_{4}+G_{1} G_{3} G_{4}}{1-G_{1} G_{4} H_{1}+G_{1} G_{2} G_{4} H_{2}+G_{1} G_{3} G_{4} H_{2}} \\
& =\frac{G_{1} G_{4}\left(G_{2}+G_{3}\right)}{1-G_{1} G_{4} H_{1}+G_{1} G_{2} G_{4} H_{2}+G_{1} G_{3} G_{4} H_{2}}
\end{aligned}
$$

## Chapter two

## Time Invariant system

## 2. Introduction

A dynamic system is Time invariant if shifting the input on the time axis leads to on equivalent shifting of the output along the time axis with no other changes in other words a time invariant system maps a given input trajectory $u(t)$ no matter when it occurs. Many physical systems can be modeled as linear time invariant (LTI) systems very general signals can be represented as linear combinations of delayed impulses by the principle of super position the response $y[n]$ of a discrete time (TLI) system is the sum of the response to the individual shifted impulses to the individual shifted impulses marking up the input signal $\mathrm{x}[\mathrm{x}]$

$$
\begin{equation*}
Y(t-\tau)=F[u(t-\tau)] \tag{2.1}
\end{equation*}
$$

The formula above says specifically that if an input signal is delayed by some amount T so will be the output and with no other changes



Figure (2.3)


An example of a physical time varying system is the pitch response of a rocket $\mathrm{y}(\mathrm{t})$ when the thrusters are being steered by an angle $\mathrm{u}(\mathrm{t})$. You can see first that this is an inverted pendulum problem and unstable without .A closed loop controller it is time varying because as the rocket burns fuel its mass is changing and so the pitch responds differently to various inputs throughout its flight. In this case the absolute time coordinate is the time since lift off. To assess whether a system is time varying or not follow these steps replace $u(t)$ with $u(t-\tau)$ on one side of equation replace $y(t)$ with $y(t-$ $\tau)$ on the other side of the equation and then check if they are equal Hoer several examples.
$Y(t)=u(t)^{3 / 2}$
This system is clearly time invariant because it is a static map Next

## 1. Example 2

$$
y(t)=\int_{0}^{t} \sqrt{u(t)} d t
$$

Replace $\mathrm{u}(\mathrm{t})$ with $\mathrm{u}(\mathrm{t}-\tau)$ in the right hand side and it the

$$
y(t)=\int_{0}^{t} \sqrt{u\left(t_{1}-\tau\right)} d t_{1}=\int_{-t}^{t-\tau} \sqrt{u\left(t_{2}\right)} d t_{2}
$$

The left hand side is simply

$$
y(t-\tau)=\int_{0}^{t-\tau} \sqrt{u\left(t_{1}\right)} d t_{1}
$$

Clearly the right and left hand sides are different hence the system is not time invariant. As another example consider

## 2.Example 2

$$
y(t)=\int_{0}^{t} u^{2}\left(t_{1}\right) d t_{1}
$$

The right hand side becomes with the time shift

$$
\int_{t-5}^{t} u^{2}\left(t_{1}-\tau\right) d t_{1=} \int_{t-5-\tau}^{t-\tau} u^{2}\left(t_{2}\right) d t_{2}
$$

Whereas the left hand side is

$$
y(t-\tau)=\int_{t-5-\tau}^{t-\tau} u^{2}\left(t_{1}\right) d t_{1}
$$

The two sides of the defining equation are equal under a time shift t and so this system is time invariant. Time invariant also known as shift invariance describe function independence from the location of $t=0$ on the time lime By definition a time invariant describe S a function's independence from the location of $t=0$ on the time lime By definition a time invariant systems
output will shift in tine if its input shifts in time but otherwise will remain exactly the same in other words a time invariant function does not words a time invariant function does not care what time it is we describe time invariance with the following notation

$$
\begin{equation*}
\text { Given } Y(t)=f(t) \text { then } y(t-s)=f(t-s) \tag{2.2}
\end{equation*}
$$

System because is a crucial property of real system because it allows us to assume that a system will respond in a predictable manner at any time modeling time dependent systems are often highly influenced by initial condition and system definition

## 3.Example 2

Is the system $y=t+x(t)$ time invariant?

To prove whether or not the above system is time invariant we must a mathematical technique called proof by contradiction proof by contradiction is often used when two separate conditions can be tested on the same system or mathematical contract. In this case we can compare the results of the time shifted system with the solution assuming that the system is time invariant thus the shifted system is represent the original system as
$y_{1}(t)=y(t-s)+x(t-s)$
Now if we define the input function to be $x(t-s)=x_{\text {shift }}(t)$ which we may assume if and only if the system is time invariant then we represent the original system as $\quad y_{2}(t)=t+x(t-s)$

Since $y_{1} \neq y_{2}$ we may conclude that the system is not time invariant through proof by contradiction

## Example 2.4

Is the system $y=\alpha x(t)+\beta \quad$ time invariant? (Note $\alpha$ and $\beta$ are constant)

We will use of some method a gain on the new system described above
The shifted system is represented as $y_{1}(t)=y(t-s)=\alpha x(t-s)+\beta$
Again if we $\quad x(t-s)=x_{\text {shift }}(t)$ and (other by assume system the system is time in variant).Than we represent the original system as

$$
y_{2}(t)=\alpha x_{\text {shift }}(t)+\beta=\alpha x(t-s)+\beta
$$

Since $y_{1}=y_{2}$ we may conclude that the system is time invariant

### 2.1 Solving the time invariant state Equation

In this part, we show obtain the general solution of the linear time invariant system equation.We shall first consider the homogeneous case and then the nonhomogeneous case.

### 2.1.1 Solution of Homogeneous state Equations

Before we solve vector matrix differential equations .Let u review the solution of the scalar differentia
$X=a x$

In solving this equation, we may assume a solution $\mathrm{x}(\mathrm{t})$ of the form

$$
\begin{equation*}
x(t)=b_{0}+b_{1} t+b_{2} t^{2+\cdots \cdots+b_{k} t^{k}+\cdots} \tag{2.4}
\end{equation*}
$$

By substituting this assumed solution in to Equation (2.3) we obtain

$$
\begin{align*}
& b_{1}+2 b_{2} t+3 b_{3} t^{2+\cdots \cdots+k b_{k} t^{k-1}+\cdots} \\
& =\mathrm{a}\left(b_{0}+b_{1} t+b_{2} t^{2+\cdots \cdots+b_{k} t^{k}+\cdots}\right) \mathrm{x}(0) \tag{2.5}
\end{align*}
$$

If the assumed solution is to be the true solution Equation (2.5) must hold for any $t$, Hence, equating the coefficients of the equal powers of $t$, we obtain
$b_{1}=a b_{0}, 2 b_{2}=a b_{1} \quad, b_{2}=\frac{1}{2} a b_{1}=\frac{1}{2} a^{2} b_{0}, \quad b_{3}=\frac{1}{3} a^{2} b_{0}=$ $\frac{1}{3 \times 2} a^{3} b_{0} \quad, b_{k}=\frac{1}{k!} a^{k} b_{o}$

The value of $b_{0}$ is determined by substituting $t=0$ into equation (2.4), or Hence, the solution $x(t)$ can be written

$$
x(t)=\left(1+a t+\frac{1}{2!} a^{2} t^{2}+\cdots+\frac{1}{k!} a^{k} t^{k}+\cdots\right) x(0)=e^{a t} x(0)
$$

We shall now solve the vector matrix differential equation
$X^{=}=A X$

Where $\mathrm{X}=\mathrm{n}-$ vector
$\mathrm{A}=\mathrm{n} \times \mathrm{n}$ constant matrix
By analogy with the scalar case, we assume that the solution is in the form of vector power series in t , or by substituting this assumed solution in to equation (2.5) we obtain
$b_{1}+2 b_{2} t+3 b_{3} t^{2+\cdots \cdots+k b_{k} t^{t-1}+\cdots}$
$=A\left(b_{0}+b_{1} t+b_{2} t^{2+\cdots \cdots+b_{k} t^{k}+\cdots}\right)$
If the assumed solution is to be true solution, equation (2.8) must hold for all $t$. thus by equating the coefficients of like powers of $t$ on both sides $g$ equation (2.8) we obtain
$b_{1}=A b_{0}, 2 b_{2}=A b_{1} \quad, b_{2}=\frac{1}{2} A b=\frac{1}{2} A^{2} b_{0}, \quad b_{3}=\frac{1}{3} A^{2} b_{0}=$ $\frac{1}{3 \times 2} A^{3} b_{0} \quad, b_{k}=\frac{1}{k!} A^{k} b_{o}$

By substituting $\mathrm{t}=0$ into equation (2.7) we obtain
$x(0)=b_{0}$.Thus, the solution $\mathrm{x}(\mathrm{t})$ can be written as
$\left.x(t)=1+A t+\frac{1}{2!} A^{2} t^{2}+\cdots+\frac{1}{k!} A^{k} t^{k}+\cdots\right) x(0)$
The expression in the parentheses on the right hand side of this last equation is an $\mathrm{n} \times \mathrm{n}$ matrix because of the similarity to the infinite power series for a scalar exponential we
$1+A t+\frac{1}{2!} A^{2} t^{2}+\cdots+\frac{1}{k!} A^{k} t^{k}+\cdots=e^{A t}$
In terms of the matrix exponential is very important in state space in state space analysis of linear systems, we shall next examine its properties

## Definition 2.1

It can be proved that the matrix exponential of an $n \times n$ matrix $A$

$$
\begin{equation*}
e^{A t}=\sum_{k=0}^{\infty} \frac{A^{k} t^{k}}{k!} \tag{2.9}
\end{equation*}
$$

Converges absolutely for all finite t Hence, computes calculations for evaluating the dements $e^{A t}$ by using the series expansion can be easily carried out.Because of the convergence of the infinite series

$$
\begin{align*}
& e^{A t} \\
& =\sum_{k=0}^{\infty} \frac{A^{k} t^{k}}{k!} \tag{2.10}
\end{align*}
$$

The series can be differential term to give

$$
\begin{align*}
& \frac{d e^{A t}}{d t}=A+A^{2} t+\frac{A^{3} t^{2}}{2!}+\cdots+\frac{A^{k} t^{K-1}}{(k-1)!}  \tag{2.11}\\
& \quad=A\left[1+A t+\frac{A^{2} t^{2}}{2!}+\cdots+\frac{A^{k-1} t^{K-1}}{(k-1)!}+\cdots\right]=A e^{A t} \\
& =A\left[1+A t+\frac{A^{2} t^{2}}{2!}+\cdots+\frac{A^{k-1} t^{K-1}}{(k-1)!}+\cdots\right]=A e^{A t}
\end{align*}
$$

The matrix exponential has the property that

$$
\begin{equation*}
e^{A(t+S)}=e^{A t} e^{A s} \tag{2.12}
\end{equation*}
$$

This can be proved as follows

$$
\begin{aligned}
& e^{A t} e^{A s}=\left(\sum_{k=0}^{\infty} \frac{A^{k} t^{k}}{k!}\right)\left(\sum_{k=0}^{\infty} \frac{A^{k} s^{k}}{k!}\right) \\
& \quad=\sum_{k=0}^{\infty} A^{k}\left(\sum_{i=0}^{\infty} \frac{t^{i} s^{k-i}}{!(k-1)!}\right)=\sum_{k=0}^{\infty} A^{k} \frac{(i+t)^{k}}{k!}=e^{A(t+S)}
\end{aligned}
$$

In particular, if $s=-t$, then

$$
e^{A t} e^{-A t}=e^{A(t-t)}=1
$$

Thus, the inverse of $e^{A t}$ ise $e^{-A t}$ since the inverse of $e^{A t}$ always exists $e^{A t}$ Is nonsingular, it is very important to remember that

$$
e^{t(A+B)}=e^{A t} e^{B t} \quad \text { if } A B=B A, e^{t(A+B)} \neq e^{A t} e^{B t} \quad \text { if } A B \neq B A
$$

To prove this note that

$$
\begin{aligned}
& e^{t(A+B)}=1+(A+B) t+\frac{(A+B)^{2}}{2!} t^{2}+\frac{(A+B)^{3}}{3!} t^{3}+\cdots \\
& e^{A t} e^{B t}=\left(1+A t+\frac{A^{2} t^{2}}{2!}+\frac{A^{3} t^{3}}{3!}+\cdots\right)\left(1+B t+\frac{B^{2} t^{2}}{2!}+\frac{A^{3} t^{3}}{3!}+\cdots\right) \\
& \quad=1+(A+B) t+\frac{A^{2} t^{2}}{2!}+\frac{A^{3} t^{3}}{3!}+\frac{A^{2} B t^{3}}{2!}+\frac{A B^{2} t^{3}}{2!}+\frac{B^{3} t^{3}}{3!} \\
& \quad+\cdots
\end{aligned}
$$

Hence

$$
\begin{aligned}
& e^{A t} e^{B t} \\
& -e^{A t} e^{B t} \\
& =\frac{B A-A B}{2!} t^{2}+\frac{B A^{2}+A B A+B^{2} A-B A B-2 A^{2} B-A B^{2}}{3!} t^{3}
\end{aligned}
$$

The difference between $e^{A t} e^{B t}$ and $e^{A t} e^{B t}$ vanishes if A and B commute
Laplace transforms Approach to the solution of Homogeneous state equation
Let us first consider the scalar case

$$
\begin{equation*}
x=a x \tag{2.13}
\end{equation*}
$$

Taking the Laplace transform of equation (2.13) we obtain
$S X(s)-X(0)=a X(s)$
Where $X(s)=\varphi[x]$ for solving (2.14) for $\mathrm{X}(\mathrm{s})$ gives

$$
X(s)=\frac{x(0)}{s-a}=(s-a)^{-1} x(0)
$$

The inverse lap lace transform of this last equation gives the solution

$$
x(t)=e^{a t}
$$

The foregoing to the approach to the solution of the homogeneous scalar differential equation can be extended to the homogeneous state equation

$$
\begin{equation*}
X^{\cdot}=A X \tag{2.15}
\end{equation*}
$$

Taking the Laplace transform of both sides of equation (2.15)

$$
S X(s)-X(0)=A X(s)
$$

Where $X(s)=\varphi[x]$.Hence $\quad(S I-A) X(s) X=X(0)$
Premultiplying both sides of this last equation by $(S I-A)^{-1}$ we obtain

$$
X(s)=(S I-A)^{-1} x(0)
$$

The inverse Laplace transformation of $\mathrm{X}(\mathrm{s})$ gives the solution $\mathrm{x}(\mathrm{t})$, Thus

$$
\begin{equation*}
x(t)=\varphi^{-1}\left[(S I-A)^{-1}\right] x(0) \tag{2.16}
\end{equation*}
$$

Note that

$$
\begin{aligned}
(S I-A)^{-1} & =\frac{1}{S}+\frac{A}{S^{2}}+\frac{A^{2}}{S^{3}} \\
& +\cdots
\end{aligned}
$$

Hence the inverse Laplace transform of $(S I-A)^{-1}$ gives

$$
\begin{gather*}
\varphi^{-1}\left[(S I-A)^{-1}\right]=1+A t+\frac{1}{2!} A^{2} t^{2}+\cdots+\frac{1}{k!} A^{k} t^{k}+\cdots \\
=e^{A t} \tag{2.17}
\end{gather*}
$$

The Laplace transform of a matrix is the matrix consisting of the inverse Laplace transform all elements form equation (2.16) and (2.17) the solution of equation (2.15) is obtained as
$X(t)=e^{A t} X(0)$
The importance of equation (2.17) lies in the fact that it provides a convenient means for finding the closed solution for matrix exponential

### 2.2 State Transition Matrix

We can write the solution of the homogeneous state equation

$$
\begin{equation*}
X^{\cdot}=A X \tag{2.18}
\end{equation*}
$$

$X(t)=\emptyset(t) X(0)$
Where $\varnothing(t)$ is an $\mathrm{n} \times \mathrm{n}$ matrix and is the unique solution of

$$
\varnothing(t)=A \emptyset(t), \quad \varnothing(0)=1
$$

To verify this note that

$$
X(0)=\emptyset(0) X(0)=X(0)
$$

And

$$
X^{\prime}(t)=\varnothing(t) X(0)=A \emptyset(t) X(0)=A X(t)
$$

We thus confirm that equation (2.19) is the solution of $t$ equation (2.16) from equation (2.17).and (2.18) we obtain

$$
\emptyset(t)=e^{A t}=\varphi^{-1}\left[(S I-A)^{-1}\right]
$$

Note that

$$
\emptyset^{-1}(t)=e^{-A t}=\emptyset(-t)
$$

From equation (2.19) we see that the solution of condition, hence the unique matrix $\emptyset(t)$ is called the state information matrix.

The state information contains all the information about the free motions of the system defined by (2.3)

If the Eigen values $\lambda_{1}, \lambda_{2}, \cdots, \lambda_{n}$ of the matrix A are distinct, than $\emptyset(t)$ will contain the $n$-exponentials

$$
e^{\lambda_{1} t}, \quad e^{\lambda_{2} t}, \cdots, e^{\lambda_{n} t}
$$

In particular, if the matrix $A$ is diagonal, then
$\varphi(t)=e^{A t}=\left[\begin{array}{lll}e^{\lambda_{1} t} & & 0 \\ & e^{\lambda_{2} t} & \\ 0 & & e^{\lambda_{n} t}\end{array}\right] \quad$ (A:diagonal)
If there is a multiplicity in the Eigen values for example $A$ are $\lambda_{1}, \lambda_{2}, \lambda_{3}, \lambda_{4}$, $\lambda_{5}, \ldots, \lambda_{n}$

Then $\phi(t)$ will contain, in addition to the exponential $e^{\lambda_{1} t}, e^{\lambda_{2}}, \ldots, e^{\lambda_{n} t}$ terms like $t e^{\lambda_{1} t}$ and $t^{2} e^{\lambda_{1} t}$

### 2.2.1 Properties of State Transition Matrix

We shall now some matrix the important properties of state transition matrix $\varnothing(t)$ for the time invariant system

$$
X^{\prime}=A X
$$

For which

$$
\emptyset(t)=e^{A t}
$$

We have the following

1. $\varnothing(t)=e^{A 0}=1$
2. $\varnothing(t)=e^{A t}=e^{(-A t)^{-1}}=[\emptyset(-t)]^{-1}$ or $\emptyset^{-1}(t)=\emptyset(-t)$
$3 . \emptyset\left(t_{1}+t_{2}\right)=e^{A\left(t_{1}+t_{2}\right)}=e^{A t_{1}} e^{A t_{2}}=\emptyset\left(t_{1}\right) \emptyset\left(t_{2}\right)=\emptyset\left(t_{2}\right) \emptyset\left(t_{1}\right)$
3. $[\varnothing(t)]^{n}=\emptyset(n t)$
4. $\varnothing\left(t_{2}-t_{1}\right) \emptyset\left(t_{1}-t_{0}\right) \emptyset=\varnothing\left(t_{2}-t_{0}\right)=\varnothing\left(t_{1}-t_{0}\right) \emptyset\left(t_{2}-t_{1}\right)$

### 5.2 Example

Obtain the state transition matrix $\phi(\mathrm{t})$ of the following system
$\left[\begin{array}{c}x^{\cdot} \\ x_{2}^{\cdot 1}\end{array}\right]=\left[\begin{array}{cc}0 & 1 \\ -2 & -2\end{array}\right]\left[\begin{array}{l}x_{1} \\ x_{2}\end{array}\right]$
Obtain also the inverse of the state transition matrix $\phi^{-1}(\mathrm{t})$ for system

$$
A=\left[\begin{array}{cc}
0 & 1 \\
-2 & -3
\end{array}\right]
$$

The state transition matrix $\emptyset(t)$ is given by

$$
\emptyset(t)=e^{A t}=\emptyset^{-1}\left[(S I-A)^{-1}\right]
$$

Since

$$
\text { SI- A }=\left[\begin{array}{ll}
\mathrm{S} & 0 \\
0 & S
\end{array}\right]-\left[\begin{array}{cc}
0 & 2 \\
-2 & -3
\end{array}\right]=\left[\begin{array}{cc}
\mathrm{S} & -1 \\
2 & \mathrm{~S}+3
\end{array}\right]
$$

The inverse of $(\mathrm{SI}-\mathrm{A})$ is given by

$$
\begin{aligned}
(\mathrm{SI}-\mathrm{A})^{-1}= & \frac{1}{(\mathrm{~S}+1)(\mathrm{S}+2)}\left[\begin{array}{cc}
\mathrm{S}+3 & 1 \\
-2 & \mathrm{~S}
\end{array}\right] \\
& =\left[\begin{array}{ll}
\frac{\mathrm{S}+3}{(\mathrm{~s}+1)(\mathrm{s}+2)} & \frac{-1}{(\mathrm{~s}+1)(\mathrm{s}+2)} \\
\frac{-2}{(\mathrm{~s}+1)(\mathrm{s}+2} & \frac{\mathrm{s}}{(\mathrm{~s}+1)(\mathrm{s}+2}
\end{array}\right]
\end{aligned}
$$

Hence

$$
\left.\Phi(\mathrm{t})=\mathrm{e}^{\mathrm{At}}=\phi^{-1}[\mathrm{SI}-\mathrm{A})^{-1}\right]=\left[\begin{array}{ll}
2 \mathrm{e}^{-\mathrm{t}}-\mathrm{e}^{-2 \mathrm{t}} & \mathrm{e}^{-\mathrm{t}}-\mathrm{e}^{-2 \mathrm{t}} \\
-2 \mathrm{e}^{\mathrm{t}}+2 \mathrm{e}^{2 \mathrm{t}} & -\mathrm{e}^{\mathrm{t}}+2 \mathrm{e}^{2 \mathrm{t}}
\end{array}\right]
$$

Noting that $\emptyset^{-1}(t)=\varnothing(-t)$ we obtain the inverse of the state transition matrix as follows
$\Phi^{-1}(t)=e^{-A t}=\left[\begin{array}{cc}2 e^{t}-e^{2 t} & e^{t}-e^{2 t} \\ -2 e^{t}+2 e^{2 t} & -e^{t}+2 e^{2 t}\end{array}\right]$

### 2.3 Solution of non-homogeneous state equations

We shall begin by considering the scalar case
$X^{\cdot}=a x+b u$
let us rewrite equation (2.20) as $X^{*}-a x=b u$ multiplying both sides of this equation by $e^{-a t}$ we obtain

$$
e^{-a t}\left[X^{\cdot}(t)-a x(t)\right]=\frac{d}{d t}\left[e^{-a t} x(t)\right]=e^{-a t} b u(t)
$$

Integrating this equation between o and t give

$$
e^{-a t} x(t)-x(0)=\int_{0}^{t} e^{-a t} b u(\tau) d \tau
$$

or

$$
x(t)=e^{a t} x(0)+e^{a t} \int_{0}^{t} e^{-a t} b u(\tau) d \tau
$$

The first term on the right hand side is the response to the initial condition and the second term is the response to the nonhomogeneous state equation described by
$X=A X+B U$
Where $\mathrm{X}=\mathrm{n}$ - vector

$$
\begin{aligned}
& U=r-\text { vector } \\
& U=n \times n \text { constant matrix } \\
& U=n \times r \text { constant matrix }
\end{aligned}
$$

By writing equating (2.21) as $X^{\cdot}-A X(t)=B U(t)$ and premultiplying both sides of this equation by $e^{-A t}$ we obtain

$$
e^{-A t}[X(t)-A X(t)]=\frac{d}{d t}\left[e^{-A t} X(t)\right]=e^{-A t} B U(t)
$$

Integration the preceding equation between o and t gives

$$
e^{-A t} x(t)-x(0)=\int_{0}^{t} e^{-A t} B u(\tau) d \tau
$$

Or
$x(t)=e^{A t} x(0)+e^{A t} \int_{0}^{t} e^{A(t-\tau)} B u(\tau) d \tau$
Equation above can also between as
$X(t)-\emptyset(t) X(0)+\int_{0}^{t} \emptyset(t-\tau) B u(\tau) d \tau$

Where $\quad \emptyset(t)=e^{A t}$
Equation (2.22) and (2.23) is the solution of equation (2.21) the solution $\mathrm{x}(\mathrm{t})$ is clearly the sum of a term consisting of the transition of the initial state and a term arising from the input vector. Laplace transforms Approach to the solution of nonhomogeneous state equation. The solution of the nonhomogeneous state equation
$X^{\cdot}=A X+B U$
Can also be obtained by the Laplace transform approach, the Laplace transform of this last equation yields
$S X(s)-X(0)=A X(t)+B U(t)$
Or
$(S I-A) X(s)=(0)+B U(s)$

Premultiplying both sides of this last equation by $(2.25)(S I-A)^{-1}$
We obtain
$X(s)=(S I-A)^{-1} X(0)+(S I-A)^{-1} B U(s)$
Using the relationship given by equation (2.17) gives
$X(s)=\emptyset\left[e^{A t}\right] X(0)+\emptyset\left[e^{A t}\right] B U(s)$
The inverse Laplace transform of this last equation can be obtained by use of the convolution integral as follows.
$x(t)=e^{A t} x(0)+e^{A t} \int_{0}^{t} e^{A(t-\tau)} B u(\tau) d \tau$
Solution in term s of $\mathrm{x}\left(\mathrm{t}_{0}\right)$, thus for we have assumed the initial time to be zero. If however, the initial time is given by to instead of o , them the saluted to above equation must be modified to

$$
x(t)=e^{A(t-\tau)} x(0)+e^{A t} \int_{0}^{t} e^{A(t-\tau)} B u(\tau) d \tau
$$

### 6.2 Example

Obtain the time response of the following system

$$
\left[\begin{array}{l}
\mathrm{x}_{1} \\
\mathrm{x}_{2}
\end{array}\right]=\left[\begin{array}{cc}
0 & 1 \\
-2 & -3
\end{array}\right],\left[\begin{array}{l}
\mathrm{x}_{1} \\
\mathrm{x}_{2}
\end{array}\right]\left[\begin{array}{l}
0 \\
1
\end{array}\right]
$$

Where $\mathrm{u}(\mathrm{t})$ is the unit step function occurring at $\mathrm{t}=0$ or $\mathrm{U}(\mathrm{t})=\mathrm{I}(\mathrm{t})$
For this system

$$
\mathrm{A}=\left[\begin{array}{cc}
0 & 1 \\
-2 & -3
\end{array}\right], \mathrm{B}=\left[\begin{array}{l}
0 \\
1
\end{array}\right]
$$

The state transition matrix $\varnothing(t)=e^{A t}$ was obtained

The response to the unit step input is then is obtained as
$\mathrm{x}(\mathrm{t})=$
$e^{A t} x(0)+\int_{0}^{t}\left[\begin{array}{cc}2 e^{-(t-\tau)}-e^{-2(t-\tau)} & e^{-(t-\tau)}-e^{-2(t-\tau)} \\ -2 e^{-(t-\tau)}+2 e^{-2(t-\tau)} & -e^{-(t-\tau)}+2 e^{-2(t-\tau)}\end{array}\right]\left[\begin{array}{l}0 \\ 1\end{array}\right][1] d \tau$
Or

$$
\left[\begin{array}{l}
X_{1}(t) \\
X_{1}(t)
\end{array}\right]=\left[\begin{array}{cc}
2 e^{-t}-e^{-2 t} & e^{-t}-e^{-2 t} \\
-2 e^{-t}+2 e^{-2 t} & -e^{-t}+2 e^{-2 t}
\end{array}\right]\left[\begin{array}{l}
X_{1}(0) \\
X_{1}(0)
\end{array}\right]+\left[\begin{array}{c}
\frac{1}{1}-e^{-t}+\frac{1}{1} e^{-2 t} \\
e^{-t}-e^{-2 t}
\end{array}\right]
$$

If the initial state is zero $\mathrm{x}(0)=0$. Then $\mathrm{u}(\mathrm{t})$ can be simplified to

$$
\left[\begin{array}{l}
X_{1}(t) \\
X_{1}(t)
\end{array}\right]=\left[\begin{array}{c}
\frac{1}{1}-e^{-t}+\frac{1}{1} e^{-2 t} \\
e^{-t}-e^{-2 t}
\end{array}\right]
$$

### 2.4 State model of linear system

Consider multiple input multiple output nth order

Number of inputs $=m$
Number of outputs $=p$


State variables

$$
U(t)=\left[\begin{array}{c}
u_{1}(t) \\
u_{2}(t) \\
\vdots \\
\vdots \\
u_{m}(t)
\end{array}\right], \quad X(t)=\left[\begin{array}{c}
x_{1}(t) \\
x_{2}(t) \\
\vdots \\
\vdots \\
x_{n}(t)
\end{array}\right], Y(t)=\left[\begin{array}{c}
Y_{1}(t) \\
Y_{2}(t) \\
\vdots \\
\vdots \\
Y_{P}(t)
\end{array}\right]
$$

All Colum vectors having order $m \times 1, n \times 1$ and $p \times 1$ respectively
For such a system the state variable representation can be arranged in the form of $n$ fist order differential equations

$$
\begin{aligned}
& \frac{d x_{1}(t)}{d t}=f_{1\left(x_{1}, x_{2}, \ldots \ldots, x_{n}, U_{1}, \ldots \ldots, U_{m}\right)} \\
& \frac{d x_{2}(t)}{d t}=f_{2\left(x_{1}, x_{2}, \ldots \ldots, x_{n}, U_{1, \ldots}, \ldots, U_{m}\right)}
\end{aligned}
$$

$\frac{d x_{n}(t)}{d t}=f_{n\left(x_{1}, x_{2}, \ldots \ldots, x_{n}, U_{1}, \ldots \ldots, U_{m},\right)}$
Where $f=\left[\begin{array}{c}f_{1} \\ f_{2} \\ \vdots \\ f_{n}\end{array}\right]$ is the functional operator
Integrating the above equation

$$
\begin{equation*}
X_{i}(t)=X_{i}\left(t_{0}\right)+\int_{t_{0}}^{t} f_{i\left(x_{1}, x_{2}, \ldots, x_{n}, U_{1}, \ldots, U_{m}\right)} d t \tag{2.26}
\end{equation*}
$$

$\qquad$ Where $\mathrm{i}=1,2$,

Thus n state variables and hence state vector at any time t can be determined uniquely any n dimensional time invariant system has state equation in functional form as
$\frac{d x}{d t}=f(X, U)$
While outputs of such system are dependent on the state and instantaneous Inputs functional output equation can be written as
$X(t)=g(x, u)$
Where g is functional output
For the time invariant system the same equation s can be written as
$\frac{d x}{d t}=f(X, U, t) \cdots \cdots$ State equation
$Y(t)=g(X, U, t) \cdots \cdots$ Output equation
Diagrammatically this can be represented in the figure above


Input-output state description of system
The functional equations can be expressed in terms of linear combination of system states and inputs as

$$
\begin{aligned}
& \dot{x}_{1}=a_{11} x_{1}+a_{12} x_{2}+\ldots+a_{1 n} x_{n}+b_{11} U_{1}+b_{12} U_{2}+\ldots+b_{1 m} U_{m} \\
& \dot{x}_{2}=a_{21} x_{1}+a_{22} x_{2} \ldots+a_{2 n} x_{n}+b_{21} U_{1}+b_{22} U_{2}+\ldots+b_{2 m} U_{m} \\
& \vdots \\
& \vdots \\
& \dot{X}_{n}=a_{n 1} x_{1}+a_{n 2} x_{2}+\ldots+a_{n n} x_{n}+b_{n 1} U_{1}+b_{n 2} U_{2}+\ldots+b_{n m} U_{m}
\end{aligned}
$$

For linear time invariant systems the coefficients $a_{i j}$ and $b_{i j}$ thus all equations can be written in the vector matrix form as
$X^{\prime}(t)=A X(t)+B U(t)$
Where
$\mathrm{X}(\mathrm{t})=$ state vector matrix of order $\mathrm{n} \times 1$
$\mathrm{U}(\mathrm{t})=$ input vector matrix of order $\mathrm{m} \times 1$
$\mathrm{A}=$ system matrix or Evaluation matrix of order $\mathrm{n} \times n$
$\mathrm{B}=$ input matrix of control matrix of order $\mathrm{n} \times m$

Similarly the output variables at time $t$ can be expressed as the linear combinations of the input variables and state variables at time $t$ as

$$
\begin{aligned}
& Y_{1}(t)=c_{11} X_{1}(t)+\ldots+c_{1 n} X_{n}(t)+d_{11} U_{1}(t)+\ldots+d_{1 m} U_{m}(t) \\
& \vdots \\
& Y_{p}(t)=c_{p 1} X_{1}(t)+\ldots+c_{p 1} X_{n}(t)+d_{p 1} U_{1}(t)+\ldots+d_{p m} U_{m}(t)
\end{aligned}
$$

For linear time invariant systems, the coefficients $C_{i j}$ and $d_{i j}$ are constants thus all the output equations can be written in the vector matrix form as

$$
\mathrm{Y}(\mathrm{t})=\mathrm{CX}(\mathrm{t})+\mathrm{DU}(\mathrm{t})
$$

Where
$\mathrm{Y}(\mathrm{t})=$ output vector matrix of order $\mathrm{p} \times 1$
$\mathrm{C}=$ output matrix or observation matrix of order $\mathrm{p} \times n$
$\mathrm{D}=$ Direct transmission matrix of order $\mathrm{p} \times m$
The two vector equations together is called state model of linear system
$X^{\cdot}(t)=A X(t)+B U(t) \quad$ State equation
$Y(t)=C X(t)+D U(t) \quad$ Output equation
This state model of a system
For linear time invariant system the matrix $\mathrm{A}, \mathrm{B}, \mathrm{C}$ and D are also time dependent. Thus
$X^{\cdot}(t)=A(t) X(t)+B(t) U(t)$
$Y(t)=C(t) X(t)+D(t) U(t)$
For linear time invariant system

### 2.5 State model of single input-single output system

Consider a single input -single output
$\rightarrow$ But its order n hence n state variable are required to define state of system in such case the state model is
$X^{\prime}(t)=A X(t)+B U(t)$
$Y(t)=C X(t)+d U(t)$
Where
$A=n \times n$ Matrix $B=n \times 1$ matrix
$C=1 \times n$ Matrix $\mathrm{d}=$ constant and $\mathrm{u}(\mathrm{t})=$ single scalar input variables
In general remember the order of various matrices
$A=$ Evoluation matrix $\rightarrow n \times n$
$B=$ Control matrix $\rightarrow n \times m$
$c=$ Observation matrix $\rightarrow p \times n$
$D=$ Transimission matrix $\rightarrow p \times m$

## Chapter Three

## Controllable Pairs of matrice

## 3. Introduction

A system is said to be controllable at time $t_{0}$, if it is possible by means of an unconstrained control vector to transfer the system from. Any initial state $\mathbf{x}\left(t_{0}\right)$ to any other state in a finite interval of time. A system is said to be observable at time $t_{0}$ if, with the system in state $\mathbf{x}\left(t_{0}\right)$, it is possible to determine this state from the observation of the output over a finite time interval. The concepts of controllability and observability were introduced by Kalman,s they Play an important role in the design of control systems in state space. In fact, the Conditions of controllability and observability may govern the existence of a complete Solution to the control system design problem. The solution to this problem may not, exist if the system considered is not controllable. Although most physical systems are .Controllable and observable, corresponding mathematical models may not possess the Property of controllability and observability Then it is necessary to know the. Conditions under which a system controllable and observable this deals with controllability and the next discusses observability in what follows, we shall first derive the condition for complete state controllability. Then we derive alternative forms of the condition for complete state controllability followed by discussions of complete output controllability. Finally, we present the concept of stabilizability

### 3.1 Complete State Controllability of Continuous-Time Systems.

Consider the, Continuous-time system.
1). (3

$$
x=\mathrm{Ax}+B U
$$

Where $\mathbf{x}=$ state vector (n-vector)
$\mathrm{U}=$ control signal (scalar)
$\mathrm{n}=\mathrm{n} \times \mathrm{n}$ matrix
$\mathrm{B}=\mathrm{n} \times 1$ matrix
The system described by Equation (3.1) is said to be state controllable at $t=t_{0}$ if it is possible to construct an unconstrained control signal that will transfer an initial state to any final state in a finite time interval If every state is controllable, then the system is said to be completely state controllable We shall now derive the condition for complete state controllability. Without loss of generality, we can assume that the final state is the origin of the state space and that the Initial time is zero .or $t_{0}=0$.The solution of Equation (3.1) is
$\mathrm{X}(\mathrm{t})=e^{A t} \mathrm{X}(0)+\int_{0}^{t} e^{A(t-\tau)} B U(\tau) d \tau$
Applying the definition of complete state controllability just given, we have $\mathrm{X}\left(t_{1}\right)=e^{A t_{1}} \mathrm{X}(0)+\int_{0}^{t_{1}} e^{A\left(t_{1}-\tau\right)} B U(\tau) d \tau$
Or
$\mathrm{X}(0)=-\int_{0}^{t_{1}} e^{-A \tau} B U(\tau) d \tau$
2). (3

By definition of $e^{-A t}$ can be written

$$
e^{-A t}=\sum_{k=0}^{n-1} \alpha_{k}(\tau) A^{k}
$$

3). (3

Substituting Equation (3.3) into Equation (3.2) gives
$\mathrm{X}(0)=-\sum_{k=0}^{n-1} A^{k} \mathrm{~B} \int_{0}^{t_{1}} \alpha_{k}(\tau) \mathrm{U}(\tau) \mathrm{d} \tau$
4). 3

Let us put $\int_{0}^{t_{1}} \alpha_{k}(\tau) \mathrm{u}(\tau) \mathrm{d} \tau=B_{k}$. Then Equation (3.4) becomes
$\mathrm{X}(0)=-\sum_{k=0}^{n-1} A^{k} \mathrm{~B} B_{k}=-\left[B: A B: \cdots A^{n-1} B\right]$
5).(3

If the system is completely state controllable, then, given any initial state $\mathbf{x}$ (0) Equation (3.5) must be satisfied. This requires that the rank of the $\mathrm{n} \times \mathrm{n}$ matrix

$$
\left[B: A B: \cdots A^{n-1} B\right]
$$

From this analysis, we can state the condition for complete state controllability as follows. The system given by Equation (3.1) is completely state controllable if and only
If the vectors $\mathrm{B}, \mathrm{AB}, \cdots, A^{n-1} \mathrm{~B}$ are linearly independent, or the $\mathrm{n} \times \mathrm{n}$ matrix

$$
\left[B: A B: \cdots A^{n-1} B\right]
$$

Is of rank n
The result just obtained can be extended to the case where the control vector $\mathbf{u}$ is n -dimensional. If the system is described by
$X^{-}=\mathrm{AX}+B U$
Where $\mathbf{U}$ is an r -vector, then it can be proved that the condition for complete state. Controllability is that the $\mathrm{n} \times \mathrm{nr}$ matrix

$$
\left[B: A B: \cdots A^{n-1} B\right]
$$

Be of rank n , or contain n linearly independent column vectors. The matrix

$$
\left[B: A B: \cdots A^{n-1} B\right]
$$

Is commonly called the controllability matrix
1 Consider the system given by.Example 3
$\left[\begin{array}{l}x_{1}{ }_{1} \\ x_{2}\end{array}\right]=\left[\begin{array}{cc}1 & 1 \\ 0 & -1\end{array}\right]\left[\begin{array}{l}x_{1} \\ x_{2}\end{array}\right]+\left[\begin{array}{l}1 \\ 0\end{array}\right] \mathrm{U}$
Since $[B: A B]=\left[\begin{array}{ll}1 & 1 \\ 0 & 0\end{array}\right]=$ singular matrix
The system is not completely state controllable.

## 3:1:1 Alternative Form of the Condition for Complete State

Controllability. Consider the system defined by
6).

$$
X^{\cdot}=\mathrm{A} X+B U
$$

Where $\mathbf{x}=$ state vector (n-vector)
$\mathrm{U}=$ control signal (scalar)
$\mathrm{n}=\mathrm{n} \times \mathrm{n}$ matrix
$\mathrm{B}=\mathrm{n} \times 1$ matrix
If the eigenvectors of A are distinct, then it is possible to find a transformation matrix P such that
$P^{-1} \mathrm{AP}=\mathrm{D}=\left[\begin{array}{cccc}\lambda_{1} & \cdots & \cdots & 0 \\ \vdots & \lambda_{1} & \cdots & \vdots \\ \vdots & \vdots & \ddots & \vdots \\ 0 & \cdots & \cdots & \lambda_{n}\end{array}\right]$
Note that if the eigenvalues of Aare distinct, then the eigenvectors of Aare distinct; however, the converse is not true. For example, an $\mathrm{n} \times \mathrm{n}$ real symmetric matrix having, multiple eigenvalues has $n$ distinct eigenvectors. Note also that each column of the Matrix is an eigenvector of $\mathbf{A}$ associated with $\lambda_{i}=(\mathrm{i}=1,2, \cdots, n)$

Let us define
7). $\quad \mathrm{X}=\mathrm{PZ}$

Substituting Equation (3.5) into Equation (3.4), we obtain

By defining

$$
\begin{equation*}
Z{ }^{=P^{-1}} A P Z+P^{-1} B U \tag{3.8}
\end{equation*}
$$

By defining

$$
\begin{aligned}
& P^{-1} \mathrm{~B}=\mathrm{F}=\left(f_{i j}\right) \text {.We can rewrite Equation (3.7) as } \\
& Z_{1=}=\lambda_{1 Z_{1+}} f_{11 u_{1+}} f_{12 u_{2+\ldots+}} f_{1 r u_{r}} \\
& Z_{2=} \lambda_{2 Z_{2+}} f_{21 u_{1+}} f_{22 u_{2+\ldots}+} f_{2 r u_{r}} \\
& \vdots \\
& Z_{n=} \lambda_{n Z_{n+}} f_{n 1 u_{1+}} f_{n 2 u_{2+\ldots}+}+f_{n r u_{r}}
\end{aligned}
$$

If the elements of any one row of the $\mathrm{n} \times r$ matrix F are all zero, then the corresponding state variable cannot be controlled by any of the $u_{i}$. Hence, the condition of complete state controllability is that if the eigenvectors of A are distinct, then the system is completely State controllable if and only if no row of $P^{-1} \mathrm{~B}$ has all zero elements. It is important. To note that, to apply this condition for complete state controllability, we must put the matrix $P^{-1} A P$ in Equation (3.8) in diagonal form If the A matrix in Equation (3.6) does not possess distinct eigenvectors, then diagonalization is impossible. In such a case, we may transform A into a Jordan canonical Form. If for example, A has eigenvalues and has n-3 distinct. Eigenvectors, then the Jordan canonical form of A is

$$
J=\left[\begin{array}{ccccc}
\lambda_{1} & 1 & 0 & 0 & 0 \\
0 & \lambda_{1} & 0 & 0 & 0 \\
0 & 0 & \lambda_{4} & 1 & \vdots \\
\vdots & 0 & \vdots & \vdots & \vdots \\
0 & 0 & 0 & 0 & \lambda_{n}
\end{array}\right]
$$

The square sub matrices on the main diagonal are called Jordan blocks.
Suppose that we can find a transformation matrix $\mathbf{S}$ such that

$$
S^{-1} A S=J
$$

If we define a new state vector $\mathbf{z}$ by

$$
\begin{equation*}
X=S Z \tag{3.9}
\end{equation*}
$$

Then substitution of Equation (9) into Equation (6) yields

$$
\begin{align*}
& Z \cdot S^{-1} A S Z+S^{-1} \quad B U \\
&=J Z+S^{-1} B U \tag{3.10}
\end{align*}
$$

The condition for complete state controllability of the system of Equation
6) may. Then be stated as follows: The system is completely state . 3
10) are .controllable if and only if, No two Jordan blocks in J of Equation (3 associated with the same eigenvalues, the elements of any row of that
3) .correspond to the last row of each Jordan block. Are not all zero, and (3 the elements of each row of that correspond to distinct, Eigenvalues are not all zero.

2 the following systems are completely state controllable:.Example3
$\left[\begin{array}{l}x_{1}{ }_{1} \\ x_{2}\end{array}\right]=\left[\begin{array}{cc}-1 & 0 \\ 0 & -2\end{array}\right]\left[\begin{array}{l}x_{1} \\ x_{2}\end{array}\right]+\left[\begin{array}{l}2 \\ 5\end{array}\right] \mathrm{U}$
Since $[B: A B]=\left[\begin{array}{cc}2 & -2 \\ 5 & -10\end{array}\right]=$ nonsingular matrix the system is completely state controllable

### 3.1.2 Conditions for Complete State Controllability in the Plane

The condition for complete state controllability can be stated in terms of transfer functions or transfer matrices. It can be proved that a necessary and sufficient condition for complete state controllability is that no cancellation occur in the transfer function or transfer matrix. If cancellation occurs; the system cannot be controlled in the direction of the canceled mode.

3 Consider the following transfer function .Example 3

$$
\frac{X(S)}{U(S)}=\frac{(s+2,5)}{(s+2,5)(s-1)}
$$

Clearly, cancellation of the factor ( $s+2.5$ ) occurs in the numerator and denominator of this. Transfer function. (Thus one degree of freedom is lost.) Because of this cancellation, this system is not completely state controllable. The same conclusion can be obtained by writing this transfer function in the form of a state, Equation a state-space representation is
$\left[\begin{array}{l}x_{1} \\ x_{2} \\ x_{2}\end{array}\right]=\left[\begin{array}{cc}0 & 1 \\ 2,5 & 1,5\end{array}\right]\left[\begin{array}{l}x_{1} \\ x_{2}\end{array}\right]+\left[\begin{array}{l}1 \\ 1\end{array}\right] \mathrm{U}$
Since $\quad[B: A B]=\left[\begin{array}{ll}1 & 1 \\ 1 & 1\end{array}\right]$
The rank of the matrix is $[B: A B]$.Therefore, we arrive at the same conclusion: The system is not completely state controllable.

4 Show that the system described by. 3 Example

$$
\begin{array}{r}
X^{\cdot}=\mathrm{AX}+B U \\
Y=C X
\end{array}
$$

Where $\mathbf{x}=$ state vector ( n -vector)

$$
\begin{aligned}
\mathrm{U} & =\text { control vector(r-vector) } \\
Y & =\text { output control }(\mathrm{m} \text {-vector) }(\mathrm{m} \leq n) \\
\mathrm{A} & =\mathrm{n} \times \mathrm{n} \text { matrix } \\
\mathrm{B} & =\mathrm{n} \times r \text { matrix } \\
C & =m \times n \text { matrix }
\end{aligned}
$$

Is completely output controllable if and only if the composite $\mathrm{m} \times \mathrm{nr}$ matrix $P$, where

$$
P=\left[C B: C A B: C A^{2} B: \cdots \cdots: C A^{n-1} B\right]
$$

Is of rank m (Notice) that complete state controllability is neither necessary nor sufficient for complete output controllability

Suppose that the system is output controllable and the output $\mathbf{y}(\mathrm{t})$ starting from any $\mathbf{y}(0)$ the initial output can be transferred to the origin of the output space in a finite time interval
$0 \leq \mathrm{t} \leq$ that is

$$
\mathbf{y}(\mathrm{T})=\mathrm{Cx}(\mathrm{~T})=\mathbf{0}
$$

Since the solution of Equation P is

$$
\begin{equation*}
X(t)=e^{A t\left[X(0)+\int_{0}^{t} e^{-A t} B u(\tau) d \tau\right]} \tag{3.11}
\end{equation*}
$$

At $\mathrm{t}=T$ we have

$$
X(t)=e^{A T\left[X(0)+\int_{0}^{T} e^{-A \tau} B u(\tau) d \tau\right]}
$$

Substituting Equation (3.12) into Equation (3.11), we obtain
$\mathrm{Y}(\mathrm{T})=\mathrm{Cx}(\mathrm{T})$

$$
\begin{equation*}
=C e^{A T\left[X(0)+\int_{0}^{T} e^{-A \tau} B u(\tau) d \tau\right]}=0 \tag{3.13}
\end{equation*}
$$

On the other hand, $\mathbf{y}(0)=\mathrm{Cx}(0)$. Notice that the complete output controllability means that the. Vector $\mathrm{Cx}(0)$ spans the m-dimensional output space. Since $e^{A T}$ is nonsingular, if $\mathrm{Cx}(0)$ spans the m-dimensional output space, so does $\mathrm{C} e^{A T} \mathbf{x}(0)$, and vice versa. From Equation (3.3) we obtain

$$
\mathbf{C} e^{A T} \mathbf{x}(0)=-\mathbf{C} e^{A T} \int_{0}^{T} e^{A \tau} B U(T-\tau) d \tau
$$

Note that $e^{A T} \int_{0}^{T} e^{A \tau} B u(T-\tau) d \tau c$ an be expressed as the sum of $A^{i} B_{j}$; that is

$$
\int_{0}^{T} e^{A \tau} B U(T-\tau) d \tau=\sum_{i=0}^{p-1} \sum_{j=0}^{r} \gamma_{i j} A^{i} \boldsymbol{B}_{j}
$$

Where

$$
\gamma_{i j}=\int_{0}^{T} \alpha_{i(\tau) U_{j}}(\tau)(T-\tau) d \tau=\text { sclar }
$$

And $\alpha_{i}(\tau)$ satisfies

$$
e^{A \tau=\sum_{i=1}^{p-1} \alpha_{i}(\tau)} A^{i}
$$

(P: degree of the minimal polynomial of (3.12))

And is the $B_{j}$; jth column of Therefore, we can write $\mathbf{C} e^{A T} \mathbf{x}(0$ as

$$
\mathbf{C} e^{A T} \mathbf{x}(0)=-\sum_{i=0}^{p-1} \sum_{j=0}^{r} \gamma_{i j} C A^{i} B_{j}
$$

From this last equation, we see that $\mathbf{C} e^{A T} \mathbf{x}(0)$ is a linear combination of $C A^{i} B_{j} .(\mathrm{i}=0,1,2, \mathrm{p}, \mathrm{P}-1 ; \mathrm{j}=1,2, \mathrm{p}, \mathrm{r})$. Note that if the rank of $\mathbf{Q}$, where

$$
\mathrm{M}=\left[C B \vdots C A B: C A^{2} B: \cdots \cdots: C A^{n-1} B\right] \quad(p \leq n)
$$

Is m , then so is the rank of P , and vice versa. [This is obvious if $\mathrm{p}=\mathrm{n}$. If $\mathrm{p}<\mathrm{n}$, then the $\mathrm{C} A^{h} B_{j}$. (Where $\mathrm{p} \leq \mathrm{h} \leq \mathrm{n}-1$ ) are linearly dependent on $\mathrm{C} B_{j}, \mathrm{CA} B_{j}$, $\mathrm{C} A^{p-1} B_{j}$. Hence, the rank of p is equal to that of Q .] If the rank of P is m , then $\mathrm{C} e^{A T} \mathrm{x}(0)$ spans the m-dimensional output, space. This means that if the rank of P ism, then $\mathrm{Cx}(0)$ also spans the m -dimensional output space, and the system is completely output controllable. Conversely, suppose that the system is completely output controllable, but the rank of P is k , where $\mathrm{k}<\mathrm{m}$. Then the set of all initial outputs that can be transferred to the origin is of,

K-dimensional space hence, the dimension of this set is less than m . This contradicts the assumption, that the system is completely output controllable. This completes the proof. Note that it can be immediately proved that, in the system of Equations (3.12) and (3.13), state controllability on
$0 \leq \mathrm{t} \leq \mathrm{T}$ implies complete output controllability on $0 \leq \mathrm{t} \leq \mathrm{T}$, if and only if m rows of C are linearly in depend

5 obtain a state-space equation and output equation and test . $\mathbf{3}$ Example controllable for the system defined by

$$
\frac{Y(S)}{U(S)}=\frac{2 s^{3}+s^{2}+s+2}{s^{3}+4 s^{2}+5 s+2}
$$

From the given transfer function, the differential equation for the system is

$$
y^{\prime \prime \prime}+4 y^{\prime \prime}+5 y^{\prime}+2 y=2 u^{\prime \prime}+u^{\prime \prime}+u^{\prime}+2 u
$$

Comparing this equation with the standard equation given by Equation

$$
y^{\cdots}+a_{1} y^{\prime \prime}+a_{2} y^{\prime}+2 y=b_{0} u^{\cdots}+b_{1} u^{\prime \prime}+b_{2} u^{\prime}+b_{3} u
$$

We find

$$
\begin{gathered}
a_{1}=4, a_{2}=5, a_{3}=2 \\
b_{0}=2, b_{1}=1, b_{2}=1, b_{3}=2 \\
\text { and }
\end{gathered}
$$

$\beta_{0}=b_{0}=2$
$\beta_{1}=b_{1}-a_{1} \beta_{0}=1-4 \times 2=-7$
$\beta_{2}=b_{2}-a_{1} \beta_{1}-a_{2} \beta_{0}=1-4 \times(-7)-5 \times 2=19$
$\beta_{3}=b_{3-a_{1} \beta_{2}-a_{2} \beta_{1-a_{3} \beta_{0}}=2-4 \times 19-5 \times(-7)-2 \times 2=-43}$
And
$x_{1}=\mathrm{y}-\beta_{0} \mathrm{u}=\mathrm{y}-2 \mathrm{u}$
$x_{2=} x_{1}-\beta_{1} \mathrm{u}=x_{1+7 u}$

$$
x_{3}=x_{2}-\beta_{2} \mathrm{u}=x_{2}-19 \mathrm{u}
$$

And

$$
\begin{aligned}
& x_{1}=x_{2-7 u} \\
& x_{2}=x_{2+19 u}
\end{aligned}
$$

$x_{3=}-a_{3} x_{1}-a_{2} x_{2}-a_{1} x_{3+} \beta_{3} u=-2 x_{1}-5 x_{2}-4 x_{2-} 43 u$
Hence, the state-space representation of the system is

$$
\begin{aligned}
& {\left[\begin{array}{l}
x_{1} \\
x \cdot{ }_{2} \\
x \cdot{ }_{3}
\end{array}\right]=\left[\begin{array}{ccc}
0 & 1 & 0 \\
0 & 0 & 1 \\
-2 & -5 & -2
\end{array}\right]\left[\begin{array}{l}
x_{1} \\
x_{2} \\
x_{3}
\end{array}\right]+\left[\begin{array}{c}
-7 \\
19 \\
-43
\end{array}\right]+2 \mathrm{U}} \\
& \mathrm{Y}=\left[\begin{array}{lll}
1 & 0 & 0
\end{array}\right]\left[\begin{array}{l}
x_{1} \\
x_{2} \\
x_{3}
\end{array}\right]
\end{aligned}
$$

Test controllable
Controllability matrix $=\left[B: A B: A^{2} B\right]$
Since

$$
\begin{aligned}
& B=\left[\begin{array}{c}
-7 \\
19 \\
-43
\end{array}\right], A B=\left[\begin{array}{ccc}
0 & 1 & 0 \\
0 & 0 & 1 \\
-2 & -5 & -2
\end{array}\right]\left[\begin{array}{c}
-7 \\
19 \\
-43
\end{array}\right]=\left[\begin{array}{c}
19 \\
-43 \\
5
\end{array}\right] \\
& A^{2}=\left[\begin{array}{ccc}
0 & 0 & 1 \\
-2 & -5 & -2 \\
4 & 8 & -1
\end{array}\right], A^{2} B=\left[\begin{array}{c}
-43 \\
5 \\
167
\end{array}\right]
\end{aligned}
$$

$\mathrm{M}=\left[B: A B: A^{2} B\right]==\left[\begin{array}{ccc}-7 & 19 & -43 \\ 19 & -43 & 5 \\ -43 & 5 & 167\end{array}\right]$
$|M|=61492 \neq 0$ is system is complete controllable
Example3.6 the following systems are completely state controllable

$$
\begin{gathered}
A=\left[\begin{array}{lll}
2 & 0 & 0 \\
0 & 2 & 0 \\
0 & 3 & 1
\end{array}\right] \quad B=\left[\begin{array}{ll}
0 & 1 \\
1 & 0 \\
0 & 1
\end{array}\right] \quad C=\left[\begin{array}{lll}
1 & 0 & 0 \\
0 & 1 & 0
\end{array}\right] \quad D=\left[\begin{array}{ll}
0 & 0 \\
0 & 0
\end{array}\right] \\
A=\left[\begin{array}{ll}
0 & 2 \\
2 & 0 \\
3 & 1
\end{array}\right] \quad A^{2}=\left[\begin{array}{lll}
4 & 0 & 0 \\
0 & 4 & 0 \\
0 & 9 & 1
\end{array}\right] \quad A^{2} B=\left[\begin{array}{ll}
0 & 4 \\
4 & 0 \\
9 & 1
\end{array}\right] \\
M=\left[\begin{array}{llllll}
0 & 1 & 0 & 2 & 0 & 4 \\
1 & 0 & 2 & 0 & 4 & 0 \\
0 & 1 & 3 & 1 & 9 & 1
\end{array}\right]
\end{gathered}
$$

$\operatorname{Rank}(M)=3, \operatorname{det}($ left half $)=-1(3)=-3 \neq 0$ fully state controllability

### 3.1.3 Vander Monde Matrix

If the state model is obtained using the phase variables then the matrix $A$ is in bush for or phase variable form as

$$
A=\left[\begin{array}{ccccc}
0 & 1 & 0 & \cdots & 0 \\
0 & 0 & 1 & \cdots & 0 \\
\vdots & & & & \\
0 & 0 & 0 & \cdots & 1 \\
-a_{n} & \cdots a_{n-1} & -a_{n-2} & \cdots & -a_{1}
\end{array}\right]
$$

And the characteristic equation i.e. denominator of $\mathrm{T}(\mathrm{s})$ is

$$
F(s)=s^{n}+a_{1} s^{n-1}+\cdots+a_{n-1} s+a_{n}=0
$$

In such a case model matrix takes a form of a special matrix as

$$
\mathrm{V}=\left[\begin{array}{cccc}
1 & 1 & \cdots & 1 \\
\lambda_{1} & \lambda_{2} & \cdots & \lambda_{\mathrm{n}} \\
\lambda_{1}^{2} & \lambda_{2}^{2} & \cdots & \lambda_{n}^{2} \\
\vdots & & \cdots & \cdots \\
\lambda_{1}^{n-i} & \lambda_{2}^{n-2} & \cdots & \lambda_{n}^{n-1}
\end{array}\right]
$$

Such a model matrix for matrix A which in phase variable is called Vander monde matrix

Example 3.7 consider a state model with matrix A as

$$
A=\left[\begin{array}{cc:c}
0 & 2 & 0 \\
4 & 0 & 1 \\
-48 & -34 & -9
\end{array}\right]
$$

Determine (a) characteristic equation (b) Eigen values (c) Eigen vector and model matrix ,also prove that the transformed $M^{-1} A M$ a diagonal matrix
(a) Characteristic equation is $|\lambda I-A|$

$$
\begin{gathered}
A=\lambda\left|\begin{array}{lll}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & 0 & 1
\end{array}\right|-\left|\begin{array}{ccc}
0 & 2 & 0 \\
4 & 0 & 1 \\
-48 & -34 & -9
\end{array}\right|=0 \quad,\left|\begin{array}{ccc}
\lambda & -2 & 0 \\
4 & 0 & 1 \\
48 & 34 & \lambda+9
\end{array}\right|=0 \\
\lambda^{2}(\lambda+9)+2 \times 48+0+0-8(\lambda+9)+34 \lambda=0 \\
\lambda^{3}+9 \lambda^{2}+26 \lambda+24=0
\end{gathered}
$$

This is required characteristic equation
(b) To find Eigen values test $\lambda=-2$ for root

$$
\begin{aligned}
& -2 \left\lvert\, \begin{array}{rrrr}
1 & 9 & 26 & 24 \\
-2 & -14 & -24 \\
\hline 1 & 7 & 12 & 0
\end{array}\right. \\
& \begin{array}{l}
(\lambda+2)\left(\lambda^{2}+7 \lambda+12\right)=0 \\
\text { i.e }(\lambda+2)(\lambda+3)(\lambda+4)=0 \\
\text { i.e } \lambda_{1}=-2, \lambda_{2}=-3 \text { and } \lambda_{3}=-4
\end{array}
\end{aligned}
$$

These are the Eigen values of the matrix A
(c) To find Eigen vectors, obtain matrix $\left[\lambda_{i} I-A\right]$ for each Eigen values by substituting value of $\lambda$ in above equation

$$
\begin{gathered}
\text { For } \lambda_{1}=-2,\left[\lambda_{i} I-A\right]=\left|\begin{array}{ccc}
-2 & -2 & 0 \\
-4 & -2 & -1 \\
48 & 34 & 7
\end{array}\right| \\
M_{1}=\left[\begin{array}{c}
C_{11} \\
C_{12} \\
C_{13}
\end{array}\right] \text { where } c_{11}, c_{12}, c_{13} \text { cofactor of row } 1 \\
M_{1}=\left[\begin{array}{c}
20 \\
-20 \\
40
\end{array}\right]=\left[\begin{array}{c}
1 \\
-1 \\
2
\end{array}\right] \text { the common factor can be taken out } \\
\text { For } \lambda_{2}=-3,\left[\lambda_{i} I-A\right]=\left[\begin{array}{ccc}
-3 & -2 & 0 \\
-4 & -3 & -1 \\
48 & 34 & 6
\end{array}\right] \\
M_{2}=\left[\begin{array}{c}
C_{11} \\
C_{12} \\
C_{13}
\end{array}\right]=\left[\begin{array}{c}
16 \\
-24 \\
8
\end{array}\right]=\left[\begin{array}{c}
2 \\
-3 \\
1
\end{array}\right]
\end{gathered}
$$

$$
\begin{gathered}
\text { For } \lambda_{3}=-4\left[\lambda_{i} I-A\right]=\left[\begin{array}{ccc}
-4 & -2 & 0 \\
-4 & -4 & -1 \\
48 & 34 & 5
\end{array}\right] \\
M_{3}=\left[\begin{array}{l}
C_{11} \\
C_{12} \\
C_{13}
\end{array}\right]=\left[\begin{array}{c}
14 \\
-28 \\
56
\end{array}\right]=\left[\begin{array}{c}
1 \\
-2 \\
4
\end{array}\right]
\end{gathered}
$$

$M_{1}, M_{2}$ and $M_{3}$ are the eigen vectors corrsesponding to the eigen values

$$
\lambda_{1}, \lambda_{2} \text { and } \lambda_{3}
$$

(d) The modal matrix is

$$
M=\left[M_{1}: M_{2}: M_{3}\right]=\left[\begin{array}{ccc}
1 & 2 & 1 \\
-1 & -3 & -2 \\
-2 & 1 & 4
\end{array}\right]
$$

Let us prove $M^{-1} A M=$ is adiagonal matrix

$$
\begin{gathered}
M^{-1}=\frac{\operatorname{Adj}[M]}{|M|}=\frac{[\text { cofactor of } M]^{T}}{|M|} \\
A d j[M]=\left[\begin{array}{ccc}
-10 & 8 & -7 \\
-7 & 6 & -5 \\
-1 & 1 & -1
\end{array}\right]^{T}=\left[\begin{array}{ccc}
-10 & -7 & -1 \\
8 & 6 & 1 \\
-7 & -5 & -1
\end{array}\right] \\
|M|=\left[\left.\begin{array}{ccc}
1 & 2 & 1 \\
-1 & -3 & -2 \\
-2 & 1 & 4
\end{array} \right\rvert\,=-1, \quad \frac{\operatorname{Adj}[M]}{|M|}=\left[\begin{array}{ccc}
10 & 7 & 1 \\
-8 & -6 & -1 \\
7 & 5 & 1
\end{array}\right]\right. \\
A M=\left[\begin{array}{ccc}
0 & 2 & 0 \\
4 & 0 & 1 \\
-48 & -34 & -9
\end{array}\right]\left[\begin{array}{ccc}
1 & 2 & 1 \\
-1 & -3 & -2 \\
-2 & 1 & 4
\end{array}\right]=\left[\begin{array}{ccc}
-2 & -6 & -4 \\
2 & 9 & 8 \\
4 & -3 & -16
\end{array}\right] \\
M^{-1} A M=\left[\begin{array}{ccc}
10 & 7 & 1 \\
-8 & -6 & -1 \\
7 & 5 & 1
\end{array}\right]\left[\begin{array}{ccc}
-2 & -6 & -4 \\
2 & 9 & 8 \\
4 & -3 & -16
\end{array}\right]=\left[\begin{array}{ccc}
-2 & 0 & 0 \\
0 & -3 & 0 \\
0 & 0 & -4
\end{array}\right] \\
=\left[\begin{array}{ccc}
\lambda_{1} & 0 & 0 \\
0 & \lambda_{2} & 0 \\
0 & 0 & \lambda_{3}
\end{array}\right]=\Delta
\end{gathered}
$$

### 3.1.4 MIMO System

For multiple inputs multiple output systems a single transfer function does not exist. There exists a mathematical relationship between each output and all the inputs. Hence for such systems there exists a transfer matrix rather than transfer function but method of obtaining transfer matrix remains same as before.

8 Determine the transfer matrix for MIMO system given by.Example 3

$$
\begin{gathered}
{\left[\begin{array}{l}
x_{1}{ }_{1} \\
x_{2}
\end{array}\right]=\left[\begin{array}{cc}
0 & 3 \\
-2 & -5
\end{array}\right]\left[\begin{array}{l}
x_{1} \\
x_{2}
\end{array}\right]=\left[\begin{array}{ll}
1 & 1 \\
1 & 1
\end{array}\right]\left[\begin{array}{l}
U_{1} \\
U_{2}
\end{array}\right]} \\
{\left[\begin{array}{l}
Y_{1} \\
Y_{2}
\end{array}\right]=\left[\begin{array}{ll}
2 & 1 \\
1 & 0
\end{array}\right]\left[\begin{array}{l}
x_{1} \\
x_{2}
\end{array}\right]}
\end{gathered}
$$

From given state model

$$
\begin{aligned}
& \mathrm{A}=\left[\begin{array}{cc}
0 & 3 \\
-2 & -5
\end{array}\right], \mathrm{B}=\left[\begin{array}{ll}
1 & 1 \\
1 & 1
\end{array}\right], \mathrm{C}=\left[\begin{array}{ll}
2 & 1 \\
1 & 0
\end{array}\right], \mathrm{D}=[0] \\
& \mathrm{T} . \mathrm{M}=C[S I-A]^{-1} B+D
\end{aligned}
$$

$$
[S I-A]=\left[\begin{array}{cc}
s & -3 \\
2 & s+3
\end{array}\right]
$$

$$
\operatorname{Adj}[S I-A]=\left[\begin{array}{ll}
C_{11} & C_{12} \\
C_{21} & C_{22}
\end{array}\right]^{T}=\left[\begin{array}{cc}
s+5 & -2 \\
3 & s
\end{array}\right]^{T}=\left[\begin{array}{cc}
s+5 & 3 \\
-2 & s
\end{array}\right]
$$

$$
|S I-A|=s^{2}+5 s+6 s=(s+2)(s+3)
$$

$$
\therefore[S I-A]^{-1}=\frac{\operatorname{Adj}[S I-A]}{|S I-A|}=\frac{\left[\begin{array}{cc}
s+5 & 3 \\
-2 & s
\end{array}\right]}{(s+2)(s+3)}
$$

$$
\therefore \mathrm{T} . \mathrm{M}=C[S I-A]^{-1} B=\frac{\left[\begin{array}{ll}
2 & 1 \\
1 & 0
\end{array}\right]\left[\begin{array}{ll}
s+5 & 3 \\
-2
\end{array}\right]\left[\begin{array}{ll}
1 & 1 \\
1 & 1
\end{array}\right]}{(s+2)(s+3)}
$$

$$
=\frac{\left[\begin{array}{ll}
2 & 1 \\
1 & 0
\end{array}\right]\left[\begin{array}{ll}
s+8 & s+3 \\
s-2 & s-2
\end{array}\right]}{(s+2)(s+3)}=\frac{\left[\begin{array}{cc}
3 s+14 & 3 s+14 \\
s+8 & s+8
\end{array}\right]}{(s+2)(s+3)}
$$

### 3.1.5 Gilbert's Test for controllability

For Gilbert's test it is necessary that the matrix A must be in canonical form .Hence the given state model required to be transformed to the canonical form first to apply the Gilbert test

Consider single input linear time invariant system represented by

$$
X^{\prime}=A X(t)+B U(t)
$$

Where A is not in the canonical form then it can be transformed to be canonical form by the transformation $\quad X(t)=M Z(t) \quad$ Where $\mathrm{M}=$ model matrix. The transformed state model, as derived earlier take the form

$$
Z^{\prime}(t)=\nabla Z(t)+B^{\sim} U(t)
$$

Where $\nabla=$ diagonal matrix

$$
B^{\sim}=M^{-1} B
$$

It is assumed that the Eigen values of A are distinct in such a case the necessary and sufficient condition for are complete state controllability is that the vector matrix $B^{\sim}$ should not have any zero elements If it has zero elements then the corresponding state variables are not controllable if the Eigen values are repeated then matrix A cannot be transformed to Jordan results Jordan canonical form, as


In such a case the condition for complete state controllability is that the elements of any row of $B^{\sim}$ that corresponds Jordan to the last row of each Jordan block are not all zero

Example 3.9 consider the system with state equation

$$
\left[\begin{array}{l}
\dot{X}_{1} \\
\dot{X}_{2} \\
\dot{X}_{3}
\end{array}\right]=\left[\begin{array}{ccc}
0 & 1 & 0 \\
0 & 0 & 1 \\
-6 & -11 & -6
\end{array}\right]\left[\begin{array}{l}
X_{1} \\
X_{2} \\
X_{3}
\end{array}\right]+\left[\begin{array}{l}
0 \\
0 \\
1
\end{array}\right] u(t)
$$

Estimate the state controllability by (i) kalman,s test and (ii) Gilbert's test
(i) Kaman,s test

$$
\begin{gathered}
M=\left[B: A B: A^{2} B\right] \ldots N=3 \\
A B=\left[\begin{array}{ccc}
0 & 1 & 0 \\
0 & 0 & 1 \\
-6 & -11 & -6
\end{array}\right]\left[\begin{array}{l}
0 \\
0 \\
1
\end{array}\right]=\left[\begin{array}{c}
0 \\
1 \\
-6
\end{array}\right] \\
A^{2}=\left[\begin{array}{ccc}
0 & 1 & 0 \\
0 & 0 & 1 \\
-6 & -11 & -6
\end{array}\right]\left[\begin{array}{ccc}
0 & 1 & 0 \\
0 & 0 & 1 \\
-6 & -11 & -6
\end{array}\right]=\left[\begin{array}{ccc}
0 & 0 & 1 \\
-6 & -11 & -6 \\
36 & 60 & 25
\end{array}\right] \\
A^{2} B=\left[\begin{array}{ccc}
0 & 0 & 1 \\
-6 & -11 & -6 \\
36 & 60 & 25
\end{array}\right]\left[\begin{array}{l}
0 \\
0 \\
1
\end{array}\right]=\left[\begin{array}{c}
1 \\
-6 \\
25
\end{array}\right], M=\left[\begin{array}{ccc}
0 & 0 & 1 \\
0 & 1 & -6 \\
1 & -6 & 25
\end{array}\right] \\
=1, \text { thus }|M| \text { is nonsingular }
\end{gathered}
$$

Hence the rank of M is 3 which is n thus the system is completely state controllable
(ii) Gilbert test. For this it is necessary to express A in the canonical form find the Eigen values of A

$$
\begin{gathered}
{\left[\lambda_{i} I-A\right]=\left[\begin{array}{ccc}
\lambda & -1 & 0 \\
0 & \lambda & -1 \\
6 & 11 & \lambda+6
\end{array}\right]=0, \lambda^{3}+6 \lambda^{2}+11 \lambda+6=0} \\
(\lambda+1)(\lambda+2)(\lambda+3)=0, \lambda_{1}=-1, \quad \lambda_{2}=-2, \lambda_{3}=-3 \\
M=\left[\begin{array}{ccc}
1 & 1 & 1 \\
\lambda_{1} & \lambda_{2} & \lambda_{3} \\
\lambda_{1}{ }^{2} & \lambda_{2}{ }^{2} & \lambda_{3}{ }^{2}
\end{array}\right]=\left[\begin{array}{ccc}
1 & 1 & 1 \\
-1 & -2 & -3 \\
1 & 4 & 9
\end{array}\right]
\end{gathered}
$$

$$
\begin{gathered}
M^{-1}=\frac{\operatorname{Adj}[M]}{|M|}=\frac{\left[\begin{array}{ccc}
-6 & 6 & -2 \\
-5 & 8 & -3 \\
-1 & 2 & -1
\end{array}\right]^{T}}{-2}=\left[\begin{array}{ccc}
3 & 2.5 & 0.5 \\
-3 & -4 & -1 \\
1 & 1.5 & 0.5
\end{array}\right] \\
B^{\sim}=M^{-1} B=\left[\begin{array}{ccc}
3 & 2.5 & 0.5 \\
-3 & -4 & -1 \\
1 & 1.5 & 0.5
\end{array}\right]=\left[\begin{array}{c}
0.5 \\
-1 \\
0.5
\end{array}\right]
\end{gathered}
$$

As matrix A is in phase variable form the model matrix M is Vander monde matrix.As none of the elements of $B^{\sim}$ are zero the system is completely state controllable. As Gilbert's test requires transforming matrix A into canonical form it is time consuming and hence kalman, s test is popularly used to test controllability

### 3.1.6 Output controllability

In the practical design of a control system, we may want to control the output rather than the state of the system. Complete state controllability. is neither necessary nor sufficient for controlling the output of the system for this reason, it is desirable to define separately complete output controllability.
Consider the system described by
14).

$$
\begin{align*}
& x=\mathrm{Ax}+\mathrm{Bu} \\
& Y=C X+\mathrm{Du} \tag{3.15}
\end{align*}
$$

Where $\mathbf{x}=$ state vector ( n -vector)

$$
\mathrm{U}=\text { control vector(r-vector) }
$$

$$
Y=\text { output control (m-vector) }
$$

$$
\mathrm{A}=\mathrm{n} \times \mathrm{n} \text { matrix }
$$

$$
\mathrm{B}=\mathrm{n} \times r \text { matrix }
$$

$$
C=m \times \text { nmatrix }
$$

$$
D=m \times r \text { matrix }
$$

The system described by Equations (3.14) and (3.15) is said to be completely output. Controllable if it is possible to construct an unconstrained control vector $\mathbf{u}$ (t) that will. Transfer any given initial output $\mathbf{y A t} 0 \mathrm{~B}$ to any final output $\mathbf{y A t 1 B}$ in a finite time interval $t_{0} \leq \mathrm{t} \leq t_{1}$, It can be proved that the condition for complete output controllability is as follows:

15 ) is completely output .14) and (3.The system described by Equations (3 controllable. If and only if the $\mathrm{m} \times(\mathrm{n}+1) \mathrm{r}$ matrix
$\left[C: C A B \vdots C A^{2} B: \cdots \cdots \vdots C A^{n-1} B: D\right]$

### 4.1.7 Uncontrollable System.

An uncontrollable system has a subsystem that is physically disconnected from the, Input

### 3.2 OBSERVABILITY

We discuss the observability of linear systems. Consider the unforced System described by the following equations

$$
\begin{equation*}
\text { 16). } x=\mathrm{Ax} \tag{3}
\end{equation*}
$$

$$
\begin{equation*}
Y=C X \tag{3.17}
\end{equation*}
$$

Where $\mathbf{x}=$ state vector ( n -vector)

$$
Y=\text { output control(m-vector) }
$$

$\mathrm{A}=\mathrm{n} \times \mathrm{n}$ matrix
$\mathrm{C}=\mathrm{m} \times n$ matrix
The system is said to be completely observable if every state $\mathbf{x}\left(t_{0}\right)$ can be determined from the observation of $\mathbf{y}(\mathrm{t})$ over a finite time interval, $t_{0} \leq \mathrm{t} \leq$ $t_{1}$

The system is therefore completely observable if every transition of the state eventually affects every element, of the output vector. The concept of observability is useful in solving the problem of reconstructing unmeasurable state variables from measurable variables in the minimum possible length of time we treat only linear, time-invariant systems.

Therefore, without loss of generality, we can assume that $t_{0}=0$ the concept of observability is very important because, in practice, the difficulty. Encountered with state feedback control is that some of the state variables are not. Accessible for direct measurement, with the result that it becomes necessary to estimate the unmeasurable state variables in order to construct the control signals. It will be. Shown in that such estimate of state variables are possible if and only if the system is completely observable. In discussing observability conditions, we consider the unforced system as given By 17).The reason for this is as follows: If the system is .16) and (3.Equations (3 described

$$
\begin{aligned}
x= & A x+B U \\
& Y=C X+D U
\end{aligned}
$$

Then
$\mathrm{X}(\mathrm{t})=e^{A t} \mathrm{X}(0)+\int_{0}^{t} e^{A(t-\tau)} B u(\tau) d \tau$
And $y(t)$ is

$$
\mathrm{Y}(\mathrm{t})=e^{A t} \mathrm{x}(0)+\int_{0}^{t} e^{A(t-\tau)} B u(\tau) d \tau+D U
$$

Since the matrices $\mathrm{A}, \mathrm{B}, \mathrm{C}$, and D are known and $\mathrm{u}(\mathrm{t})$ is also known, the last two terms, on the right-hand side of this last equation are known quantities. Therefore they may be subtracted from the observed value of $\mathbf{y}(\mathrm{t})$. Hence, for investigating a necessary and sufficient condition for complete

Observability, it suffices to consider the system described by Equations
17).16) and (3.(3

### 3.2.1 Complete Observability of Continuous-Time Systems.

17).The output .16) and (3. Consider the system described by Equations (3 vector $\mathrm{Y}(\mathrm{t})$ is
$\mathrm{Y}(\mathrm{t})=\mathrm{C} e^{A t} \mathrm{X}(0)$.
12), we have.10) or (3.Referring to Equation (3

$$
e^{A t}=\sum_{k=0}^{n-1} \alpha_{k}(\tau) A^{k}
$$

Where n is the degree of the characteristic polynomial. Note that Equations (3.8) and (3.10) with m replaced by n can be derived using the characteristic polynomial. Hence we obtain.
$y(t)=\sum_{k=0}^{n-1} \alpha_{k}(\tau) C A^{k} X(0)$
$y\left(t=\alpha_{0}(t) C X(0)+\alpha_{1}(t) C A X(0)+\cdots+\alpha_{n-1}(t) C A^{n-1} X(0\right.$
If the system is completely observable, then, given the output $\mathbf{y}(\mathrm{t})$ over a (3.18) .time interval $t_{0} \leq \mathrm{t} \leq t_{1} \mathbf{X}(0)$ is uniquely determined from Equation It can be shown that this requires the rank of the $n m \times n$ matrix

$$
\left[\begin{array}{c}
C \\
C A \\
\vdots \\
\vdots \\
C A^{n-1}
\end{array}\right]
$$

From this analysis, we can state the condition for complete observability as follows the system described by Equations (3.16)and 3.17) is completely observable if and only if the $\mathrm{n} \times \mathrm{nm}$ matrix

$$
\left[C^{T}: A^{T} C^{T}: \cdots \cdots:\left(A^{T}\right)^{n-1} C^{T}\right]
$$

Is of rank n or has n linearly independent column vectors. This matrix is called the.observability matrix

### 10.3 Example

consider the system .Is this system controllable and observable. Since the rank of the matrix

$$
\begin{gathered}
{\left[\begin{array}{l}
x_{1} \\
x_{2}
\end{array}\right]=\left[\begin{array}{cc}
1 & 1 \\
-2 & -1
\end{array}\right]\left[\begin{array}{l}
x_{1} \\
x_{2}
\end{array}\right]+\left[\begin{array}{l}
0 \\
1
\end{array}\right]} \\
\mathrm{Y}=\left[\begin{array}{ll}
1 & 0
\end{array}\right]\left[\begin{array}{l}
x_{1} \\
x_{2}
\end{array}\right] \\
{[B \vdots A B]=\left[\begin{array}{cc}
0 & 1 \\
1 & -1
\end{array}\right]}
\end{gathered}
$$

Is 2 , the system is completely state controllable. For output controllability, let us find the rank of the matrix $[C B: C A B]$ since

$$
[C B: C A B]=\left[\begin{array}{ll}
1 & 0
\end{array}\right]
$$

The rank of this matrix is 1 . Hence, the system is completely output controllable. To test the observability condition, examine the rank of

$$
\left[C^{T}: A^{T} C^{T}\right]
$$

Since

$$
\left[C^{T}: A^{T} C^{T}\right]=\left[\begin{array}{ll}
1 & 1 \\
0 & 1
\end{array}\right]
$$

The rank of $\left[C^{T}: A^{T} C^{T}\right]$ is 2 . Hence, the system is completely observable

### 3.2.2 Conditions for Complete Observability in the Plane

The conditions for complete observability can also be stated in terms of transfer functions or transfer matrices. The necessary and sufficient conditions for complete observability are that no cancellation, occur in the transfer function or transfer matrix. If cancellation occurs, the canceled, mode cannot be observed in the output

## Example 3:11

Show that the following system is not completely observable.

$$
\begin{gathered}
X=A X+B U \\
Y=C
\end{gathered}
$$

Where
$\mathrm{X}=\left[\begin{array}{l}X_{1} \\ X_{2} \\ X_{3}\end{array}\right], \quad \mathrm{A}=\left[\begin{array}{ccc}0 & 1 & 0 \\ 0 & 0 & 1 \\ -6 & -11 & -6\end{array}\right], \quad \mathrm{B}=\left[\begin{array}{l}0 \\ 0 \\ 1\end{array}\right], \quad \mathrm{C}=\left[\begin{array}{lll}4 & 5 & 1\end{array}\right]$
Note that the control function $u$ does not affect the complete observability of the system to examine complete observability, we may simply set $u=0$. For this system, we have

$$
\left[C^{T}: C^{T} A^{T}:\left(A^{T}\right)^{2} C^{T}\right]=\left[\begin{array}{ccc}
4 & -6 & 6 \\
5 & -7 & 5 \\
1 & -1 & -1
\end{array}\right]
$$

Note that

$$
\left|\begin{array}{ccc}
4 & -6 & 6 \\
5 & -7 & 5 \\
1 & -1 & -1
\end{array}\right|=0
$$

Hence, the rank of the matrix $\left[C^{T}: C^{T} A^{T}: \cdots \cdots\left(A^{T}\right)^{2} C^{T}\right]$ is less than3.Therefore, the system is not, completely observable.
Comments. The transfer function has no cancellation if and only if the system is completely, state controllable and completely observable. This means that the canceled transfer, function does not carry along all the information characterizing the dynamic system.

### 3.2.3 Alternative Form of the Condition for Complete Observability.

Consider the System described by Equations (13) and (14), rewritten

$$
\begin{equation*}
X=A X \tag{3.19}
\end{equation*}
$$

$$
\begin{equation*}
Y=C X \tag{2.20}
\end{equation*}
$$

Suppose that the transformation matrix $\mathbf{P}$ transforms $\mathbf{A}$ into a diagonal matrix, or

$$
P^{-1} A P=D
$$

Where $\mathbf{D}$ is a diagonal matrix let us define

$$
X=P Z
$$

Then Equations (16) and (17) can be written

$$
\begin{aligned}
Z & =P^{-1} A P Z=D Z \\
& Y=C P Z
\end{aligned}
$$

Hence

$$
Y(t)=C P e^{A t} Z(0)
$$

The system is completely observable if (1) no two Jordan blocks in $\mathbf{J}$ are associated with, the same eigenvalues, (2) no columns of $\mathbf{C S}$ that correspond to the first row of each. Jordan block consist of zero elements, and (3) no columns of $\mathbf{C S}$ that correspond to, distinct eigenvalues consist of zero elements

### 3.2.4 Gilbert test observability

It is known that for Gilbert test the state model must be expressed in the canonical form; consider the state model of linear time invariant system, as

$$
X^{\prime}(t)=A X(t)+B U(t) \text { and } Y(t)=C X(t)
$$

Use the transformation $X(t)=M Z(t)$ where $M \equiv$ model matrix

$$
\therefore \quad Y(t)=C M Z(t)=C^{\sim} Z(t) \quad \text { where } C^{\sim}=C M
$$

For a single input single output system

$$
\begin{aligned}
Y(t) & =C^{\sim} Z(t)=\left[\begin{array}{llll}
C^{\sim}{ }_{11} & C^{\sim}{ }_{12} & \cdots & C^{\sim}{ }_{1 n}
\end{array}\right]\left[\begin{array}{c}
Z_{1}(t) \\
Z_{2}(t) \\
\vdots \\
z_{n}(t)
\end{array}\right] \\
& =C^{\sim}{ }_{11} Z_{1}(t) \\
C^{\sim}{ }_{12} Z_{2}(t) & \cdots
\end{aligned} C^{\sim}{ }_{1 n}{ }_{1 n} z_{n}(t) \quad \$
$$

Due to the canonical form all the state are decoupled and not like the each other. Hence for the system to be observable each term corresponding to each state must be observed in the output Hence none of the coefficient of the $C^{\sim}$ must be zero thus the system is the complete observable if all the coefficients of $C^{\sim}$ are nonzero coefficient none of the coefficient is zero if any element is zero the corresponding state remains unobservable i.e shielded from observation

Example 3.12 evaluate the observability of the system with using Gilbert's test

$$
A=\left[\begin{array}{ccc}
0 & 1 & 0 \\
0 & 0 & 1 \\
0 & -2 & -3
\end{array}\right], B=\left[\begin{array}{l}
0 \\
0 \\
1
\end{array}\right] \text { and } C=\left[\begin{array}{lll}
3 & 4 & 1
\end{array}\right]
$$

For Gilbert's test find the Eigen values

$$
\begin{gathered}
{[\lambda I-A]=\left|\begin{array}{ccc}
\lambda & -1 & 0 \\
0 & \lambda & -1 \\
0 & 2 & \lambda+3
\end{array}\right|=0 \quad, \lambda^{3}+3 \lambda^{2}+2 \lambda=0} \\
\lambda\left(\lambda^{2}+3 \lambda+2\right)=0 \quad \therefore \lambda(\lambda+1)(\lambda+2)=0 \\
\lambda_{1}=0, \lambda_{2}=-1, \lambda_{3}=-2
\end{gathered}
$$

As the Eigen values are distinct the model matrix $M$ is the Vander Monde matrix

$$
M=\left[\begin{array}{ccc}
1 & 1 & 1 \\
\lambda_{1} & \lambda_{2}{ }^{2} & \lambda_{1} \\
\lambda_{1}{ }^{2} & \lambda_{2}{ }^{2} & \lambda_{3}{ }^{2}
\end{array}\right]=\left[\begin{array}{ccc}
1 & 1 & 1 \\
0 & -1 & -2 \\
0 & 1 & 4
\end{array}\right]
$$

When the model transformed form

$$
C^{\sim}=C M=\left[\begin{array}{lll}
3 & 4 & 1
\end{array}\right]\left[\begin{array}{ccc}
1 & 1 & 1 \\
0 & -1 & -2 \\
0 & 1 & 4
\end{array}\right]=\left[\begin{array}{lll}
3 & 0 & -1
\end{array}\right]
$$

As there is one zero elements in $C^{\sim}$ the system is not complete observsble

Example 3.13 Use controllability and observability matrices to determine whether the system represented by the flow graph shown the figure is completely controllability and completely observability


The value of variable at the node is an algebraic sum of all signals entering at the node

$$
\begin{array}{ll}
\therefore & \dot{x}_{3}=\mathrm{x}_{1}+\mathrm{x}_{2}+2 \mathrm{u}, \quad \dot{\mathrm{X}}_{2}=\mathrm{X}_{1}-2 \mathrm{x}_{2}+\mathrm{x}_{3}+2 \mathrm{u} \\
& \dot{\mathrm{X}}_{1}=-3 \mathrm{x}_{1}+\mathrm{x}_{2}, \quad \mathrm{Y}=\mathrm{x}_{1}+2 \mathrm{X}_{2}-\mathrm{X}_{3} \\
\therefore & \mathrm{~A}=\left[\begin{array}{ccc}
-3 & 1 & 0 \\
1 & -2 & 1 \\
1 & 1 & 0
\end{array}\right], \mathrm{B}=\left[\begin{array}{l}
0 \\
2 \\
2
\end{array}\right], \mathrm{C}=\left[\begin{array}{ll}
1 & 2
\end{array}-1\right]
\end{array}
$$

For controllability $M_{=\left[\begin{array}{lll}B: & A B: & A^{2} B\end{array}\right] \text { for } n=3}$

$$
\begin{aligned}
A B=\left[\begin{array}{ccc}
-3 & 1 & 0 \\
1 & -2 & 1 \\
1 & 1 & 0
\end{array}\right]\left[\begin{array}{l}
0 \\
2 \\
3
\end{array}\right] & =\left[\begin{array}{c}
2 \\
-2 \\
2
\end{array}\right] \\
A^{2}=\left[\begin{array}{ccc}
-3 & 1 & 0 \\
1 & -2 & 1 \\
1 & 1 & 0
\end{array}\right]\left[\begin{array}{ccc}
-3 & 1 & 0 \\
1 & -2 & 1 \\
1 & 1 & 0
\end{array}\right] & =\left[\begin{array}{ccc}
10 & -5 & 1 \\
-4 & 6 & -2 \\
-2 & -1 & 1
\end{array}\right]
\end{aligned}
$$

$$
A^{2} B=\left[\begin{array}{ccc}
10 & -5 & 1 \\
-4 & 6 & -2 \\
-2 & -1 & 1
\end{array}\right]\left[\begin{array}{l}
0 \\
2 \\
2
\end{array}\right]=\left[\begin{array}{c}
-8 \\
8 \\
0
\end{array}\right]
$$

$$
M=\left[\begin{array}{ccc}
0 & 2 & -8 \\
2 & -2 & 8 \\
2 & 2 & 0
\end{array}\right] \quad \therefore|M|=-32 \neq 0 \text { hence rank } n=3=n
$$

The system is fully state controllable
For observablility $M_{o}\left[C^{T}: A^{T} C^{T}: A^{T^{2}} C^{T}\right], n=3$

$$
\begin{gathered}
A^{T} C^{T}=\left[\begin{array}{ccc}
-3 & 1 & 1 \\
1 & -2 & 1 \\
0 & 1 & 0
\end{array}\right]\left[\begin{array}{c}
1 \\
2 \\
-1
\end{array}\right]=\left[\begin{array}{c}
-2 \\
-4 \\
2
\end{array}\right] \\
A^{T^{2}} C^{T}=\left[\begin{array}{ccc}
-3 & 1 & 1 \\
1 & -2 & 1 \\
0 & 1 & 0
\end{array}\right]\left[\begin{array}{c}
-2 \\
-4 \\
2
\end{array}\right]=\left[\begin{array}{c}
4 \\
4 \\
-4
\end{array}\right] \\
\therefore M_{o}=\left[\begin{array}{ccc}
1 & -2 & 4 \\
2 & -4 & 8 \\
-1 & 2 & -4
\end{array}\right] \quad,\left|M_{o}\right|=0 \text { and }\left|\begin{array}{cc}
2 & -4 \\
-1 & 2
\end{array}\right|=0
\end{gathered}
$$

Thus $2 \times 2$ determinant is zero hence the system is not observable

### 3.3 Principle of Duality

We shall now discuss the relationship between controllability and observability .We shall introduce the principle of duality, due to Kaman, to clarify apparent analogies between controllability and observability.
Consider the system $S_{1}$ described by

$$
\begin{array}{r}
x=\mathrm{Ax}+B u \\
Y=C X
\end{array}
$$

Where $\mathbf{x}=$ state vector ( n -vector)

$$
\mathrm{U}=\text { control vector(r-vector) }
$$

$$
Y=\text { output control }(m \text {-vector })
$$

$$
\mathrm{A}=\mathrm{n} \times \mathrm{n} \text { matrix }
$$

$$
\mathrm{B}=\mathrm{n} \times r \text { matrix }
$$

$$
C=m \times \text { nmatrix }
$$

$$
D=m \times r \text { matrix }
$$

And the dual system $S_{2}$ defined by

$$
\begin{gathered}
Z \cdot=A^{*} Z+C^{*} V \\
n=B^{*} Z
\end{gathered}
$$

Where
$\mathrm{Z}=$ State vector ( n -vector)
$\mathrm{V}=$ control vector (m-vector)
$\mathrm{n}=$ output vector(r-vector)
$A^{*}=$ conjugate transpose of $A$
$B^{*}=$ conjugate transpose of $B$
$C^{*}=$ conjugate transpose of $C$
The principle of duality states that the system $S_{1}$ is completely state controllable.(Observable) if and only if system $S_{2}$ is completely observable
(state controllable).To verify this principle, let us write down the necessary and sufficient conditions for complete state controllability and complete Observability for systems $S_{1}$ and $S_{2}$

## For system $S_{1}$ :

1. A necessary and sufficient condition for complete state controllability is that the. Rank of the $n \times n r$ matrix

$$
\left[B: A B: \cdots A^{n-1} B\right]
$$

Be n
2. A necessary and sufficient condition for complete observability is that the rank of the $n \times n m$ matrix

$$
\left[C^{T}: A^{T} C^{T} \vdots \cdots \cdots: A^{T^{n-1}} C^{T}\right]
$$

Be n:

## For system $S_{2}$

A. A necessary and sufficient condition for complete state controllability is that the. Rank of the $n \times n m$ matrix

$$
\left[C^{T} \vdots A^{T} C^{T} \vdots \cdots \cdots \vdots\left(A^{T}\right)^{n-1} C^{T}\right]
$$

Ben
B. A necessary and sufficient condition for complete observability is that the rank of the $n \times n r$ matrix

$$
\left[B: A B: \cdots A^{n-1} B\right]
$$

Ben.
By comparing these conditions, the truth of this principle is apparent. By use of this, principle, the observability of a given system can be checked by testing the state controllability, of its dual

### 3.4 Detectability.

For a partially observable system, if the unobservable modes are, stable and the observable modes are unstable, the system is said to be detectable. Note, that the concept of detectability is dual to the concept of stabilizability

## Chapter Four

## Bounded controls

## 4. Introduction

Time delay often occurs in engineering system. Since the existence of time delay usually instability of the system the study on the Time delay systems has received considerable attentions and many stability criteria for time delay systems can be found in the literature. Stability criteria for time delay systems. Tend to fall into one of the two categories: delay independent and delay dependent .As the name implies delay independent criteria provide conditions which guarantee stability for any length of the time delay on the other hand delay dependent criteria exploit a priori knowledge of upper bounds on the a mount of time delay, These criteria since more information about the Time delay is assumed to be known

## 4.1 stability

## Definition of stability 4.1

A continuous system (discrete time system) is stable if its impulse response $y_{f}(t)$ (koneke delta response $y_{f}(t)$ approaches zero at time approaches infinity.

Alternatively the definition of a stable system can be upon the response of the system to bounded inputs that is inputs whose magnitudes are less than some finite value for all time.

## Other definition of stability 4.2

A continuous or discrete time system is stable if every bounded input produces a bounded output (BIBO) also definition. The system is stable if
$\mathrm{f}(\mathrm{t}) \rightarrow \infty \quad$ as $\quad \mathrm{t} \rightarrow \infty$


The system is an unstable if $\mathrm{f}(\mathrm{t}) \rightarrow \infty \quad$ as $\quad \mathrm{t} \rightarrow \infty$


The system is a marginally stable $f(t)$ does not decay to 0 or go to $\infty$ as $t \rightarrow \infty$


Consideration of degree of stability of a system often provides valuable information about its behavior. That is if it stable how close is it to being unstable and Definition of stability


Poles on the iw axis: marginally stable in above figure we define stability the closed loop transfer function left hand side (BIBO).

### 4.1.1 Characteristic Root locations for continuous systems

A major result of chapter 3 is that the impulse response of a linear time invariant continuous system is a sum of exponential time functions. Whose exponents are the roots of the system characteristic equation A necessary and sufficient condition for the system to be stable is that real parts of the roots of the characteristic equation have negative real parts. This ensures that the impulse response will decay exponentially with time. If the system has some roots with real parts equal to zero but none with positive real parts the system is said to be marginally stable. In this instance the impulse response does not decay to zero although it is bounded but certain other inputs with produce unbounded outputs therefore marginally stable system unstable.

## Example 4.1

Differential Equation
$\left(S^{2}+1\right)$ y $(S)=U(S)$
Has the characteristic equation
$S^{2}+1=0$
This equation has the two roots $\pm \mathrm{j}$
Since these roots have zero real parts the system is not stable it is however marginally stable since the equation has no roots with positive real parts. In response to most input or output will contain term of the form $y=t$ constant This is unbounded

### 4.1.2 Routh stability criterion

The Routh criterion is a method of determining continuous system stability for system with an nth order characteristic equation of the form
$a_{n} s^{n}+a_{n-1} s^{n-1}+\ldots \ldots \ldots \ldots \ldots \ldots \ldots \ldots . .+a_{1} s+a_{0}=0$
The criterion is applied using a Routh table defined as follows


Where $a_{n}, a_{n-1}, \ldots \ldots \ldots \ldots, a_{0}$ are the coefficient of the characteristic equation and

$$
b_{1=\frac{a_{n-1} a_{n-2}-a_{n} a_{n-3}}{a_{n-1}}}^{a_{n-1}}
$$

$$
\begin{array}{cc}
b_{2=} \frac{a_{n-1} a_{n-4}-a_{n} a_{n-5}}{a_{n-1}} & \text { etc } \\
C_{1=} \frac{b_{1} a_{3}-a_{n-1} b_{2}}{b_{1}} \quad, c_{2}=\frac{b_{1} a_{n-5}-a_{n-1} b_{3}}{b_{1}} \quad \text { etc }
\end{array}
$$

The table is continued horizontally and vertically unit only zero are obtained before the next row is computed without disturbing the properties of the table.

The Routh criterion .all the roots of the characteristic if the elements of the first column of the Routh table have the same sign. Otherwise the number of roots with positive real parts is equal to the number of change of sign

## Example 4.2

$S^{3}+6 S^{2}+12 S+8=0$


Since there are no changes of sign in the first column of the table all the roots of the equation have negative real parts. Often it is desirable to determine arrange of value of a particular system parameter for which the system is stable. This can be accomplished by writing the inequalities the ensure that there is no change of sign in the first column of the Routh table for the system

These inequalities then specify the range of allowable values of the parameter.

For no sign change in the first column it is necessary that the condition

$$
1+K>0 ، 8-K>0
$$

Be satisfied .thus the characteristic equation has roots with negative real parts if $-1<\mathrm{k}<8$, the simultaneous solution of these two inequalities. a row of zero the $\grave{S}$ row of the Routh Stable indicates that the polynomial has a pair of roots which satisfy the auxiliary equation formed as flows
$\mathrm{A} S^{2}+\mathrm{B}=0$
Where A and B are the first and second elements of the $S^{2}$ row zero in the To continue the stable the zero in the $\grave{S}$ row are replaced with the coefficients of the derivative of auxiliary equation is $2 \mathrm{AS}+0=0$

The coefficients 2A and 0 are then entered into the Srow and the table is continued as described above

## Example 4.3

$s^{4}+2 s^{3}+3 s^{2}+4 s+5=0$
$s^{4}$
$s^{3}$
$s^{2}$
$s^{1}$

$s^{0}$ | 1 | 3 | 5 |
| :---: | :---: | :---: | :---: |
| $l$ |  |  |

The system unstable 2 poles RHP

## Special case 1

## Example4.4

$s^{3}+2 s^{2}+s+2=0$

```
lac
\[
\mathrm{E}=0.0001
\]
```

The system is stable

## Special case 2

## Example 4.5

$$
s^{5}+2 s^{4}+24 s^{3}+48 s^{2}-25 s-50=0
$$

| $s^{5}$ |  | 24 |  | -25 |
| :---: | :---: | :---: | :---: | :---: |
| $s^{4}$ |  | 2 | 48 | -50 |
| $s^{3}$ |  | 0 | $0_{98}$ | 0 |
| $s^{2}$ |  |  | -50 | o |
| $s^{1}$ | 112,7 |  |  | 0 |
| $s^{0}$ |  |  |  | -50 |

The system is unstable
I poles RHP

### 4.1.3 Hurwitz stability criterion

The Hurwitz criterion is another method for determining whether all the roots of the characteristic equation of a continuous system have negative real parts.

This criterion is applied using determents formed form the coefficients of the characteristic equation.

It is assumed that the first coefficient $a_{n}$ is positive the determinants $A_{i} I=$ 1,2,3,..................n-1

Are formed as the principal minor determinants of the determinant


The determinant are thus formed as follows
$\Delta_{1}=a_{n-1}$

$$
\begin{gathered}
\Delta_{2}=\left|\begin{array}{cc}
a_{n-1} & a_{n-3} \\
a_{n} & a_{n-2}
\end{array}\right|=a_{n-1} a_{n-2}-a_{n} a_{n-3} \\
\Delta_{3}=\left|\begin{array}{ccc}
a_{n-1} & a_{n-3} & a_{n-5} \\
a_{n} & a_{n-2} & a_{n-4} \\
0 & a_{n-1} & a_{n-3}
\end{array}\right|=a_{n-1} a_{n-2} a_{n-3}+a_{n} a_{n-1} a_{n-6}-a_{n} a_{n-3}^{2}-a_{n-4} a_{n-1}^{2}
\end{gathered}
$$

And so on up to $\Delta_{n-1}$
All the roots of the characteristic equation have negative real pats if and only
If $n=3$

$$
\Delta_{3}=\left|\begin{array}{ccc}
a_{2} & a_{0} & 0 \\
a_{3} & a_{1} & 0 \\
0 & a_{2} & a_{0}
\end{array}\right|=a_{2} a_{1} a_{0}-a_{0}{ }^{2} a_{3} \quad, \Delta_{2=}\left|\begin{array}{ll}
a_{2} & a_{0} \\
a_{3} & a_{1}
\end{array}\right|=a_{2} a_{1}-a_{3} a_{0}
$$

$\Delta_{i}>0$

$$
\mathrm{i}=1,2, \ldots \ldots \ldots \ldots \ldots . ., \mathrm{n}
$$

$\Delta_{1}=a_{2}$
Thus all the roots of the characteristic equation have negative real parts if

$$
\begin{gathered}
a_{2} a_{1} a_{0}-a_{0}^{2} a_{3}>0 \quad a_{2} a_{1}-a_{0} a_{1}>0 \quad a_{2}>0 \\
\Delta_{3}=\left|\begin{array}{c}
a_{n-1} a_{n-3} \\
a_{n} a_{n-2} a_{n-4} \\
\vdots \\
0 \\
\vdots \\
\vdots \\
a_{n-1} \\
\vdots
\end{array}\right|=\left|\begin{array}{ccr}
8 & 24 & 0 \\
1 & 14 & 0 \\
0 & 8 & 24
\end{array}\right|
\end{gathered}
$$

## Example4.6

Determinant if the characteristic equation below represents a stable or au stable system using the Hurwitz criterion method

$$
s^{3}+8 s^{3}+24 s+24=0
$$

$$
\Delta_{3}=\left|\begin{array}{ccc}
8 & 24 & 0 \\
1 & 14 & 0 \\
0 & 8 & 24
\end{array}\right|
$$

$$
\Delta_{3}=8(14 \times 24-0 \times 8)-24(1 \times 24-0)+0(1 \times 8-0)=
$$

$2688-576=211$

$$
\Delta_{2}=\left|\begin{array}{ll}
a_{2} & a_{0} \\
a_{3} & a_{1}
\end{array}\right|=\left|\begin{array}{ll}
8 & 24 \\
1 & 14
\end{array}\right|=88, \quad \Delta_{1=8}
$$

In this method the system is a stable iff $\Delta_{3}>0$
$\therefore \Delta_{3}=2112>0$ the system is stable
$\Delta_{1}=a_{1}=8$
$\therefore$ Each determinant is positive
$\therefore$ The system is a stable.

## Example4.7

For what range of values of K is system with following characteristic equation stable using Hurwitz method
$S^{2}+k s+2 k-1=0$
The Hurwitz determinant for this system

$$
\Delta_{2=}\left|\begin{array}{cc}
k & 0 \\
1 & 2-1 k
\end{array}\right|=2 k^{2}-k=k(2 k-1)
$$

$$
\Delta_{1}=k
$$

In order for this determinant stable positive
It is necessary that $\mathrm{K}>0$ an $2 \mathrm{~K}-1>0$
Thus the system is stable if $K>1 / 2$

## Example4.8

Determine the Hurwitz condition for stability of the following general fourth order characteristic equation assuming $a_{4}$ is positive
$a_{4} \mathrm{~S}^{4}+a_{3} \mathrm{~S}^{3}+a_{2} \mathrm{~S}^{2}+a_{1} \mathrm{~S}+a_{0}=0$
The Hurwitz determinants are

$$
\begin{aligned}
& \Delta_{4}=\left|\begin{array}{cccc}
a_{3} & a_{1} & 0 & 0 \\
a_{4} & a_{2} & a_{0} & 0 \\
0 & a_{3} & a_{1} & 0 \\
0 & a_{4} & a_{2} & a_{0}
\end{array}\right|=a_{3}\left(a_{2} a_{1} a_{0}-a_{3} a_{0}^{2}\right)-a_{1}^{2} a_{0} a_{4} \\
& \Delta_{3}=\left|\begin{array}{ccc}
a_{3} & a_{1} & 0 \\
a_{2} & a_{0} & 0 \\
0 & a_{2} & a_{0}
\end{array}\right|=a_{3} a_{2} a_{1}-a_{0} a_{3}^{2}-a_{4} a_{1}^{2}
\end{aligned}
$$

$\Delta_{1}=a_{3}$

$$
\Delta_{2}=\left|\begin{array}{ll}
a_{3} & a_{1} \\
a_{4} & a_{2}
\end{array}\right|=a_{3} a_{2}-a_{4} a_{1}
$$

$a_{3} a_{2} a_{1}-، a_{3} a_{2}-a_{4} a_{1}>0$ ‘The conditions for stability are then $a_{3}>0$
$a_{0} a_{3}^{2}-a_{4} a_{1}^{2}>0$
$a_{3}\left(a_{2} a_{1} a_{0}-a_{3} a_{0}^{2}\right)-a_{1}^{2} a_{0} a_{4}>0$
Is the system with the following characteristic equation stable?
$S^{4}+3 S^{3}+6 S^{2}+9 S+12=0$

Substituting the appropriate values for the coefficient in the general condition of above Example of Example (3 )we have

$$
\begin{aligned}
& a_{0}=12 \cdot a_{1}=9 \quad a_{2}=6 \quad \checkmark a_{3}=3 \quad a_{4}=1 \\
& a_{3}>0 \Rightarrow 3>0 \\
& a_{3} a_{2}-a_{4} a_{1}>0 \Rightarrow 3.6-1.9>0 \Rightarrow 18-9=9>0 \\
& a_{3} a_{2} a_{1}-a_{0} a_{3}^{2}-a_{4} a_{1}^{2}>0 \Rightarrow(3.6 . a)-(3)^{2}-(1)(a)^{2}>0 \Rightarrow 162-108-81>0 \Rightarrow \\
& -27<0 \\
& 3\left((6.9,12)-3(12)^{2}\right)-(9)^{2}(12)(1)>0 \\
& =3(648)-432)-972>0=648-972=-324<0
\end{aligned}
$$

Since the last two conditions are not satisfied the system is unstable

### 4.1.4 Continued Fraction stability criterion

This criterion is applied to the characteristic equation of a continuous system by forming a continued fraction form the odd and even portion of the equation in the following manner

Let $Q(\mathrm{~s})=a_{n} \mathrm{~S}^{\mathrm{n}}+a_{n-1} \mathrm{~S}^{\mathrm{n}-1}+\ldots .+a_{1} \mathrm{~S}+a_{0}$
$Q_{1}(\mathrm{~s})=a_{n} \mathrm{~S}^{\mathrm{n}}+a_{n-2} \mathrm{~S}^{\mathrm{n}-2}+\ldots \ldots$.
$Q_{2}(\mathrm{~s})=a_{n-1} \mathrm{~S}^{\mathrm{n}-1}+a_{n-3} \mathrm{~S}^{\mathrm{n}-3}+\ldots \ldots .$.
Form the Fraction $\frac{Q_{1}(S)}{Q_{2}(S)}$ and then divide the denominator into the numerator and Invert the remainder to form a continued Fraction as follow:

$$
\begin{array}{r}
\frac{Q_{1}(S)}{Q_{2}(S)}=\frac{\frac{a_{n} S}{a_{n-1}}+\left(a_{n-2}-\frac{a_{n} a_{n-3}}{a_{n-1}}\right) S^{n-2}+\left(\frac{\left.a_{n-4-a_{n-5}}^{a_{n-1}}\right) S^{n-4}+\cdots}{}\right.}{Q_{2}} \\
=h_{1} S+\frac{1}{1} \\
h_{2} S+\frac{1}{1} \\
\\
\quad h_{3} S+\frac{1}{1}
\end{array}
$$

$$
\frac{1}{h_{n} S}
$$

$\mathrm{h}_{2}, \ldots \ldots \ldots . \mathrm{h}_{\mathrm{n}}$ are all positive then all roots of $\mathrm{Q}(\mathrm{S})=0$ have negative 'If $\mathrm{h}_{1}$ real parts

## Example4.9

Using the continued Fraction stability criterion the polynomial
$(S)=S^{3}+4 S^{2}+8 S+12$

## Is divided into the two parts

$(S)=S^{3}+8 S$
$(S)=4 S^{2}+12$
The continued Fraction for is
$\frac{Q_{1}(\mathrm{~s})}{Q_{2}(\mathrm{~s})}=\frac{s^{3}+8 \mathrm{~s}}{4 s^{2}+12}=1 / 4 \mathrm{~s}+\frac{5 s}{4 s^{2}+12}=1 / 4 \mathrm{~s}+\frac{1}{\frac{4}{5} s+\frac{1}{\frac{5}{12}} s}$
$\therefore$ All coefficient of s are positive, the polynomial has all roots in the left plane and the system with the characteristic equation

$$
(\mathrm{s})=0
$$

$\therefore$ The system is stable

## Example4.10

Determine bounds the parameter k which a system the following characteristic equation is stable
$S^{3}+14 s^{2}+56 s+K=0$
The continued Fraction for is
$\frac{Q_{1}(S)}{Q_{2}(S)}=\frac{s^{3}+56 s}{14 s^{2}+k}=\frac{1}{14} s+\left(\frac{56-k}{14}\right) S=\frac{1}{14} s+\overline{\left[\frac{14}{56-\frac{k}{14}}\right] S+\frac{1}{\left[\frac{56-k / 14}{k}\right]} s}$

For the system to be stable the following condition must be satisfied $56-\frac{k}{14}>0$

And $\mathrm{k}>0$ that is $0<\mathrm{k}<784$

## Example 4.11

Derive conditions for all the roots of a general third order polynomial to have negative real parts

For $\mathrm{Q}(\mathrm{s})=a_{3} s^{3}+a_{2} s^{2}+a_{1} s+a_{0}$
The continued Fraction for is

$$
\frac{Q_{1}(S)}{Q_{2}(s)}=\frac{a_{3} s^{3}+a_{1} s}{a_{2} s^{2}+a_{0}}=\frac{a_{3}}{a_{2}} s+\frac{\left[a_{1}-\frac{a_{3} a_{0}}{a_{2}}\right]}{a_{2} s^{2}+a_{0}} s=\frac{a_{3}}{a_{2}} s+\left[\frac{a_{2}}{a_{1}-\frac{a_{3 a}}{a_{2}}}\right] s+\frac{1}{\left[\frac{a_{1}-\frac{a_{3} a_{0}}{a_{0}}}{a_{0}}\right]}
$$

The condition for all the roots of $\mathrm{Q}(\mathrm{s})$ to have negative real parts are then
$\xrightarrow[a_{0}-\frac{a_{3} a_{0}}{a_{2}}]{a_{0}}>0 \underbrace{a_{2}}_{a_{1}-\frac{a_{2}}{a_{2}}}>0 \quad \underset{a_{2}}{a_{3}}>0$
Thus if $a_{3}$ is positive the required conditions are $a_{2}, a_{1}, a_{0}>0$ and $a_{3} a_{0}>0-a_{1} a_{2}$

Note that if $a_{3}$ is not positive (s) should be multiplied by -1 before checking the above condition

## 4:1:5 Stability criterions for Discrete Time systems

The stability of discrete systems is determined by the roots of the discrete system characteristic equation

$$
a_{n} z^{n}+a_{n-1} z^{n-1}+\cdots \ldots \ldots \ldots+a_{1} z+a_{0=} 0
$$

However in this case the stability region is defined by the unit circle $|Z|=1$ in the Z- plane .A necessary and sufficient condition for system stability is that all roots of the characteristic equation have a magnitude less than one. That is be within the unit circle this ensures that the kronecker delta response decays with time .A stability criterion for discrete system similar to Routh criterion is called the jury test for this test the coefficients of the characteristic equation are first arranged in the Jury array row.


Where
$2 \mathrm{n}-3 \quad S_{0} \quad S_{1} \quad S_{2}$

$$
\left.\begin{aligned}
& b_{k}=\left|\begin{array}{cc}
a_{0} & a_{n-k} \\
a_{n} & a_{k}
\end{array}\right|
\end{aligned} \quad c_{k}=\left|\begin{array}{cc}
b_{0} & b_{n-1-k} \\
b_{n-1} & b_{k}
\end{array}\right|-1 . \begin{array}{ll}
S_{0}=\left|\begin{array}{ll}
r_{0} & r_{3} \\
r_{3} & r_{0}
\end{array}\right| \quad, S_{1}=\left|\begin{array}{cc}
r_{0} & r_{2} \\
r_{3} & r_{1}
\end{array}\right| \quad, \left.S_{2=\mid}^{r_{0}} \begin{array}{l}
r_{1} \\
r_{3}
\end{array} \right\rvert\, & r_{2}
\end{array} \right\rvert\,
$$

The first two rows are written using the characteristic equation coefficients and the next two rows are computed using the determinant relationships shown above .The process is continued with each succeeding pair of rows having one less column than the previous pair unit row $2 \mathrm{n}-3$ is computed which only has three entries the array is then terminated .Jury test: Necessary and sufficient conditions for the $\operatorname{root} Q(Z)=0$, To have magnitudes less than one are
$Q(1)>0$

$\left|a_{0}\right|<a_{n}$
$\left|b_{0}\right|>\left|b_{n-1}\right|$
$\left|c_{0}\right|<\left|c_{n-2}\right|$
$\left|r_{0}\right|>\left|r_{3}\right|$
$\left|s_{0}\right|>\left|s_{2}\right|$
Note that if the $\mathrm{Q}(1)$ or $\mathrm{Q}(-1)$ condition above are not satisfied the system is unstable and it is not necessary to construct the array

### 4.2 Stabilization by bounded delayed control

Time delay often occurs in engineering system. Since the existence of time delay usually instability of the system the study on the Time delay systems has received considerable attentions and many stability criteria for time delay systems can be found in the literature. Stability criteria for time delay
systems. Tend to fall into one of the two categories: delay independent and delay dependent, As the name implies delay independent criteria provide conditions which guarantee stability for any length of the time delay on the other hand delay dependent criteria exploit a priori knowledge of upper bounds on the a mount of time delay. These criteria since more information about the Time delay is assumed to be known

Consider the following linear time delay stem
$. \dot{X}(\mathrm{t})=\mathrm{Bx}(\mathrm{t})+B_{d} x(t-T(t))$
Where $\mathrm{T}(\mathrm{t})$ is a unknown time varying parameter which satisfies
$|\mathrm{T}(\mathrm{t})| \leq \mathrm{d} \quad \forall t \geq 0<\mathrm{o} \leq \mathrm{T}(\mathrm{t}) \leq \mathrm{h}$
$\mathrm{x} \epsilon R^{n}$ is the state B and $B_{d} \epsilon R^{n \times n}$ are constant matrices
In is used to denote n-dimensional identity matrix. Given a matrix M , the transposition an the conjugate transposition are denoted by $\grave{M}$ and $\mathrm{M}^{*}$ ${ }^{m}$ and ${ }^{\times}$respectively .A matrix M is called positive definite if $m$ belongs to $\mathrm{H}^{\mathrm{n}}$ $\dot{X} M X>0$ for all $x \in C^{n} \quad x \neq 0$, The notation $M>0$ is used to denote positive definiteness the positive semi definiteness negative definiteness and negative semi definiteness have similar definitions except that the $>$ is replaced by $" \geq ", "<"$ and $\leq$ respective. We use $L_{2}$ to denote the space of square assumable Function defined to time interval $[0, \infty)$. Given a signal f in $\mathrm{L}_{2}$ space. We use $\|f\| \mathrm{L}_{2}$ denote the $\mathrm{L}_{2}$ norm off to denote the $\mathrm{L}_{2}$ induced norm of $G D_{e}$ note the time delay operator and let
$\Delta \mathrm{Be}\left(\mathrm{D}_{\mathrm{T}}-\mathrm{I}\right)^{\circ} \frac{1}{s}$ that is
$\mathrm{D}_{\mathrm{T}}(\mathrm{v}):=\mathrm{v}(\mathrm{t}-\mathrm{T}(\mathrm{t}))$, and
$\Delta(\mathrm{v})=\int_{t}^{t-T(\mathrm{t})} \mathrm{v}(\mathrm{A}) \mathrm{d} \theta$

## Lemma (A)

Operator $\Delta$ is bounded $L_{2}$ space the $L_{2}$ induced norm of $\Delta_{T}$ is equal to $h$ and $\Delta$

Satisfies integral quadratic defined by
$\sigma(\mathrm{v}, \mathrm{w})=\int_{0}^{\infty} h^{2} \mathrm{v}(\mathrm{t}) \mathrm{Xv}(\mathrm{t})-\mathrm{w}(\mathrm{t}) \mathrm{X} \mathrm{w}(\mathrm{t}) \mathrm{dt}$
Where $\mathrm{X}=\bar{X}$ is any positive definite matrix
Proof
Let $X=\grave{X}>o$ and $w=\Delta$ we have
$\mathrm{W}(\mathrm{t})=-\int_{t-T(\mathrm{t})}^{t} \mathrm{v}(\theta) \mathrm{d} \theta \quad$ and
$\grave{\mathrm{w}}(\mathrm{t}) \mathrm{We}(\mathrm{t})=-\left(\int_{t-T(\mathrm{t})}^{T} u(\theta) X \mathrm{v}(\eta) \mathrm{d} \theta \mathrm{d} \eta\right)^{\prime} \mathrm{X}-\left(\int_{t-T(t)}^{t} \mathrm{v}(\eta) \mathrm{d} \eta\right)$
$=\int_{t-T(t)}^{t} \int_{t-T(t)}^{t} v(\theta) X v(\eta) d \theta d \eta$
Using the Cauchy Schwartz in equality formula we get.
$\grave{\mathrm{w}} \times \mathrm{W} \leq \int_{t-h}^{T} \sqrt{\left(h \times v(\eta)^{\prime}\right)} \times v(\eta) \sqrt{\int_{\mathrm{t}-\mathrm{T}(\mathrm{t})}^{t} \mathrm{v}}(\theta) \times \mathrm{v}(\eta) \mathrm{d} \theta \mathrm{d} \eta$
$\leq \mathrm{h}\left(\sqrt{\int_{\mathrm{t}-\mathrm{h}}^{t} \mathrm{v}}(\theta) \times \mathrm{v}(\eta) \mathrm{d} \theta \mathrm{d} \eta \sqrt{\int_{\mathrm{t}-\mathrm{h}}^{t} \mathrm{v}}(\eta) \mathrm{X} \mathrm{v}(\eta) \mathrm{d} \eta \leq \mathrm{h} \int_{\mathrm{t}-\mathrm{h}}^{t} \mathrm{v}(\theta) \times \mathrm{v}(\theta) \mathrm{d} \theta\right.$
This is turn implies

$$
\begin{aligned}
& \int_{0}^{\infty} w(t) \grave{\times} w(\mathrm{t}) d t \leq \int_{0}^{\infty} \mathrm{h} v(\theta) \times(\theta) \mathrm{d} \theta=\mathrm{h} \int_{0}^{\infty}\left(\int_{-h}^{0} \mathrm{v}(\mathrm{t}+\mathrm{s}) \times \mathrm{v}(\mathrm{t}+\mathrm{s}) \mathrm{ds}\right) \mathrm{dt} \\
& \leq \mathrm{h} \int_{-h}^{0}\left(\int_{0}^{\infty} \mathrm{v}(\mathrm{t}) \times \mathrm{v}(\mathrm{t}) \mathrm{dt}\right) \mathrm{d} \mathrm{~s}=\int_{0}^{\infty} h^{2} \grave{\mathrm{v}}(\mathrm{t}) \times \mathrm{v}(\mathrm{t}) \mathrm{dt}
\end{aligned}
$$

$\therefore \Delta$ Bounded by h \#

### 4.3 Stability criteria based on integral quadratic constraints

## Lemma (B)

Let $\mathrm{m} \in R^{n \times n}$ be a constant matrix then the time delay system
$\dot{X}(t)=\mathrm{Bx}(t)+\mathrm{B}_{\mathrm{d}} \mathrm{x}(\mathrm{t}-\mathrm{T}(\mathrm{t})) \quad$ can be equivalent formulated as.
$\dot{X}(t)=\left(\mathrm{B}+\mathrm{mB}_{\mathrm{d}}\right) \mathrm{xt}(\mathrm{x})+\left(\mathrm{I}_{\mathrm{n}}-\mathrm{m}\right) \mathrm{B}_{\mathrm{d}} w_{1}(\mathrm{t})+\mathrm{mB}_{\mathrm{d}} \mathrm{B} w_{2}(\mathrm{t})+\mathrm{mB}^{2} w_{3}(\mathrm{t}) \ldots$ (4.2)
Where $w_{1}(\mathrm{t})=\mathrm{x}(\mathrm{t}-\mathrm{T}(\mathrm{t}))$
$w_{3}(\mathrm{t})=\int_{t}^{t-T(t)} w_{1}(\theta) d \theta \quad w_{2}(\mathrm{t})=\int_{t}^{t-T(t)} x(\theta) d \theta$
Using one can put the linear time delay system standard linear fractional transformation setup for robustness analysis as shown in figure

The linear time invariance (LTI) system $G$ has a state space representation
$\dot{X}(\mathrm{t})=\overline{\mathrm{A}}_{1} \mathrm{x}(\mathrm{t})+\overline{\mathrm{B}}_{1} \breve{W}_{1}(\mathrm{t})+\overline{\mathrm{B}}_{2} \breve{W}_{2}(\mathrm{t})$
$\check{v}_{1}(\mathrm{t})=\mathrm{x}(\mathrm{t})$
$\check{v}_{2}(\mathrm{t})=\left[\begin{array}{c}\mathrm{x}(\mathrm{t}) \\ \check{w} 1(\mathrm{t})\end{array}\right]$
$\breve{w}_{2}=\mathrm{D}_{\mathrm{T}}\left(\breve{v}_{2}\right)$ and matrices $\quad$ Where $\breve{w}_{1}=\mathrm{D}_{\mathrm{T}}\left(\check{v}_{1}\right)$
$\overline{\mathrm{B}}_{2}=\left[\mathrm{mB}_{\mathrm{d}} \mathrm{B} \quad \mathrm{mB}_{\mathrm{d}}^{2}\right] 厄 \overline{\mathrm{~B}}_{1}=\left(\mathrm{I}_{\mathrm{n}}-\mathrm{m}\right) \mathrm{B}_{\mathrm{d}} \quad \measuredangle \bar{A}=\mathrm{B}+\mathrm{mB}_{\mathrm{d}}$
Since system(4.1),(B) and (4.3) are equivalent stability of any one system implies stability of the other Two.


Figure (4.4)

Stability criteria derived in the previous topic are used on simple norm bounded type of integral quadratic constraint for $\mathrm{D}_{\mathrm{T}}$ and $\Delta$ which might be very conservative .Less conservative criteria can be derived provided (ICS) which better characterize $\mathrm{D}_{\mathrm{T}}$ and $\Delta$ are a viable in this topic stronger (ICS) for $\mathrm{D}_{\mathrm{T}}$ and $\Delta$ are derived .

Lemma (C) ( swapping lemma for operator $\mathrm{D}_{\mathrm{T}}$ )
Let H be a stable linear invariant system with state and let T denote operator of multiplying i- e
$\mathrm{T}(\mathrm{v}(\mathrm{t})):=\dot{\mathrm{T}}(\mathrm{t}) \mathrm{v}(\mathrm{t})$ then
$\mathrm{D}_{\mathrm{T}} \mathrm{oH}(\mathrm{s})=\mathrm{H}(\mathrm{s})^{\circ} \mathrm{D}_{\mathrm{T}}-\mathrm{H}(\mathrm{s})^{\circ} \mathrm{T}^{\circ} \mathrm{sH}(\mathrm{s})$

## Proof

Let V be any $\mathrm{L}_{2}$ Function and defined y and $\underline{Z}$ to be

$$
\begin{aligned}
& \mathrm{y}(0)=0 \quad \dot{\mathrm{y}}(\mathrm{t})=A_{n} \mathrm{y}(\mathrm{t})+\mathrm{v}(\mathrm{t}) \\
& \mathrm{z}(0)=0 \quad \dot{\mathrm{z}}=A_{n} \mathrm{z}(\mathrm{t})+\mathrm{v}(\mathrm{t}-\mathrm{T}(\mathrm{t}))
\end{aligned}
$$

Let $r(t)=y(t-T(t))-z(t)$ and we have
$\dot{r}(\mathrm{t})=\dot{\mathrm{y}}(\mathrm{t}-\mathrm{T}(\mathrm{t}))(1-\mathrm{T}(\mathrm{t}))-\dot{z}(\mathrm{t})$

$$
\begin{aligned}
& \dot{\mathrm{y}}(\mathrm{t}-\mathrm{T}(\mathrm{t})) \theta=A_{n}(\mathrm{y}(\mathrm{t}-\mathrm{T}(\mathrm{t}))-\mathrm{z}(\mathrm{t}))-\dot{\mathrm{T}}(\mathrm{t}) \\
& =A_{n} \mathrm{r}(\mathrm{t})-\dot{\mathrm{T}}(\mathrm{t}) \mathrm{OD} \mathrm{D}_{\mathrm{T}}(\dot{\mathrm{y}}(\mathrm{t}))
\end{aligned}
$$

Which implies $\mathrm{D}_{\mathrm{T}}(\mathrm{HV})=\mathrm{HoD}_{\mathrm{T}}(\mathrm{V})-\mathrm{HoD}_{\mathrm{T}}\left(\frac{d}{d t}(\mathrm{HV})\right)$
This conclude the proof
Lemma ( $\mathbf{D}$ )(swapping lemma for operator $\Delta$ )
Let H and T be the operators as defined in lemma 2 then
$\Delta \mathrm{oH}(\mathrm{s})=\mathrm{H}(\mathrm{s}) \mathrm{o} \Delta-\mathrm{H}(\mathrm{s}) \mathrm{oToD}_{\mathrm{T}} \mathrm{OH}(\mathrm{s})$

## Proof

The proof is similar to the of lemma
$z$ and $x$ to be 'Let $V$ be any $L_{2}$ Function and defined $y$

$$
\begin{aligned}
& \mathrm{y}(0)=0 \dot{\mathrm{y}}(\mathrm{t})=A_{n} \mathrm{y}(\mathrm{t})+\mathrm{v}(\mathrm{t}) \\
& \mathrm{z}(0)=0 \dot{\mathrm{z}}=A_{n} \mathrm{z}(\mathrm{t})+\int_{T}^{\mathrm{t}-\mathrm{T}(\mathrm{t})} \mathrm{v}(\theta) \mathrm{d} \theta \\
& \mathrm{r}(\mathrm{t})=\int_{T}^{\mathrm{t}-\mathrm{T}(\mathrm{t})} \mathrm{y}(\theta) \mathrm{d} \theta-\mathrm{z}(\mathrm{t})
\end{aligned}
$$

One can easily verify that

$$
\begin{aligned}
& \dot{\mathrm{r}}(\mathrm{t})=\mathrm{y}(\mathrm{t}-\mathrm{T}(\mathrm{t}))(1-\dot{\mathrm{T}}(\mathrm{t}))-\mathrm{y}(\mathrm{t})-\dot{\mathrm{z}}(\mathrm{t}) \\
& \mathrm{y}(\mathrm{t}-\mathrm{T})(\mathrm{t}) \theta=\int_{T}^{\mathrm{t}-\mathrm{T}(\mathrm{t})} \dot{\mathrm{y}}(\theta) \mathrm{d} \theta-\dot{\mathrm{z}}(\mathrm{t})-\dot{\mathrm{T}}(\mathrm{t}) \\
& =A_{n} \int_{T}^{\mathrm{t}-\mathrm{T}(\mathrm{t})} \mathrm{y}(\theta) \mathrm{d} \theta-A_{n} \mathrm{z}(\mathrm{t})-\dot{\mathrm{T}}(\mathrm{t}) 0 \mathrm{y}(\mathrm{t}-\mathrm{T})(\mathrm{t}) \\
& =A_{n} \mathrm{r}(\mathrm{t})-\dot{\mathrm{T}}(\mathrm{t}) 0 \mathrm{D}_{\mathrm{T}}(\mathrm{y}(\mathrm{t}))
\end{aligned}
$$

Which implies $\Delta(\mathrm{HV})=\operatorname{Ho} \Delta(\mathrm{V})-\operatorname{HoToD}_{\mathrm{r}}(\mathrm{HV})$ this concludes the proof

### 4.4 Bounded by a neighborhood

## Corollary(i)

Let $\mathrm{y} \subseteq R^{m}$ and pick any $\mathrm{T} \geq 0$ and any integer $\mathrm{q} \geq 1$
Then $R_{u}^{T}(0)+e^{T A} R_{u}^{T}(0)+\ldots \ldots \ldots .+e^{(\mathrm{q}-1) \mathrm{TA}} R_{u}^{T}(0)=R_{u}^{q T}(0)$
$\operatorname{Lemma}(\mathbf{E})$ if C is an open convex sub set of $R^{n}$ and L is a sub space of $R^{n}$ contained in C then $\mathrm{c}+\mathrm{L}=\mathrm{C}$

## Proof

so we only need prove the other inclusion pick any ${ }^{\text {Cleary }} \mathrm{c}=\mathrm{c}+0 \leq \mathrm{C}+\mathrm{L}$ $x \in C$ and $y \in L$ then for all $\in \neq 0$
$\left.\mathrm{x}+\mathrm{y}=\left(\frac{1}{1+\epsilon}\right)[1+\epsilon) \mathrm{x}\right]+\left(\frac{\epsilon}{1+\epsilon}\right)\left[\left(\frac{1+\epsilon}{\epsilon}\right) y\right]$
Since $C$ is open $(1+\in) x \in C$ for some sufficiently small $\in>0$
Since L is a sub space $\left(\frac{1+\epsilon}{\epsilon}\right) y \in \mathrm{~L} \subseteq \mathrm{c}$
That $x+y \in C$ by convexity

## Lemma (H)

$\mathrm{T} \geq 0$ then Let $\mathrm{u} \subseteq R^{m}$ and pick any Two S
$R_{u}^{T}(0)+e^{T A} R_{u}^{s}(0)=R_{u}^{s+T}(0)$

## Proof

Pick.
$x_{1}=\int_{0}^{T} \mathrm{e}^{(\mathrm{T}-\mathrm{T}) \mathrm{A}} \mathrm{B} w_{1}(\mathrm{~T}) \mathrm{dt}=\int_{S}^{S+T} \mathrm{e}^{(\mathrm{s}+\mathrm{T}-\mathrm{T}) \mathrm{A}} \mathrm{B} w_{1}(\mathrm{~T}-\mathrm{s}) \mathrm{dt} \quad$ and
$x_{2}=\int_{0}^{T} \mathrm{e}^{(\mathrm{s}-\mathrm{T}) \mathrm{A}} \mathrm{B} w_{2}(\mathrm{~T}) \mathrm{dt}$
Note that ${ }^{\prime}$ With inputs $w_{i} \mathrm{U}$ - valued
$e^{T A} x_{2}=\int_{0}^{S} \mathrm{e}^{(\mathrm{s}+\mathrm{T}-\mathrm{T}) \mathrm{A}} \mathrm{B} w_{2}(\mathrm{~T}) \mathrm{dt}$
Thus $x_{1}+e^{T A} x_{2}=\int_{0}^{s+T} \mathrm{e}^{(s+\mathrm{T}-\mathrm{T}) \mathrm{A}} \mathrm{B} w(\mathrm{~T}) \mathrm{dt}$
Where $\mathrm{w}(\mathrm{s})=\left\{\begin{array}{lr}w_{2}(\mathrm{~T}) & 0 \leq \mathrm{T} \leq \mathrm{S} \\ w_{1}(\mathrm{~T}-\mathrm{s}) \mathrm{s} \leq \mathrm{T} \leq \mathrm{S}+\mathrm{T}\end{array}\right.$
$w(s) \in U$ for all $T \in[0, s+T]$
Thus $R_{u}^{T}(0)+e^{T A} R_{u}^{S}(0) \leq R_{u}^{S+T}(0)$
The converse inclusion follows by reversing these steps
There fore $R_{u}^{T}(0)+e^{T A} R_{u}^{S}(0)=R_{u}^{S+T}(0)$

## Lemma (I)

Assume that $(\mathrm{A}, \mathrm{B})$ is controllable and $\mathrm{u} \leq R^{m}$ is an eigh borhood of 0 then $\mathrm{J}_{\mathrm{k}}^{\mathrm{R}} \subseteq R_{u}(0)$ for all k

## Proof

First replacing if necessary $U$ by a convex subset we may assume without loss of generality that $u$ is a convex neighborhood of 0we prove that statement by induction on k the case $\mathrm{k}=0$ being trivial so assume that
$\widetilde{\mathrm{V}}=\operatorname{any} \widetilde{\mathrm{V}}_{1}+\mathrm{T}_{2} \widetilde{\mathrm{~V}}_{1} \in R_{u}(0) \cdot \mathrm{J}_{\mathrm{k}-1}^{\mathrm{R}} \subseteq R_{u}(0)$ and take any $\widetilde{\mathrm{V}} \in \mathrm{J}_{\mathrm{k}}$ ، $\lambda$
First pick any $\mathrm{T}>\mathrm{o}$ so that $e^{\lambda T j}=e^{d T j}$ for all $\mathrm{j}=$ $0,1, \ldots \ldots \ldots \ldots \ldots \ldots \ldots \ldots . .$.

If $\beta=0$ one may take any $\mathrm{T}>0$
Otherwise, we may use for instance
$\mathrm{T}=\frac{2 \pi}{|\beta|}$ next choose any $\delta>0$ with the property that $\mathrm{V}_{1}:=\delta \mathrm{V}_{1}^{2} \in R_{u}^{T}(0)$
There is such a $\delta$ because $R_{u}(0)$ contains 0 in its interior
Since $\mathrm{V} \in \operatorname{Ker}(\lambda I-A)^{\mathrm{k}}$ where $\mathrm{V}=\delta \widetilde{\mathrm{V}}$
$\mathrm{e}^{(\mathrm{A}-\lambda \mathrm{T}) \mathrm{T}} \mathrm{V}=\left(1+\mathrm{T}(\mathrm{A}-\lambda \mathrm{I})+\frac{t^{2}}{2}(\mathrm{~A}--\lambda \mathrm{I})^{2}+\ldots \ldots \ldots.\right) \mathrm{V}=\mathrm{V}+\mathrm{W} \quad \forall \mathrm{T}$
Where $\mathrm{w} \in \mathrm{J}_{\mathrm{k}-1}$
Thus
$e^{\alpha T} \mathrm{~V}=e^{\lambda T} \mathrm{~V}=e^{T A} \mathrm{~V}-e^{\lambda T} \mathrm{~W}=e^{T A} \mathrm{~V}-e^{\alpha T} \mathrm{~V} \quad \forall \mathrm{~T}=j \mathrm{~T}$
$j=0,1, \ldots \ldots \ldots \ldots \ldots$
Decomposing into real and imaginary parts
$\mathrm{w}=w_{1}+\mathrm{i} w_{2}$ and taking real parts in above equation
$\mathrm{j}=0,1, \ldots \ldots \ldots . e^{\alpha T} \mathrm{v}_{1}=e^{T A} \mathrm{v}_{1}-e^{\alpha T} \mathrm{w}_{1} \forall \mathrm{t}=\mathrm{j} \mathrm{T}$
Now pick any integer $\mathrm{q} \geq 1$ then
$\left(\sum_{\mathrm{j}=0}^{q-1} e^{\alpha \mathrm{T}}\right) \mathrm{v}_{1}=\sum_{\mathrm{j}=0}^{q-1} \mathrm{e}^{\mathrm{j} T \mathrm{~A}} \mathrm{v}_{1}+\grave{W}$
Where $\grave{w}=-\sum e^{\alpha j \mathrm{~T}}$ belong to the sub space $\mathrm{j}_{\mathrm{k}-1}^{\mathrm{R}}$
Applying First corollary (i) and lemma (A)
We conclude
$\mathrm{P} V_{1} \in R_{u}^{q T}(0)+\mathrm{J}_{\mathrm{k}-1}^{\mathrm{R}} \subseteq \mathrm{R}_{\mathrm{u}}(0)$
Where $\mathrm{P}=\sum_{j=0}^{q} e^{\alpha j T} \geq \sum_{j=0}^{q-1} 1=\mathrm{q} \geq \frac{1}{\delta}$

Here is precisely where we used that $\alpha \geq 0$
Therefore $\delta \breve{V}_{1}=P V_{1} \in R_{u}(0)$
On the other hand $\delta \mathrm{P} \geq 1$ means that
$\breve{V}_{1}=\frac{1}{\phi p} \delta \mathrm{p} \widetilde{V}_{1}+\left(1-\frac{1}{\delta \phi}\right) 0$
Is a convex combination
Since $\delta P \breve{V}_{1}$ and 0 both belong to $\mathrm{R}_{\mathrm{u}}(0)$ we conclude by convexity of the latter that indeed $\breve{V}_{1} \in \mathrm{R}_{\mathrm{u}}(0)$

## Corollary (ii)

Assume that (A, B) is controllable and $U \subseteq R^{m}$ is a convex and bounded neighborhood of 0

There exists a set B such that
$R_{u}(0)=B+L \quad$ and $B$ is bounded convex and open relative to $m$

## Proof

We claim that $\mathrm{R}_{\mathrm{u}}(0)=\left(\mathrm{R}_{\mathrm{u}}(0) \cap m\right)+L$
On inclusion is clear form

$$
\left(\mathrm{R}_{\mathrm{u}}(0) \cap m\right)+L \subseteq \mathrm{R}_{\mathrm{u}}(0)+\mathrm{L}=\mathrm{R}_{\mathrm{u}}(0)
$$

Applying lemma (B)
Conversely any $\mathrm{V} \epsilon \mathrm{R}_{\mathrm{u}}(0)$ can by decomposed as $\mathrm{V}=\mathrm{x}+\mathrm{y} \epsilon \mathrm{m}+\mathrm{L}$
We need to show that $\mathrm{x} \in \mathrm{R}_{\mathrm{u}}(0)$ But
$\mathrm{x}=\mathrm{v}-\mathrm{y} \epsilon \mathrm{R}_{\mathrm{u}}(0)+\mathrm{L}=\mathrm{R}_{\mathrm{u}}(0)$
Applying the same lemma (B) yet again this establishes the claim
We let $\mathrm{B}:=\mathrm{R}_{\mathrm{u}}(0) \cap m$
This set is convex and open in $m$ because $R_{u}(0)$ is open and convex
We only need to proof that it is bounded
Let $\mathrm{P}: R^{n \rightarrow} R^{n}$ be the projection on m a long L
$\mathrm{y} \in L$ observe that $\mathrm{PA}=\mathrm{AP}$ because each of L ،That is $\mathrm{P}(\mathrm{x}+\mathrm{y})=\mathrm{x} \quad$ if $\quad \mathrm{x} \epsilon m$
Ay $\epsilon L$ imply ' $\mathrm{Ax} \epsilon \mathrm{m}$ cand m are A invariant (so $\mathrm{v}=\mathrm{x}+\mathrm{y}$
$\mathrm{PAV}=\mathrm{Ax}=\mathrm{APV})$.
Pick any $\mathrm{x} \in R_{u}(0) \cap m$
Since $\mathrm{x} \in R_{u}$ (0) there are some T and some w so that
$\mathrm{x}=\int_{0}^{T} e^{(T-\tilde{\mathrm{I}}) A} B w(\tilde{\mathrm{I}}) d \tilde{\mathrm{I}}$
$\mathrm{x}=\mathrm{px}$ ،On the other hand since $\mathrm{x} \epsilon m$
Thus $\quad \mathrm{x}=\mathrm{px}=\int_{0}^{T} e^{(T-\tilde{\mathrm{I}}) A} B w(\tilde{\mathrm{I}}) d \tilde{\mathrm{I}}=\int_{0}^{T} e^{(T-\tilde{\mathrm{I}}) A} \quad \mathrm{x}(\tilde{\mathrm{I}}) \mathrm{d} \tilde{\mathrm{I}}$
Where $\mathrm{x}(\tilde{\mathrm{I}})=\mathrm{PW}(\tilde{\mathrm{I}}) \in m \cap p B(u)$ for all $\tilde{\mathrm{I}}$ since the restriction of A to m has all its Eigen values negative real part
There are positive constants $\mathrm{C}, \mathrm{m}>0$ such that
$\left\|e^{T A}\right\| \leq c e^{-m t}\|x\|$ For all $\mathrm{t} \geq 0$
That if x is also in $\mathrm{PB}(\mathrm{u}),\left\|e^{T A}\right\| \leq c^{1} e^{-m t} \quad$ for all $\quad \mathrm{t} \geq 0$
So for as above we conclude

$$
\|x\| \leq c^{1} \int_{0}^{T} e^{-m(T-\tilde{1})} d \tilde{\mathrm{I}} \leq \frac{c^{1}}{m}\left(1-e^{-m t}\right) \leq \frac{c^{/ /}}{m}
$$

### 4.5 Lower Bounded control lyapunov Functions

We will consider systems of the form
$X \dot{(t)}=\mathrm{F}(\mathrm{x}(\mathrm{t})+\mathrm{u}(\mathrm{t}))$
Where the state $\mathrm{x}(\mathrm{t})$ evolves in $R^{n}$ and the controls $\mathrm{u}(\mathrm{t})$ take values in a sub set $u R^{m}$ containing the origin. For simplicity we assume that $v=R^{m}$
The map $\mathrm{f}: R^{n} R^{m} \rightarrow R^{n}$ is locally lipschitzat $(x, v)$ and $\mathrm{f}(0,0)=0$

To stabilize this system to $\mathrm{x}=0$ we will use energy Function v which can be made to decrease a long system trajectories. A Function $\mathrm{v}: R^{n} \rightarrow \mathrm{R} \geq 0$

Is positive definite if $\mathrm{v}(0)=0 \mathrm{v}(\vartheta)>0$ for $\neq 0$
And properif $\mathrm{v}(\vartheta) \rightarrow \infty$ as $\|\vartheta\| \rightarrow \infty$ for $\epsilon \mathrm{R}^{\mathrm{n}}$
We denote by $\|\vartheta\|$ theEucliden norm
For a locally Lipchitz continuous Function
$\mathrm{V}: \mathrm{R}^{\mathrm{n}} \rightarrow \mathrm{R}$ and $\mathrm{P} \in \mathrm{R}^{\mathrm{n}}$ we define the Dini derivative of v in the direction of P At to be
$\overline{\mathrm{D}+\mathrm{v}}(\vartheta, \mathrm{p})=\lim _{t \rightarrow 0^{t}} \sup \frac{[v(\vartheta+t p)-v(\vartheta)]}{t}$
In what follows we will be considering closed loop systems of the form $\left.\mathrm{u}_{*}(\vartheta)\right)$ • $\mathrm{f}(\vartheta$

Where the feedback controller $u_{*}$ is measurable and locally bounded but possibly discontinuous hence the classical notion of solution need not apply to deal with this situation we will use the generalized solution due to flipper Set valued Function
$\vartheta) \backslash \mathrm{N}\} f u_{*}\left(\mathrm{~B}\left(£\left\{\mathrm{~F} u_{*}(\vartheta)=\cap \cap \overline{C O}\right.\right.\right.$

$$
\vartheta>0 \quad \mu(\mathrm{~N})=0
$$

$\delta$ ) is a ball of radius $\delta$ centered at $£$ 'Whereby ( $£$ $\overline{C O}$ Denoted the closure of the convex hull and $m$ is used Lévesque measure on $\mathrm{R}^{\mathrm{n}}$,AFilippov solution on an interval I C R is absolution continuous on any interval 0)Is Function $\mathrm{x}: \mathrm{I} \rightarrow \mathrm{R}^{\mathrm{n}}$ such that $\mathrm{x}($ , Al most everywhere on I solution $\mathrm{x}(\mathrm{t})$ to (c) ) $\left[t_{1}, t_{2}\right] \mathrm{I}$, And $\dot{x}(\mathrm{t}) \in \mathrm{F} u_{*}(\mathrm{x}(\mathrm{t})$ are thus state trajectories of the system under. The Feedback controller $u=$ $u_{*}(\mathrm{~s})$ which are differentiable al mostevery where with respect to $\tau$. Since $\mathrm{f}(£$ $\mathfrak{£})$ ) is measurable and locally bounded .The set valued Function $u_{*}(\mathfrak{£}) u_{*}$ ' is upper semi continuous compact and convex value and locals bounded in
particular the differentiable inclusion satisfies ,The basic condition and thus has oFilippov solution for each intrastate .the solution $\mathrm{x}(\mathrm{t})=0$ of a different inclusion $\dot{x}(\mathrm{t})=(\mathrm{x}))$ is called stable if for each $\epsilon>0$ there exists $\delta>0$ with the following property, For each $x_{0}$ such that $\left\|x_{0}\right\|<\delta$ each solution $\mathrm{x}(\mathrm{t})$ to $\mathrm{x}(\mathrm{t})$ to $\dot{x}(\mathrm{t}) \epsilon \quad\left(x(t)\right.$ with intial data $\mathrm{x}(0)=x_{0}$

Exists for $0 \leq t \leq \infty$ and satisfies the inequality

$$
\|x\|<\epsilon(0 \leq t<\infty)
$$

*asymptotically stable if $\mathrm{x}(\mathrm{t})=0$ is stable and in addition, $\mathrm{x}(\mathrm{t}) \rightarrow 0$ as $\mathrm{t} \rightarrow 0$ The possibly discontinuous Feedback controller
$\mathrm{X} \rightarrow u_{*}(x)$ stabilize systm (A) iff $\mathrm{x}(\mathrm{t})=0$
Is a stable solution of the corresponding differential inclusion

### 4.5.1 Lower bounded lyapunov pairs

Lipchitz continuous lyapunov pair ( $\mathrm{v}, \mathrm{w}$ ) consists of a locally Lipchitz
continuous positive definite proper Function $\mathrm{v}: R^{n} \rightarrow R$ and anon negative continuous Function
$\mathrm{W}: R^{n} \rightarrow R$ such that for $£ \neq 0$
V) will $\cdot T h e r e ~ e x i s t s ~ p ~ f f ~(£ ~$
$\mathrm{p}) \leq-\mathrm{w}(£) . \overline{D+V}(£$

## Definition 4.1

A lower bounded lyapunov pair for the system
$\dot{x}(\mathrm{t})=\mathrm{f}(\mathrm{x}(\mathrm{t}), \mathrm{u}(\mathrm{t})$ is alipschitz continuous lyapunov pair $(\mathrm{v}, \mathrm{w})$ such that
V) such ${ }^{*}$ For $£ \neq 0$ there exists $p_{1}, p_{2} \in f(£$
$\left.\left.\mathrm{p}_{2}\right) \ldots \ldots \ldots \ldots \ldots \ldots \ldots \ldots \ldots \ldots \ldots \ldots(\mathrm{c}) \mathrm{p}_{1}\right) \leq-\mathrm{w}(£) \leq \overline{D+V}(£ \sqrt{\bar{D}+V}(£$
**for $\epsilon>0$ there exists $\delta>0$ such that for $0<\|E\|<$ (i) holds with
$\left.\left\|u_{i}\right\|<{ }_{u} u^{\prime}\right)$

## Definition4.2

A lower bounded lyapunov ( $\mathrm{v}, \mathrm{w}$ ) said to be regular if

1. The set valued Function (£) is upper semi continuous, compact and convex valued and locally bounded.
2. There is appositive definite continuous map $\check{w}$ on $R^{n}$ such that at points $£$ where the Function $v$ is not different table ( or more generally where
p ) is not linear) $\mathrm{P} \rightarrow \overline{D+V}(£$
$\mathrm{p}) \leq-\breve{w}(£) \forall \mathrm{p} \epsilon^{\mathrm{V}, \mathrm{w}}(£) \stackrel{\overline{D+V}}{ }(£$

## Definition 4.3

Alocally lipschtz continuous, positive definite and prper Fuction V is called alower bounded control lyapunov Function (LB- CLF), if these Exist apositive definite Function w such that ( $\mathrm{v}, \mathrm{w}$ ) is regular lower bounded lyapunov pair the pair ( $\mathrm{v}, \mathrm{w}$ ) will be call regular lyapunov pair for we note that if $(v, w)$ is aregular pair for $V$ then the differentiable inclusion

## Chapter five

## Reachability and controllability under sampling

## 5. Introduction

When studying the input to state interaction we can take two different points of view in the former we assume that the initial state of the system, the state of the system at $t=0$ is the fixed and we consider the problem of determining the states of system that can be reached applying a certain input in this case we study the so called Reachability property in the latter.We assume that the final state of the system at some time T is fixed and we aim at determining all initial states that can be steered, by means of a certain input signal to the selected final state. In this case we study the so called Controllability property. In the study of reachability and controllability whenever the input signal that drives a certain initial state to a certain final state is not unique we could impose constraints on such an input signal. E.g. we could consider the input signal with minimum time if minimum amplitude, or the input signal which achieves the transfer are equivalent no constrain is imposed all input signals a achieving the considered transfer in minimum time. For linear systems the properties of reachability and controllability are referred to the state $x=0$, hence we say that a state is reachable to mean that it is reachable form $x=0$ and that a state is controllable to mean that it is controllable to $x=0$.Note more over that, Because these properties are used to describe the input to state interaction they trivially depend only upon properties of the matrices A and B

### 5.1 Reachability of discrete time systems

Consider a linear, time invariant, discrete time system
Let $x(0)=0$ and consider an input sequence $\mathrm{u}(0), \mathrm{u}(1), \mathrm{u}(2), \ldots, \mathrm{u}[\mathrm{k}-1]$, the state reached at $\mathrm{t}=\mathrm{k}$ is given

$$
x[k]=\left[\mathrm{B} \mathrm{AB} \ldots . \mathrm{A}^{\mathrm{k}-1} \mathrm{~B}\right]\left[\begin{array}{c}
u[k-1] \\
u[k-2] \\
\vdots \\
u(0)
\end{array}\right]
$$

This implies that the set of states that can be reached at $\mathrm{t}=\mathrm{k}$ is a linear space i-e it is the subspace $R_{k}$ spanned by all linear combinations of the columns of the matrix
$R_{k}=\left[\begin{array}{llll}B A B & \ldots & A^{k-1} B\end{array}\right]$
The set $R_{k}$ is a vector space, denoted as the reachable subspace ink steps if $R_{k}=x$,I-e rank $R_{k}=n$ then all state of the system are reachable in (at most) k steps and the system is said to be reachable in k steps. Ask varies we have a sequence of subspace namely $R_{1}, R_{2}, \ldots, R_{k}$. this sequence of subspaces is such that the following properties hold.

## Proposition (5.1)

The sequence of subspace $(A)$ is such that $R_{1} \subseteq R_{2} \subseteq \ldots \subseteq R_{k} \subseteq \ldots$
Moreover if for some $\bar{k}, R_{\bar{k}}=R_{\bar{k}+1}$ then for all $k \geq \bar{k}, R_{k}=R_{\bar{k}}$ finally

$$
\mathrm{R}_{1} \subseteq \mathrm{R}_{2} \subseteq \ldots \subseteq \mathrm{R}_{\mathrm{n}}=\mathrm{R}_{\mathrm{n}+1}
$$

## Proof

To prove the first claim note that if a state $\bar{x}$ is reached from zero in k steps, using

The input sequence $u(0), u(1), \ldots ., u[k-1]$ then the same state is also reached from zero in $k+1$ steps, using the input sequence $0, u(0), u(1)$, ...., $u(k-1)$,hence for all $k \geq 1$
$\mathrm{R}_{\mathrm{k}} \subseteq \mathrm{R}_{\mathrm{k}+1}$
To prove the second claim it is enough to show, Or equivalently that if then any, Belongs also to .For let be an element of this mean that there is an input sequence which steers the state of the system from $x=0$ to $\bar{x}$ in $\overline{\mathrm{k}}+\mathrm{z}$ step consider now the state reached after $\overline{\mathrm{k}}+1$ steps using the same input sequence which we denote with $\tilde{x}$.By assumption hence there is an input sequence which steers the state of the system from $x(0)=0$ to $\tilde{x}$ in $\tilde{k}$ steps However by definition of $\tilde{x}$ it is possible to steer $\tilde{x}$ to $\bar{x}$ in one step, hence there is an input sequence which steers $x(0)=0$ to $\tilde{x}$ in $\overline{\mathrm{k}}+1$ steps which prove claim to prove the third claim note that if for some $\mathrm{k}<\mathrm{n}, \mathrm{R}_{\mathrm{k}}=\mathrm{R}_{\mathrm{k}+1}$ then the claim follows from equation. Suppose now that for all $k$ the dimension of $R_{k+1}$ is strictly larger that the dimension of $R_{k}$ this implies that the sequences $\operatorname{dim} R_{k}$.Is strictly increasing at each step however this sequence
is bounded (from above) by n and this proving claim?

## Definition5.1

Consider the discrete time system (A) the subspace $\mathfrak{R}=\mathfrak{R}_{\mathrm{n}}$ is reachability subspace of the system, the matrix $R=R_{n}$ is reachability matrix of the system

The system is said to be reachable if $R=x=R^{n}$ remark by definition $R=\operatorname{ImR}$, Hence the discrete time system
$\sigma x=\mathrm{A} x+\mathrm{Bu}, \quad \mathrm{y}=\mathrm{c} x+\mathrm{Du}$
With $x \in \mathrm{x}=\mathfrak{R}, \mathrm{u}(\mathrm{t}) \in \mathfrak{R}^{\mathrm{m}}, \mathrm{y}(\mathrm{t}) \in \mathfrak{R}^{\mathrm{p}}$ and $\mathrm{A}, \mathrm{B}, \mathrm{C}$ and D matrices of appropriate dimensions and with constant entries. The discrete time system is said to be reachable if and only if $. \operatorname{Rank} \mathrm{R}=\mathrm{n}$. Equation is known as Kaman reachability rank condition

## Remark(5.1)

From the above discussion it is obvious that in an. Dimensional linear discrete time system if a state $\bar{x}$ is reachable, Them it is reachable in at most n steps this does not mean that n steps are necessarily required i -e , The state $\bar{x}$ could be reached in less than n steps .In a reachable system. The smallest integer $\mathrm{k}^{*}$ such that rank $\mathrm{k}_{\mathrm{k}^{*}}=\mathrm{n}$.Is called the reachability index of the system Note that for single input reachable systems necessarily $\mathrm{k}^{*}=\mathrm{n}$

## Example 5.1

Consider a discrete time system with $x \in \mathscr{R}$

$$
\mathrm{A}=\left[\begin{array}{ccc}
0 & 0 & \Phi \\
0 & 0 & \Phi \\
0 & 0 & 1
\end{array}\right] \quad \mathrm{B}=\left[\begin{array}{l}
0 \\
1 \\
1
\end{array}\right]
$$

Then $R_{1}=\operatorname{span} B \quad R_{2}=R_{3}=R \operatorname{span}\left[\begin{array}{ll}0 & 1 \\ 1 & 0 \\ 1 & 1\end{array}\right]$
Hence the system is not reachable and its reachable subspace has dimension two

## Example5.2

Consider a discrete time system with $x \in \mathfrak{R}^{3}$

$$
A=\left[\begin{array}{lll}
0 & 1 & 0 \\
0 & 0 & 1 \\
0 & 0 & 0
\end{array}\right] \quad B=\left[\begin{array}{l}
d \\
B \\
y
\end{array}\right]
$$

Then

$$
R_{3}=\operatorname{span}\left[\begin{array}{lll}
\alpha & \beta & \gamma \\
\beta & \gamma & 0 \\
\gamma & 0 & 0
\end{array}\right], R_{2}=\operatorname{span}\left[\begin{array}{ll}
\alpha & \beta \\
\beta & \gamma \\
\gamma & 0
\end{array}\right], R_{1}=\operatorname{span} B
$$

As a result the system is reachable if and only if $\gamma \neq 0$ moreoverif $\gamma=$ 0 and
$\beta \neq 0$ The system is not reachable and the reachable subspace has dimension two. Finally if $\gamma=\beta=0$ and $\alpha \neq 0$

The system is not reachable and the reachable subspace has dimension one the reachability subspace R has the following important property

The proof of which is a simple consequence of the definition of the subspace

## Proposition (5.2)

The reachability sub space contains the subspace span B i-e Span $B \subseteq R$

And it is A invariant i.e. $\mathrm{AR} \subseteq \mathfrak{R}$ we conclude this case noting that algebraically equivalent system, with state x and $\tilde{\mathrm{x}}$ respectively

Let be the coordinates transformation matrix, as a given in equation below consider a continuous time Finite dimensional linear system described by the equations
$\dot{x}=A(t) x+B(t) u \quad, \quad y=C(t) x+D(t) u$
$\mathrm{y}(\mathrm{t}) \in \mid \mathrm{R}^{\mathrm{p}}$ and the change of coordinates $\mathrm{r} \mathrm{u}(\mathrm{t}) \in \mathfrak{R}^{\mathrm{m}} \quad$,With $\mathrm{x} \in \mathrm{X}=\mathfrak{R}^{\mathrm{n}}$
$x(t)=l(t) \hat{x}(\mathrm{t})$
With $l(t)$ invertible for all $t$, the state space representation in the new coordinates is given by
$\dot{\tilde{x}}=\left[\mathrm{l}^{-1}(\mathrm{t})(\mathrm{A}) \mathrm{x}+\mathrm{B}(\mathrm{t})(\mathrm{u})+\mathrm{l}^{-1}(\mathrm{t}) \mathrm{x}\right]_{x}=l(\mathrm{t}) \hat{\mathrm{x}}=\left(\mathrm{l}(\mathrm{t})^{-1}(\mathrm{t})+\mathrm{i}^{-1}(\mathrm{t}) \mathrm{l}(\mathrm{t}) \tilde{\mathrm{x}}+\mathrm{l}^{-}\right.$ ${ }^{1}(\mathrm{t}) \mathrm{B}(\mathrm{t}) \mathrm{u}$

And $y(t)=C L(t) \tilde{x}+D(t) u$
Similarly for a discrete time finite dimensional linear system described by the state 'the equations change of coordinates, With $u(t)$ invertible for all $t$ space representation in the new coordinates is given by and $R_{k}$ and $\widetilde{R}_{k}$ the reachability subspaces and $R$ and $\widetilde{R}$ the reachability matrixes respectively then

Hence .And one of the two systems is reachable if and only if the other is

### 5.2 Controllability of discrete time system

The result established for the reachability the controllability property
In fact for a linear time invariant, discrete time system .A state $x^{*}$ is controllability (to zero) in $k$ steps if there exists an input sequence $u(0), u(1)$ , $\ldots \ldots \ldots \ldots, \mathrm{u}(\mathrm{k}-1)$ that drives the state from. This last equation implies that is controllable if the state, is reachable in K steps, Hence if; It is easy to see
that the set of all such that equation. Hold is a vector space denotes by and called controllability subspace in k steps, A linear discrete time system is controllable in k steps if Im

## Example 5.3

Consider the system in Example the system is controllable in two steps in fact

$$
A^{2}=A \in R_{2}=
$$

## Example 5.4

Consider the system in Example the system is controllable in three steps no matrix the values of $\alpha, \beta$ and $\gamma$,Note in fact that $\mathrm{A}^{3}=0$. The system is controllable in two steps .If $\gamma=0$ and $\alpha \neq 0 \quad$ or $\beta \neq 0$ finally it is controllable in one step if $\gamma=0$ and $\alpha \beta \neq 0$, Similarly to the reachability subspaces ask varies we have a sequence of controllability subspace namely This sequence of subspace is such that following properties hold.

## Proposition (5.3)

The sequence of subspaces is such that, More over if for some the for all

## Proof

The proof of this statement is similar to the one of proposition 7-1 we simply remark that if a state is controllable in $K$ steps using the input sequence $u(0)$ $, \mathrm{u}(1), \ldots, \mathrm{u}(\mathrm{k}-1)$ then the same state is also controllable in $\mathrm{k}+1$ steps using Input sequence, $u(0), \mathrm{u}(1), \ldots, \mathrm{u}(\mathrm{k}-1)$,

## Definition 5.2

Consider the discrete time system
$\sigma x=A x+B u, y=c x+D u$
With $x \epsilon X=\mathfrak{R}^{\mathrm{n}}, \mathrm{u}(\mathrm{t}) \epsilon \mathfrak{R}^{\mathrm{m}}, \mathrm{y}(\mathrm{t}) \in \mathfrak{R}^{p}$ and $\mathrm{A}, \mathrm{B}, \mathrm{C}$ and D matrices
The subspace $\mathrm{C}=c_{n}$ is the controllability subspace of the system
The system is said to be controllable if $\mathrm{C}=x=\mathfrak{R}^{\mathrm{n}}$, the discrete time system above is controllable if and only if $\operatorname{Im} A^{n} \subseteq \mathfrak{R}$.In particular if A is nilpotent i.e. $A^{q}=0$ For some $\mathrm{q} \leq n$ then for any $\mathrm{B}($ even $\mathrm{B}=0)$

The system is controllable
Note:
That reachable system is controllable but the converse statement does not hold in particular

$$
\mathfrak{R \subseteq C \subseteq X = \mathfrak { R } ^ { \mathrm { n } } , ~}
$$

### 5.3 Construction of input signals

The study of the properties of reachability and controllability leads to the following equation .Is it possible to explicitly constructan input sequence which steers the state of the system from an initial condition?
$x(0)=x_{0}$ to a final condition $x_{f}$ in $\mathrm{k}-$ steps i.e $x(k)=x_{f} ، x_{0}$.i.e
To answer this equation consider the problem of determining an input sequence $u(0), u(1), \ldots, u(k-1)$ such that
$x_{f}-A^{k} x_{0}=R_{k} U_{k-1}$

Where

$$
\mathrm{U}_{\mathrm{k}-1}=\left[\begin{array}{c}
u(k-1) \\
u(k-2) \\
\vdots \\
u(1) \\
u(0)
\end{array}\right] \epsilon \mathfrak{R}^{\mathrm{km}}
$$

Consider now an input signal defined as $U_{n-1}=R_{n}^{1} \mathrm{~V}$
Where V has to be determined using 1 his definition and setting $\mathrm{k}=\mathrm{n}$ equation

$$
x_{f}-A^{k} x_{0}=\mathrm{R}_{\mathrm{k}} \mathrm{U}_{\mathrm{k}-1}
$$

Becomes $x_{f}-A^{n} x_{0}=\mathrm{R}_{\mathrm{n}} \mathrm{R}_{\mathrm{n}}^{1} \mathrm{v}$
Where the matrix $\mathrm{R}_{\mathrm{n}} \mathrm{R}_{\mathrm{n}}^{1}$ is square and invertible Hence a control sequence solving the considered problem in n steps is given by $\mathrm{U}_{n-1}=R_{n}^{1}\left(R_{n} R_{n}^{1}\right)^{-1}\left(x_{f}-A^{k} x_{0}\right)$

It is possible to show that among all input sequence steering the state of the system from $x_{0}$ to $x_{f}$ in n steps the one constructed has minimal norm (energy).

### 5.4 Reachability and Controllability of Continuous Time

## System:

The properties of reachability and controllability for linear, time invariant continuous time systems can be assessed using the same ideas exploited in the case of discrete time systems however the tools are more involved as the input state relation is expressed by means of an integral

$$
x(t)=e^{A t} x_{0}+\int_{0}^{\tau} \mathrm{e}^{(\tau-\mathrm{T})} \mathrm{Bu}(\mathrm{~T}) \mathrm{dT}
$$

Consider the reachability problem i.e. the initial state of the system is $x(0)=0$

And we want to characterize all states $\bar{x}$ that can be reached in some interval of time $t$ i.e. all states such that for some input function $u(t)$

$$
\bar{x}=\int_{0}^{\tau} \mathrm{e}^{(\tau-\mathrm{T})} \mathrm{Bu}(\mathrm{~T}) \mathrm{dT}
$$

Note:
Now that by Clayey -Hamilton theorem
$e^{A t}=\delta_{0}(t) I+\delta_{1}(t) A+\cdots+\delta_{n-1}(t) A^{n-1}$
For some scalar function $\delta_{i}(t)$ Hence
This implies that a state $\bar{x}$ is reachable only if
$\bar{x} \in \operatorname{Im}\left[\mathrm{~B}: \mathrm{AB}: \ldots: A^{n-1} B\right]=\operatorname{ImR}$
We now prove the converse fact i-e that is in the image of R it is reachable to this end define the controllability Gramian
$w_{t}=\int_{0}^{t} e^{A(t-T)} B \grave{B} e^{A(t-T)} d T$
With $\mathrm{t}>0$ and note that $\operatorname{ImR}=\operatorname{ImWt}$
Selecting $\mathrm{U}(\mathrm{T})=B^{\prime} e^{\dot{A}(t-T)} \beta$
Where $\beta$ is a constant vector yield $\bar{x}=w_{\mathrm{t}} \beta$

Hence to assess reach ability of the state $\bar{x} \in \operatorname{ImR}$ it is sufficient to show that equation $\bar{x}=w_{\mathrm{t}} \beta$ has (at least) one solution $\beta$

However this fact holds trivially by condition $\operatorname{ImR}=\operatorname{Im} w_{\mathrm{t}}$

## Remark (5.2)

Unlike the case of discrete time systems where the set of reachable states depends upon the length of the input sequence for continuous time systems if a state is reachable them it is reachable in any (possibly small) interval of time.

## Definition of continuous time system 5.3

$\sigma x=A x+B u, \quad y=C x+D u$ The subspace R is the reach ability sub space of the system the matrix R is the reach ability matrix of the system the system is said to be reach able if $R=X=\Re^{n}$ we summarize the above discussion with a formal statement.

## Proposition (5.4)

Consider the continuous time system $\sigma x=A x+B u, \quad y=C x+D u$ the following statements are equivalent

* The system is reach able
** $\operatorname{rank} \mathrm{R}=\mathrm{n}$
*** For all $\mathrm{t}>0$ the controllability Gramian $w_{\mathrm{t}}$ is positive definite


## Remark (5.3)

If a system is reachable then it is possible to explicitly determine one input signal which steers the state of the system from $x(0)=0$ to any $\bar{x}$ is a give
time $t>0$ in fact determine one such input signal it is sufficient to solve the equation $\bar{x}=w_{\mathrm{t}} \beta$ which if the system is reach able has the unique solution $\beta=w_{t}^{-1} \bar{x}$ I.e. the input signal
$\mathrm{U}(\mathrm{T})=B^{\prime} e^{\hat{A}(t-T)} w_{t}^{-1} \bar{x}$
Steers the state of the system from $x(0)=0$ to $x(t)=\bar{x}$
Similarly to what discussed in construction of input signals it is possible steering the state from 0 to $\bar{x}$ in time t the input signal $\mathrm{u}(\mathrm{T})=B^{\prime} e^{\grave{A}(t-T)} w_{t}^{-1} \bar{x}$ Similar consideration can be done to determine an input signal steering anon Zero initial state to a given final state to discuss the property of controllability note that a state $\bar{x}$ is controllable (to Zero) in time $\mathrm{t}>0$ if there exists an input signal such that
$0=e^{A t} \bar{x}+\int_{0}^{T} e^{A(t-T)} B u(T) d T$
This however implies that $e^{A t} \bar{x} \in \Re$ hence $\bar{x} \in e^{-A t} R$
This implies that the set of controllable states in time $\mathrm{t}>0$ is the set $\mathrm{C}_{\mathrm{t}}=$ $e^{-A t} R$

Which has the same dimension as R by invariability of $e^{-A t}$ for all t and it is contained in R by the fact R is A invariant hence it is trivially $e^{-A t}$.Invariant.

As a consequence for all $\mathrm{t}>0 \mathrm{C}_{\mathrm{t}}=R$ which shows that set C of controllable states does not depend upon $\mathrm{t}>0$ and that continuous time system is controllable if and only if it is reach able C unlike what happens for discrete time system, for which reachability implies but it is not implied by controllability .

## Example5.4

Consider the linear electric network in Figure H. Assume $R_{1}>0$,
$R_{2}>0$,
$\mathrm{L}>0$ and $\mathrm{C}>0$ the input u is the driving voltage y is current supplied.
Let $x_{1}$ be the current through L and $x_{2}$ the voltage a cross C
By Kirchhoff, laws we have
$\dot{x}_{1}=\frac{-R_{1}}{L} x_{1}-\frac{1}{L} x_{2}+\frac{1}{L} u \quad$ ' $\dot{x}_{2}=\frac{1}{C} x_{1}$
And $\mathrm{y}=x_{1}+\frac{1}{R_{2}} u$
There $\mathrm{A}=\left[\begin{array}{cc}\frac{-R_{1}}{L} & \frac{-1}{L} \\ \frac{1}{C} & 0\end{array}\right] \quad \mathrm{B}=\left[\begin{array}{c}\frac{1}{L} \\ 0\end{array}\right]$
The reachability matrix is $\quad \mathrm{R}=\left[\begin{array}{cc}\frac{1}{L} & \frac{-R_{1}}{L^{2}} \\ 0 & \frac{1}{L C}\end{array}\right]$
Hence the system is reach ability and controllable for any L and C

## Example5.5

Consider the linear electric network in Figure G Assume $R_{1}>0, R_{2}>0$ and
$\mathrm{C}>0$ the input u is the deriving voltage and the output y is the current supplied

Let $x_{1}$ be the voltage across C and $x_{2}$ current through L
By Kirchhoff's laws we have
$\dot{x}_{1}=\frac{1}{R_{1} C} x_{1}+\frac{1}{R_{1} C} u \quad ، \quad \dot{x}_{2}=\frac{-R_{2}}{L} x_{2}+\frac{1}{L} u$
And $\mathrm{y}=\frac{-1}{R_{1}} x_{1}+x_{2}+\frac{1}{R_{1}} \mathrm{u}$
The reach ability net work
$\mathrm{R}=\left[\begin{array}{cc}\frac{1}{R_{1} C} & \frac{-1}{R_{1}{ }^{2} C^{2}} \\ \frac{1}{L} & \frac{-R_{2}}{L^{2}}\end{array}\right]$
And $\quad \operatorname{det}(\mathrm{R})=\frac{1}{R_{1} C L}\left(\frac{1}{R_{1} C}-\frac{-R_{2}}{L}\right)$
Hence the system is reachable and controllable provided $R_{1} R_{2} C \neq L$

## 5:5 Canonical form for reachable systems

we focus on single input systems and we show that the property of reachability allows to write the system in a special from known as reachability canonical from consider the system $6 x=A x+B u \quad y=C x+$ Du

With $x \in X=R^{n}, \mathrm{u}(\mathrm{t}) \in R^{m}, \mathrm{y}(\mathrm{t}) \in R^{P}$ and $\mathrm{A}, \mathrm{B}, \mathrm{C}$ and D matrixes of appropriate dimensions and with constant entries with $\mathrm{m}=1$ and suppose the system is reachable i.e the rank of the reachability matrix is equal to $n$

By reachability there is (row) vector L such that $\mathrm{LB}=0, \mathrm{LAB}=0, \mathrm{LA}{ }^{n-1} \mathrm{~B}=0$
$\mathrm{L} A^{n-1} \mathrm{~B}=1$
In fact conditions above formula can be rewritten as
$\mathrm{IR}=\left[\begin{array}{llll}0 & 0 & \ldots . & 0\end{array}\right]$
Hence $\mathrm{L}=\left[\begin{array}{llll}0 & 0 & \ldots . & 0\end{array}\right] R^{-1}$

Is well defined
The vector L has the following important property

## Lemma (5.1)

Let L be as in equation $\mathrm{LR}=\left[\begin{array}{llll}0 & 0 & \ldots & 0\end{array}\right] \mathrm{R}$
Then the square matrix
$\mathrm{T}=\left[\begin{array}{c}L \\ A L \\ \cdot \\ \cdot \\ L A^{n-1} \\ L A^{n-1}\end{array}\right]$ is invertible

## Proof

Consider the matrix
$T_{R}=\left[\begin{array}{c}L \\ A L \\ \cdot \\ \cdot \\ L^{n-1} \\ L A^{n-1}\end{array}\right]\left[B: A B: \ldots \ldots . A^{n-1} B: A^{n-2} B\right]$

And note that by condition IB $=0 \quad \mathrm{LAB}=0 \quad \mathrm{LA}^{\mathrm{n}-2} B \quad \mathrm{LA}^{\mathrm{n}-2} \mathrm{~B}=1$

Which shows that $\left|\operatorname{det}\left(\mathrm{T}_{\mathrm{R}}\right)\right|=1$
Hence T is invertible

The matrix $T$ can be used to define a new set of coordinate $\hat{x}=T_{x}$
To derive the state space representation of the system in the $\hat{x}$ coordinates we could use the general discussions however it is easier to proceed in an alternative way.

For consider the auxiliary signal $\tilde{x}_{1}=\mathrm{L} x$
And note that $6 \tilde{x}_{1}=\mathrm{LA} x=\tilde{x}_{2}$

$$
\begin{gathered}
\text { for } \mathrm{i}=1, \ldots, \mathrm{n}-1 \text { and } \quad 6 \tilde{x}_{i}=\mathrm{L} A^{i} x=\tilde{x}_{i+1} \\
6 \tilde{x}_{n}=\mathrm{L} A^{n} x+\mathrm{u}=\mathrm{LT}^{-1} x^{2}+u
\end{gathered}
$$

As a result in the new coordinates $\tilde{x}$ one has

$$
6 \tilde{x}=A_{r} \tilde{x}+B_{r} \tilde{x}
$$

Where

$$
A_{r}=\left[\begin{array}{cccccc}
0 & 1 & 0 & \ldots & \ldots & \ldots \\
0 & 0 & 1 & \ldots & \ldots & \ldots \\
\vdots & & & & 0 \\
0 & 0 & 0 & \ldots & \ldots & \ldots \\
-\propto_{0} & -\propto_{1} & -\propto_{2} & \ldots & -\propto_{n-1}
\end{array}\right] \quad B_{r}=\left[\begin{array}{c}
0 \\
0 \\
\vdots \\
0 \\
1
\end{array}\right]
$$

$$
\text { And }\left[-\propto_{0}-\propto_{1}-\propto_{2} \ldots-\propto_{n-1}\right]=\operatorname{LA}^{n} \mathrm{~T}^{-1}
$$

Note that for any $\propto_{i}$ the system $6 \tilde{x}=A_{r} \tilde{x}+B_{r} \tilde{x}$ is reachable hence
a system described by equation $t$ $6 \tilde{x}=A_{r} \tilde{x}+B_{r} \tilde{x}$

Is said to be in reachability canonical form
The matrix $A_{r}$ is in comanion from it is worth boting that its characteristic polynomial is
$\mathrm{P}(\mathrm{s})=s^{n}+\propto_{n-1} s^{n-1}+\propto_{n-2} s^{n-2}+\cdots+\propto_{1} s+\propto_{0}$
I.e. it depends only upon the elements of the last row.

## Example 6.5

Consider the system is reachable and controllable
For any L system C
To write the system in reachability canonical from we have to find a (row) vector L such that conditions $\mathrm{LB}=0, \mathrm{LAB}=0 \mathrm{~L} A^{n-2} B=0$
$\mathrm{L} A^{n-1} B=1$
Hold namely
$\mathrm{L}\left[\begin{array}{l}\frac{1}{L} \\ 0\end{array}\right]=0 \quad \mathrm{~L}\left[\begin{array}{c}-\frac{\mathrm{R}_{1}}{\mathrm{~L}^{2}} \\ \frac{1}{\mathrm{Lc}}\end{array}\right]=1 \quad$ yielding $\mathrm{L}=\left[\begin{array}{ll}0 & \mathrm{LC}\end{array}\right]$
Finally $T=\left[\begin{array}{c}L \\ L A\end{array}\right]\left[\begin{array}{cc}0 & \mathrm{LC} \\ \mathrm{L} & 0\end{array}\right]$
And the system in the transformed coordinates in described by
$\dot{\tilde{x}}=\left[\begin{array}{cc}0 & 1 \\ \frac{-1}{L C} & \frac{R_{1}}{L}\end{array}\right] \tilde{x}+\left[\begin{array}{l}0 \\ 1\end{array}\right] \mathrm{u}$
I.e. it is reachability canonical from

Description of non-reachable system:
We study systems which are not 7-6 reachable i.e. systems described by the equation $6 x=\mathrm{A} x+\mathrm{B} u, y=\mathrm{C} x+D u$ and such that rank $=\mathrm{p}<\mathrm{n}$

Under this assumption consider a set of coordinates $\tilde{x}$ such that $x=\mathrm{L} \tilde{x}$
And the matrix L is constructed as follows. The First P columns of L are P linearly independent columns of the matrix R and the last $\mathrm{n}-\mathrm{p}$ columns are selected in such a way that the matrix L is invertible the system is the $\tilde{x}$ coordinates which is algebraically equivalent to the system in the the $x$ coordinates, i.e. described by the equations
$6 \tilde{x}=\tilde{A} \tilde{x}+\tilde{B} \mathrm{u}=\mathrm{L}^{-1} \mathrm{AL} \tilde{x}+\mathrm{L}^{-1} \mathrm{~A} \mathrm{u}$
We now show that because of the way in which n L has been constructed the matrices $\tilde{A}$ and $\tilde{B}$ have a special structure to this end note that
$\mathrm{L} \tilde{A}=A L \quad L \tilde{B}=\mathrm{B}$
And partition the matrices $\mathrm{L}, \mathrm{A}, \tilde{A}, \mathrm{~B}$ and $\tilde{B}$ as
P , And have $\times \mathrm{P}$, have dimensions $\mathrm{n}-\mathrm{P} \times$ Where and have dimensions P an $\mathrm{m} \times$ dimension p

## Chapter six: Results and Conclusions

Model (1) an inverted pendulum mounted
An inverted pendulum mounted on a motor-driven cart is shown in figure 6.1 this is a model, of the attitude control of a space booster on takeoff. The objective of the attitude control problem is to keep the space booster in a vertical position.) The inverted pendulum is unstable in that, it may fall over any time in any direction unless a suitable control force is applied. Here we consider


Figure(6.1)
To derive the equations of motion for the system, consider the free-body diagram shown in Figure 6.1 the rotational motion of the pendulum rod about its center of gravity can be described by

$$
\begin{equation*}
I \theta^{\prime \prime}=v l \sin \theta-H l \cos \theta \tag{6.1}
\end{equation*}
$$

Where $I$ is the moment of inertia of the rod about its center of gravity the horizontal motion of center of gravity of pendulum rod is given by

$$
\begin{equation*}
m \frac{d^{2}}{d t^{2}}(x+I \sin \theta)=H \tag{6.2}
\end{equation*}
$$

The vertical motion of center of gravity of pendulum rod is

$$
\begin{equation*}
m \frac{d^{2}}{d t^{2}}(I \cos \theta)=v-m g \tag{6.3}
\end{equation*}
$$

The horizontal motion of cart is described by

$$
\begin{equation*}
m \frac{d^{2} x}{d t^{2}}=u-H \tag{6.4}
\end{equation*}
$$

Since we must keep the inverted pendulum vertical, we can assume that $\theta(t)$ and $\theta^{\cdot}(t)$ are, small quantities such that $\sin \theta \doteqdot \theta, \cos \theta=1$, and $\theta \dot{\theta}^{2}=0$.

And Then equations (6.1) through (6.3) can be linearized the linearized equations are

$$
\begin{gather*}
I \theta^{\prime \prime}=v I \theta-H l  \tag{6.5}\\
m\left(x^{\prime \prime}+I \theta^{*}\right) \\
=H \tag{6.6}
\end{gather*}
$$

0

$$
\begin{equation*}
=v-m g \tag{6.7}
\end{equation*}
$$

From Equations (6.4) and (6.6), we obtain

$$
\begin{equation*}
(M+m) x^{"}=u \tag{6.8}
\end{equation*}
$$

From Equations (6.5), (6.6), and (6.7), we have

$$
\begin{aligned}
I \ddot{\theta} & =m g l \theta-H l \\
& =m g l \theta-l(m \ddot{x}+m l \ddot{\theta})
\end{aligned}
$$

Or

$$
\begin{equation*}
\left(I+m l^{2}\right) \theta^{\prime \prime}+m l x^{\prime \prime}=m g l \theta \tag{6.10}
\end{equation*}
$$

Equations (6.9) and (6.10) describe the motion of the inverted-pendulum-on-the-cart system they constitute a mathematical model of the systems the inverted pendulum is unstable in that, it may fall over any time in any direction unless a suitable control force is applied. Here we consider Model (2) dashpot system mounted on a massless cart


Figure (6.2)

## Consider the spring - mass -

dashpot system mounted on a massless cart as shown in figure(6.2)
Let us obtain mathematical models of this system by assuming that the cart is standing still for $t<0$ and, the spring-mass-dashpot system on the cart is also standing still for $t<0$ In this system, $\mathrm{u}(\mathrm{t})$ is the. Displacement of the cart and is the input to the system. At $\mathrm{t}=0$, the cart is moved at a constant speed, or $u=$ constant. The displacement $y(t)$ of the mass is the output the displacement is relative to.The ground.) In this system, $m$ denotes the mass, b denotes the viscous-friction coefficient, and k denotes, the spring constant. We assume that the friction force of the dashpot is proportional to $y^{\cdot}-u$ and that the spring is a linear spring; that is, the spring force is proportional to $y-u$.For translational systems, Newton's second law states that

$$
m a=\sum F
$$

where m is amass, a is the acceleration of the mass, and $\quad \sum F$ is the sum of the forces acting on the, mass in the direction of the acceleration a. Applying Newton's second law to the present system, and noting that the cart is massless, we obtain,

$$
m \frac{d^{2} y}{d x^{2}}=-b\left(\frac{d y}{d t}-\frac{d u}{d t}\right)-k(y-u)
$$

Or

$$
m \frac{d^{2} y}{d t^{2}}+b \frac{d y}{d t}+k y=b \frac{d u}{d t}+k u
$$

This equation represents a mathematical model of the system considered. Taking The Laplace, transform of this last equation, assuming zero initial condition, gives. $\left(m s^{2}+b s+k\right) Y(s)=(b s+k) U(s)$

Taking the ratio of $\mathrm{Y}(\mathrm{s})$ to $\mathrm{U}(\mathrm{s})$, we find the transfer function of the system to be transfer function $=G(s)=\frac{Y(s)}{U(s)}=\frac{b s+k}{m s^{2}+b s+k}$

Such a transfer-function representation of a mathematical model is used very frequently in control engineering. Next we shall obtain a state-space model of this system. We shall first compare the differential, equation for this system

$$
\begin{gathered}
y^{\prime \prime}+\frac{b}{m} y^{\prime}+\frac{k}{m} y=\frac{b}{m} u+\frac{k}{m} u \\
y^{\prime \prime}+a_{1} y^{\prime}+a_{2} y=b_{0} u^{\prime \prime}+b_{1} u^{\prime}+b_{2} u
\end{gathered}
$$

And identify $a_{1}, a_{2}, b_{0}, b_{1}$, and $b_{2}$, as follows

$$
a_{1,}=\frac{b}{m}, a_{2=\frac{k}{m}}, b_{0,}=0, b_{1}=\frac{b}{m}, b_{2}=\frac{k}{m}
$$

Where

$$
\begin{gathered}
\beta_{0=b_{0}=0} \\
\beta_{1=b_{1}-a_{1} \beta_{0}}=\frac{b}{m} \\
\beta_{2=b_{2}}-a_{1} \beta_{1}-a_{2} \beta_{0}=\frac{k}{m}-\left(\frac{b}{m}\right)^{2}
\end{gathered}
$$

And

$$
x_{1}=y-\beta_{0} u=y
$$

$x_{2}=x_{1} \cdot \beta_{1} u=x_{1} \cdot-\frac{b}{m} u$
Also

$$
\begin{gathered}
x_{1} \cdot=x_{2}+\beta_{1} u=x_{2}+\frac{b}{m} u \\
x_{2}=-a_{2 x_{1}}-a_{1 x_{1}}+\beta_{2} u=-\frac{k}{m} x_{1}-\frac{b}{m} x_{2}+\left[\frac{k}{m}-\left(\frac{b}{m}\right)^{2}\right] u
\end{gathered}
$$

And the output equation becomes
Or
$\left[\begin{array}{l}x_{1} \\ x_{2}\end{array}\right]=\left[\begin{array}{cc}0 & 1 \\ -\frac{k}{m} & -\frac{b}{m}\end{array}\right]\left[\begin{array}{l}x_{1} \\ x_{2}\end{array}\right]+\left[\begin{array}{c}\frac{b}{m} \\ \frac{k}{m}-\left(\frac{b}{m}\right)^{2}\end{array}\right] \mathrm{u}$
And $y=\left[\begin{array}{ll}1 & 0\end{array}\right]\left[\begin{array}{l}x_{1} \\ x_{2}\end{array}\right]$

Test controllability

$$
M=[B \vdots A B]
$$

Since

$$
\begin{gathered}
B=\left[\begin{array}{c}
\frac{b}{m} \\
\frac{k}{m}-\left(\frac{b}{m}\right)^{2}
\end{array}\right] \\
A B=\left[\begin{array}{cc}
0 & 1 \\
-\frac{k}{m} & -\frac{b}{m}
\end{array}\right]\left[\begin{array}{c}
\frac{b}{m} \\
\frac{k}{m}-\left(\frac{b}{m}\right)^{2}
\end{array}\right]=\left[\begin{array}{c}
\frac{k}{m}-\left(\frac{b}{m}\right)^{2} \\
\frac{-2 k b}{m^{2}}-\left(\frac{b}{m}\right)^{3}
\end{array}\right] \\
\therefore M=\left[\begin{array}{c}
\frac{b}{m} \\
\frac{k}{m}-\left(\frac{b}{m}\right)^{2} \\
\frac{-2 k b}{m^{2}}-\left(\frac{b}{m}\right)^{2}
\end{array}\right] \\
|M|=\frac{b}{m}\left(\frac{-2 k b}{m^{2}}-\left(\frac{b}{m}\right)^{3}\right)-\left(\frac{k}{m}-\left(\frac{b}{m}\right)^{2}\right)\left(\frac{k}{m}-\left(\frac{b}{m}\right)^{2}\right) \neq 0
\end{gathered}
$$

The system is fully controllable

## Results

It has been shown that properties of zero mean error and bounded variance expansion that were seen for single input-single output (SISO) case a design execution is presented to achieved this ideal result However, the SISO control case inherent in practical makings application the knowledge of these effects presented here as undesirable transmission zero and slower than desired settling times is critical in designing a universal any control system or a dynamic system is Time invariant system is shifting the input on the time axis leads to on equivalent shifting of the output along the time axis with no other changes in other words the condition for complete state controllability can be stated in terms of transfer Functions or transfer matrices if $M$ matrix is square determinant not equal zero and $M$ matrix not square we can find sub square matrix determinant not equal zero and Conditions for Complete Observability $M_{Q}$ is full rank .A continuous or discrete time system is stable if every bounded input produces a bounded output (BIBO) and the closed loop transfer function left hand side (BIBO) the Routh criterion of any system is stable all the roots of the characteristic if the elements of the first column of the Routh table have the same sign

## Conclusions

A first solution follows from a general result on the global stabilization of null controllable linear system with delay in the input by bounded control laws with a distributed term. Next, it is shown through alaypunov analysis that the stabilization can be achieved as well neglecting the distributed terms. The multiple input-multiple output control system .It more sensitive to modeling errors, which are in bereft in practical process applications. Any system not we cannot representations square matrix controllability test taking sub square matrix detriment not equal zero or full rank, any system is called stable all roots of characteristic equation left hand side

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