

# Signals and Systems

Robby McKilliam

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# Contents

<b>1 Signals and systems</b>	<b>1</b>
1.1 Properties of signals	1
1.2 Systems (functions of signals)	4
1.3 Some important systems	7
1.4 Properties of systems	9
1.5 Exercises	11
<b>2 Systems modelled by differential equations</b>	<b>17</b>
2.1 Passive electrical circuits	19
2.2 Active electrical circuits	20
2.3 Masses, springs, and dampers	28
2.4 Direct current motors	29
2.5 Exercises	30
<b>3 Linear time invariant systems</b>	<b>35</b>
3.1 Convolution, regular systems and the delta “function”	35
3.2 Properties of convolution	40
3.3 Linear combining and composition	41
3.4 Eigenfunctions and the transfer function	42
3.5 The spectrum	45
3.6 Exercises	49
<b>4 The Laplace transform</b>	<b>55</b>
4.1 The transfer function and the Laplace transform	58
4.2 Solving differential equations	61
4.3 First order systems	62
4.4 Second order systems	65
4.5 Poles, zeros, and stability	68
4.6 Exercises	77
<b>5 The Fourier transform</b>	<b>85</b>
5.1 The inverse transform and the Plancherel theorem	87
5.2 Analogue filters	89

5.3	Real and complex valued sequences	97
5.4	Bandlimited signals	99
5.5	The discrete Fourier transform	100
5.6	Exercises	103
<b>6</b>	<b>Discrete time systems</b>	<b>111</b>
6.1	Discrete time signals (complex valued sequences)	111
6.2	The discrete time impulse response	113
6.3	The z-transform and the discrete time Fourier transform	114
6.4	Finite impulse response filters	114
6.5	Difference equations	114
6.6	Exercises	114

# Chapter 1

## Signals and systems

A **signal** is a function mapping an input variable to some output variable.  
For example

$$\sin(\pi t), \quad \frac{1}{2}t^3, \quad e^{-t^2}$$

all represent **signals** with real input variable  $t \in \mathbb{R}$  and real output variable. These signals are plotted in Figure 1.1. If  $x$  is a signal and  $t$  an input variable we write  $x(t)$  for the output variable corresponding with  $t$ . Signals can be multidimensional. This page is an example of a 2-dimensional signal, the independent variables are the horizontal and vertical position on the page, and the signal maps this position to a colour, in this case either black or white. A moving image such as seen on your television or computer monitor is an example of a 3-dimensional signal, the three independent variables being vertical and horizontal screen position and time. The signal maps each position and time to a colour on the screen. In these notes we focus exclusively on 1-dimensional signals such as those in Figure 1.1 and we will only consider signals where the output variable is real or complex valued. Many of the results presented here can be extended to deal with multidimensional signals.

### 1.1 Properties of signals

A signal  $x$  is **bounded** if there exists a real number  $M$  such that

$$|x(t)| \leq M \quad \text{for all } t \in \mathbb{R}$$

where  $|\cdot|$  denotes the (complex) magnitude. Both  $\sin(\pi t)$  and  $e^{-t^2}$  are examples of bounded signals because  $|\sin(\pi t)| \leq 1$  and  $|e^{-t^2}| \leq 1$  for all  $t \in \mathbb{R}$ . However,  $\frac{1}{2}t^3$  is not bounded because its magnitude grows indefinitely as  $t$  moves away from the origin.

A signal  $x$  is **periodic** if there exists a positive real number  $T$  such that

$$x(t) = x(t + kT) \quad \text{for all } k \in \mathbb{Z} \text{ and } t \in \mathbb{R}.$$

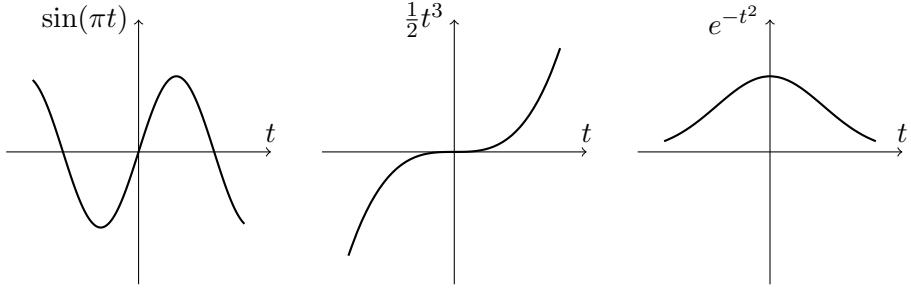


Figure 1.1: 1-dimensional signals

If there exists a smallest such positive  $T$  it is called the **fundamental period** or simply the **period**. For example, the signal  $\sin(\pi t)$  is periodic with period  $T = 2$ . Neither  $\frac{1}{2}t^3$  or  $e^{-t^2}$  are periodic.

A signal  $x$  is **right sided** if there exists a  $T \in \mathbb{R}$  such that  $x(t) = 0$  for all  $t < T$ . Correspondingly  $x$  is **left sided** if  $x(t) = 0$  for all  $T > t$ . For example, the **step function**

$$u(t) = \begin{cases} 1 & t \geq 0 \\ 0 & t < 0 \end{cases} \quad (1.1.1)$$

is right-sided. Its reflection in time  $u(-t)$  is left sided (Figure 1.2). A signal  $x$  is called **finite in time** if it is both left and right sided, that is, if there exists a  $T \in \mathbb{R}$  such that  $x(t) = x(-t) = 0$  for all  $t > T$ . A signal is called **unbounded in time** if it is neither left nor right sided. For example, the signals  $\sin(\pi t)$  and  $e^{-t^2}$  are unbounded in time, but the **rectangular pulse**

$$\Pi(t) = \begin{cases} 1 & |t| < \frac{1}{2} \\ 0 & \text{otherwise} \end{cases} \quad (1.1.2)$$

is finite in time.

A signal  $x$  is **even** (or **symmetric**) if

$$x(t) = x(-t) \quad \text{for all } t \in \mathbb{R}$$

and **odd** (or **antisymmetric**) if

$$x(t) = -x(-t) \quad \text{for all } t \in \mathbb{R}.$$

For example,  $\sin(\pi t)$  and  $\frac{1}{2}t^3$  are odd and  $e^{-t^2}$  is even.

A signal  $x$  is **locally integrable** if

$$\int_a^b |x(t)| dt < \infty$$

for all finite constants  $a$  and  $b$ , where by  $< \infty$  we mean that the integral evaluates to a finite number. An example of a signal that is not locally

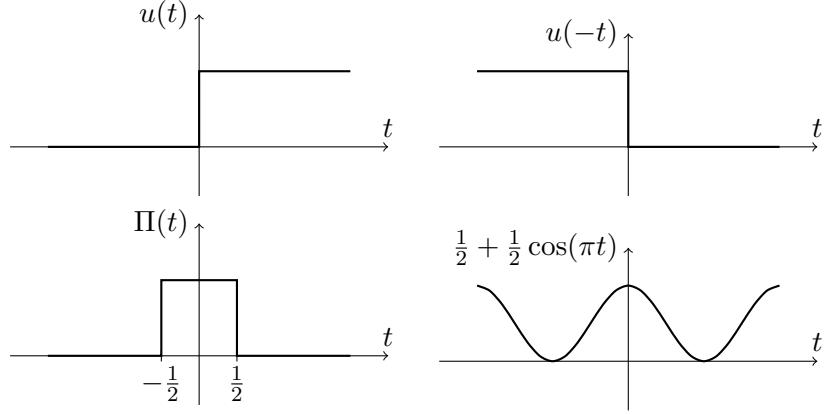


Figure 1.2: The right sided step function  $u(t)$ , its left sided reflection  $u(-t)$ , the finite in time rectangular pulse  $\Pi(t)$  and the unbounded in time signal  $\frac{1}{2} + \frac{1}{2} \cos(\pi t)$ .

integrable is  $x(t) = \frac{1}{t}$  (Exercise 1.2). A signal  $x$  is **absolutely integrable** if

$$\|x\|_1 = \int_{-\infty}^{\infty} |x(t)| dt < \infty. \quad (1.1.3)$$

Here we introduce the notation  $\|x\|_1$  called the  **$L^1$ -norm** of  $x$ . For example  $\sin(\pi t)$  and  $\frac{1}{2}t^3$  are not absolutely integrable, but  $e^{-t^2}$  is because [Nicholas and Yates, 1950]

$$\int_{-\infty}^{\infty} |e^{-t^2}| dt = \int_{-\infty}^{\infty} e^{-t^2} dt = \sqrt{\pi}. \quad (1.1.4)$$

It is common to denote the set of absolutely integrable signals by  $L^1$  or  $L^1(\mathbb{R})$ . So,  $e^{-t^2} \in L^1$  and  $\frac{1}{2}t^3 \notin L^1$ . A signal  $x$  is called **square integrable** if

$$\|x\|_2^2 = \int_{-\infty}^{\infty} |x(t)|^2 dt < \infty.$$

The real number  $\|x\|_2$  is called the  **$L^2$ -norm** of  $x$ . Square integrable signals are also called **energy signals**, and the squared  $L^2$ -norm  $\|x\|_2^2$  is called the **energy** of  $x$ . For example  $\sin(\pi t)$  and  $\frac{1}{2}t^3$  are not energy signals, but  $e^{-t^2}$  is (Exercise 1.5). The set of square integrable signals is often denoted by  $L^2$  or  $L^2(\mathbb{R})$ .

We write  $x = y$  to indicate that two signals  $x$  and  $y$  are **equal pointwise**, that is,  $x(t) = y(t)$  for all  $t \in \mathbb{R}$ . This definition of equality is often stronger than we desire. For example, the step function  $u$  and the signal

$$z(t) = \begin{cases} 1 & t > 0 \\ 0 & t \leq 0 \end{cases}$$

are not equal pointwise because they are not equal at  $t = 0$ , that is,  $u(0) = 1$  and  $z(0) = 0$ . It is useful to identify signals that differ only at isolated points and for this we use a weaker definition of equality. We say that two signals  $x$  and  $y$  are equal **almost everywhere** if

$$\int_a^b |x(t) - y(t)| dt = 0$$

for all finite constants  $a$  and  $b$ . So, in the previous example, while  $u \neq z$  pointwise we do have  $u = z$  almost everywhere. Typically the term almost everywhere is abbreviated to a.e. and one writes

$$x = y \text{ a.e.} \quad \text{or} \quad x(t) = y(t) \text{ a.e.}$$

to indicate that the signals  $x$  and  $y$  are equal almost everywhere.

## 1.2 Systems (functions of signals)

A **system** is a function that maps a signal to another signal. For example

$$x(t) + 3x(t-1), \quad \int_0^1 x(t-\tau) d\tau, \quad \frac{1}{x(t)}, \quad \frac{d}{dt} x(t)$$

represent systems, each mapping the signal  $x$  to another signal. Consider the electric circuit in Figure 1.3 called a **voltage divider**. If the voltage at time  $t$  is  $x(t)$  then, by Ohm's law, the current at time  $t$  satisfies

$$i(t) = \frac{1}{R_1 + R_2} x(t),$$

and the voltage over the resistor  $R_2$  is

$$y(t) = R_2 i(t) = \frac{R_2}{R_1 + R_2} x(t). \quad (1.2.1)$$

The circuit can be considered as a system mapping the signal  $x$  representing the voltage to the signal  $i = \frac{1}{R_1 + R_2} x$  representing the current, or a system mapping  $x$  to the signal  $y = \frac{R_2}{R_1 + R_2} x$  representing the voltage over resistor  $R_2$ .

We denote systems with capital letters such as  $H$  and  $G$ . A system  $H$  is a function that maps a signal  $x$  to another signal denoted  $H(x)$ . We call  $x$  the **input signal** and  $H(x)$  the **output signal** or the **response** of system  $H$  to signal  $x$ . The value of the signal  $H(x)$  at  $t$  is denoted by  $H(x, t)$  or  $H(x, t)$  and we do not distinguish between these notations. It is sometimes useful to depict systems with a block diagram. Figure 1.4 is a simple block diagram showing the input and output signals of a system  $H$ .

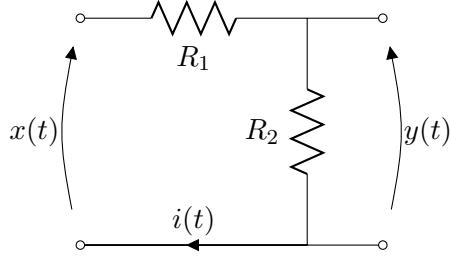
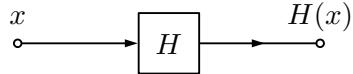
Figure 1.3: A **voltage divider** circuit.

Figure 1.4: System block diagram with input signal \$x\$ and output signal \$H(x)\$.

The electric circuit in Figure 1.3 corresponds with the system

$$H(x) = \frac{R_2}{R_1 + R_2} x = y.$$

This system multiplies the input signal \$x\$ by \$\frac{R\_2}{R\_1 + R\_2}\$. This brings us to our first practical test.

**Test 1 (Voltage divider)** In this test we construct the voltage divider from Figure 1.3 on a breadboard with resistors \$R\_1 \approx 100\Omega\$ and \$R\_2 \approx 470\Omega\$ with values accurate to within 5%. Using a computer soundcard (an approximation of) the voltage signal

$$x(t) = \sin(2\pi f_1 t) \quad \text{with} \quad f_1 = 100$$

is passed through the circuit. The approximation is generated by sampling \$x(t)\$ at rate \$F = \frac{1}{P} = 44100\text{Hz}\$ to generate samples

$$x(nP) \quad n = 0, \dots, 2F$$

corresponding to approximately 2 seconds of signal. These samples are passed to the soundcard which starts playback. The voltage over resistor \$R\_2\$ is recorded (also using the soundcard) that returns a list of samples \$y\_1, \dots, y\_L\$ taken at rate \$F\$. The voltage over \$R\_2\$ can be (approximately) reconstructed from these samples as

$$\tilde{y}(t) = \sum_{\ell=1}^L y_\ell \operatorname{sinc}(Ft - \ell) \tag{1.2.2}$$

where

$$\operatorname{sinc}(t) = \frac{\sin(\pi t)}{\pi t} \tag{1.2.3}$$

is called the **sinc function** and is plotted in Figure 5.1. We will justify this reconstruction in Section ???. Simultaneously the (stereo) soundcard is used to record the input voltage  $x$  producing samples  $x_1, \dots, x_L$  taken at rate  $F$ . An approximation of the input signal is

$$\tilde{x}(t) = \sum_{\ell=1}^L x_\ell \operatorname{sinc}(Ft - \ell). \quad (1.2.4)$$

In view of (1.2.1) we would expect the approximate relationship

$$\tilde{y} \approx \frac{R_2}{R_1 + R_2} \tilde{x} = \frac{47}{57} \tilde{x}.$$

A plot of  $\tilde{y}$ ,  $\tilde{x}$  and  $\frac{47}{57} \tilde{x}$  over a 20ms period from 1s to 1.02s is given in Figure 1.5. The hypothesised output signal  $\frac{47}{57} \tilde{x}$  does not match the observed output signal  $\tilde{y}$ . A primary reason is that the circuitry inside the soundcard itself cannot be ignored. When deriving the equation for the voltage divider we implicitly assumed that current flows through the output of the soundcard without resistance (a short circuit), and that no current flows through the input device of the soundcard (an open circuit). These assumptions are not realistic. Modelling the circuitry in the sound card wont be attempted here. In Section 2.2 we will construct circuits that contain external sources of power (active circuits). These are less sensitive to the circuitry inside the soundcard.

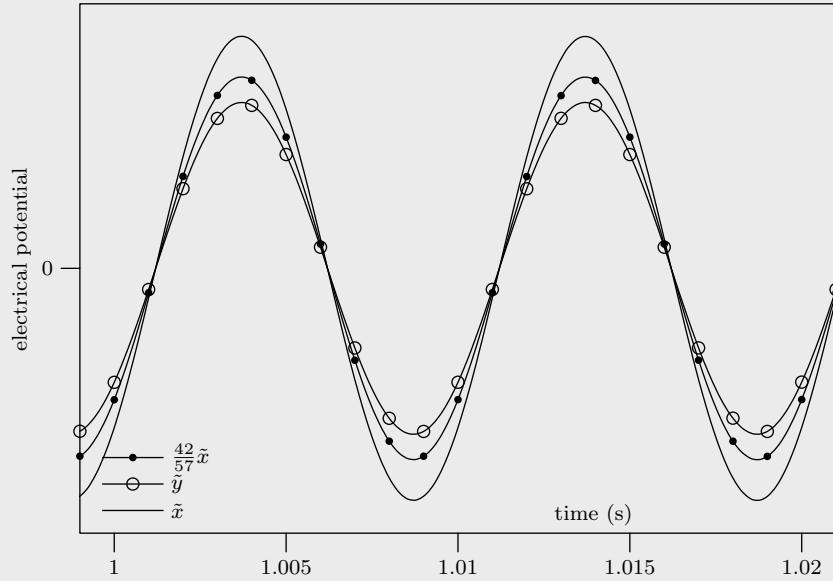


Figure 1.5: Plot of reconstructed input signal  $\tilde{x}$  (solid line), output signal  $\tilde{y}$  (solid line with circle) and hypothesised output signal  $\frac{47}{57}\tilde{x}$  (solid line with dot) for the voltage divider circuit in Figure 1.3. The hypothesised signal does not match  $\tilde{y}$ . One reason is that the model does not take account of the circuitry inside the soundcard.

Not all signals can be input to all systems. For example, the system

$$H(x, t) = \frac{1}{x(t)}$$

is not defined at those  $t$  where  $x(t) = 0$  because we cannot divide by zero. Another example is the system

$$I_\infty(x, t) = \int_{-\infty}^t x(\tau) d\tau, \quad (1.2.5)$$

called an **integrator**. The signal  $x(t) = 1$  cannot be input to the integrator because the integral  $\int_{-\infty}^t dt$  is not finite for any  $t$ .

When specifying a system it is necessary to also specify a set of signals that can be input. This is called a **domain** for the system. We are free to choose the domain at our convenience. For example, a domain for the system  $H(x, t) = \frac{1}{x(t)}$  is the set of signals  $x(t)$  which are not zero for any  $t$ . An example of a domain for the integrator  $I_\infty$  is the set  $L^1$  of absolutely integrable signals because, if  $x$  is absolutely integrable, then

$$|I_\infty(x, t)| \leq \left| \int_{-\infty}^t x(\tau) d\tau \right| \leq \int_{-\infty}^t |x(\tau)| d\tau < \int_{-\infty}^{\infty} |x(\tau)| d\tau = \|x\|_1 < \infty$$

and so,  $I_\infty(x, t)$  is finite for all  $t$ . In this text, the domain used for a given system will usually be obvious from the context in which the system is defined. For this reason we will not usually state the domain explicitly. We will only do so if there is chance for confusion.

### 1.3 Some important systems

The system

$$T_\tau(x, t) = x(t - \tau)$$

is called a **time-shifter**. This system shifts the input signal along the  $t$  axis ('time' axis) by  $\tau$ . When  $\tau$  is positive  $T_\tau$  delays the input signal by  $\tau$ . The time-shifter will appear so regularly in this course that we use the special notation  $T_\tau$  to represent it. Figure 1.6 depicts the action of time-shifters

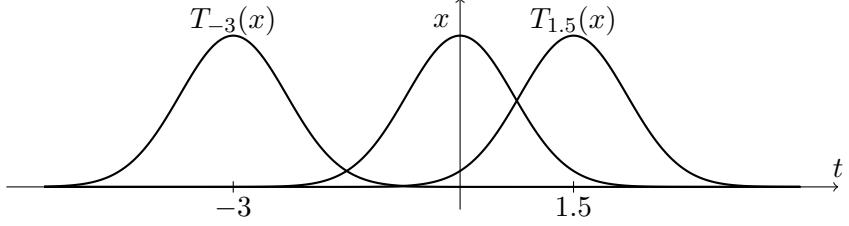


Figure 1.6: Time-shifter system  $T_{1.5}(x, t) = x(t - 1.5)$  and  $T_{-3}(x, t) = x(t + 3)$  acting on the signal  $x(t) = e^{-t^2}$ .

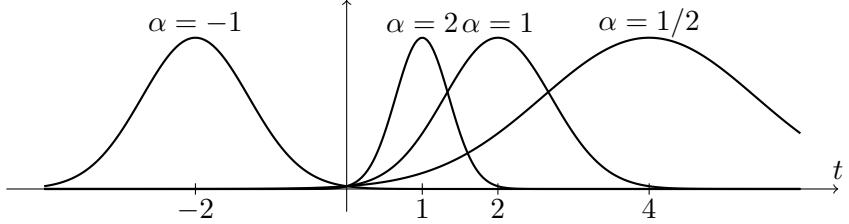


Figure 1.7: Time-scaler system  $H(x, t) = x(\alpha t)$  for  $\alpha = -1, \frac{1}{2}, 1$  and  $2$  acting on the signal  $x(t) = e^{-(t-2)^2}$ .

$T_{1.5}$  and  $T_{-3}$  on the signal  $x(t) = e^{-t^2}$ . When  $\tau = 0$  the time-shifter is the **identity system**

$$T_0(x) = x$$

that maps the signal  $x$  to itself.

Another important system is the **time-scaler** that has the form

$$H(x, t) = x(\alpha t), \quad \alpha \in \mathbb{R}.$$

Figure 1.7 depicts the action of a time-scaler with a number of values for  $\alpha$ . When  $\alpha = -1$  the time-scaler reflects the input signal in the time axis. When  $\alpha = 1$  the time-scaler is the identity system  $T_0$ .

Another system we regularly encounter is the **differentiator**

$$D(x, t) = \frac{d}{dt}x(t),$$

that returns the derivative of the input signal. We also define a  $k$ th differentiator

$$D^k(x, t) = \frac{d^k}{dt^k}x(t)$$

that returns the  $k$ th derivative of the input signal.

A related system is the **integrator**

$$I_a(x, t) = \int_{-a}^t x(\tau) d\tau.$$

The parameter  $a$  describes the lower bound of the integral. In this course it will often be that  $a = \infty$ . For example, the response of the integrator  $I_\infty$  to the signal  $tu(t)$  where  $u$  is the step function (1.1.1) is

$$\int_{-\infty}^t \tau u(\tau) d\tau = \begin{cases} \int_0^t \tau d\tau = \frac{t^2}{2} & t > 0 \\ 0 & t \leq 0. \end{cases}$$

Observe that the integrator  $I_\infty$  cannot be applied to the signal  $x(t) = t$  because  $\int_{-\infty}^t \tau d\tau$  is not finite for any  $t$ . A domain for  $I_\infty$  would not contain the signal  $x(t) = t$ .

## 1.4 Properties of systems

In this section we define a number of important properties that systems can possess. In what follows  $H$  will be a system and the phrase “for all signals” will mean for all signals inside some domain for  $H$ .

A system  $H$  is called **memoryless** if the output signal  $H(x)$  at time  $t$  depends only on the input signal  $x$  at time  $t$ . For example  $\frac{1}{x(t)}$  and the identity system  $T_0$  are memoryless, but

$$x(t) + 3x(t-1) \quad \text{and} \quad \int_0^1 x(t-\tau) d\tau$$

are not. A time-shifter  $T_\tau$  with  $\tau \neq 0$  is not memoryless.

A system  $H$  is **causal** if the output signal  $H(x)$  at time  $t$  depends on the input signal only at times less than or equal to  $t$ . Memoryless systems such as  $\frac{1}{x(t)}$  and  $T_0$  are also causal. Time-shifters  $T_\tau$  are causal when  $\tau \geq 0$ , but are not causal when  $\tau < 0$ . The systems

$$x(t) + 3x(t-1) \quad \text{and} \quad \int_0^1 x(t-\tau) d\tau$$

are causal, but the systems

$$x(t) + 3x(t+1) \quad \text{and} \quad \int_0^1 x(t+\tau) d\tau$$

are not causal.

A system  $H$  is called **bounded-input-bounded-output (BIBO) stable** or just **stable** if the output signal  $H(x)$  is bounded whenever the input signal  $x$  is bounded. That is,  $H$  is stable if for every positive real number  $M$  there exists a positive real number  $K$  such that for all signals  $x$  satisfying

$$|x(t)| < M \quad \text{for all } t \in \mathbb{R},$$

it also holds that

$$|H(x, t)| < K \quad \text{for all } t \in \mathbb{R}.$$

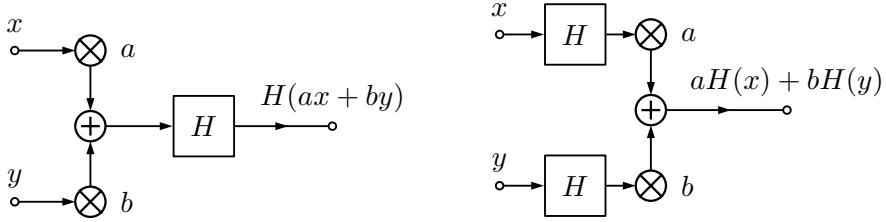


Figure 1.8: If  $H$  is a linear system the outputs of these two diagrams are the same signal, i.e.  $H(ax + by) = aH(x) + bH(y)$ .

For example, the system  $x(t) + 3x(t - 1)$  is stable with  $K = 4M$  since if  $|x(t)| < M$  then

$$|x(t) + 3x(t - 1)| \leq |x(t)| + 3|x(t - 1)| < 4M = K.$$

The integrator  $I_a$  for any  $a \in \mathbb{R}$  and differentiator  $D$  are not stable (Exercises 1.6 and 1.7).

A system  $H$  is **linear** if

$$H(ax + by) = aH(x) + bH(y)$$

for all signals  $x$  and  $y$  and all complex numbers  $a$  and  $b$ . That is, a linear system has the property: If the input consists of a weighted sum of signals, then the output consists of the same weighted sum of the responses of the system to those signals. Figure 1.8 indicates the linearity property using a block diagram. For example, the differentiator is linear because

$$\begin{aligned} D(ax + by, t) &= \frac{d}{dt}(ax(t) + by(t)) \\ &= a\frac{d}{dt}x(t) + b\frac{d}{dt}y(t) \\ &= aD(x, t) + bD(y, t) \end{aligned}$$

whenever both  $x$  and  $y$  are differentiable. However, the system  $H(x, t) = \frac{1}{x(t)}$  is not linear because

$$H(ax + by, t) = \frac{1}{ax(t) + by(t)} \neq \frac{a}{x(t)} + \frac{b}{y(t)} = aH(x, t) + bH(y, t)$$

in general.

The property of linearity trivially generalises to more than two signals. For example, if  $x_1, \dots, x_k$  are signals and  $a_1, \dots, a_k$  are complex numbers for some finite  $k$ , then

$$H(a_1x_1 + \dots + a_kx_k) = a_1H(x_1) + \dots + a_kH(x_k).$$

A system  $H$  is **time invariant** if

$$H(T_\tau(x), t) = H(x, t - \tau)$$

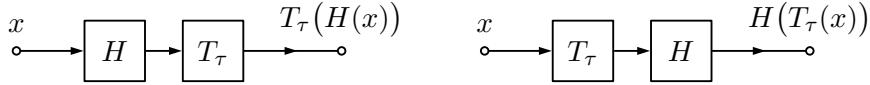


Figure 1.9: If  $H$  is a time-invariant system the outputs of these two diagrams are the same signal, i.e.  $H(T_\tau(x)) = T_\tau(H(x))$ .

for all signals  $x$  and all time-shifts  $\tau \in \mathbb{R}$ . That is, a system is time-invariant if time shifting the input signal results in the same time-shift of the output signal. Equivalently,  $H$  is time-invariant if it commutes with the time-shifter  $T_\tau$ , that is, if

$$H(T_\tau(x)) = T_\tau(H(x))$$

for all  $\tau \in \mathbb{R}$  and all signals  $x$ . Figure 1.9 represents the property of time-invariance with a block diagram.

## 1.5 Exercises

- 1.1. State whether the step function  $u(t)$  is bounded, periodic, absolutely integrable, an energy signal. **Solution:** The magnitude of  $u$  is less than or equal to one, so the signal is bounded. The signal is not period, since for any hypothesised period  $T > 0$  we have  $u(T) = 1$  but  $u(0) = 0$ . The signal is not absolutely integrable, nor an energy signal since

$$\|u\|_1 = \|u\|_2 = \int_{-\infty}^{\infty} |u(t)| dt = \int_0^{\infty} dt$$

is not finite.

- 1.2. Show that the signal  $t^2$  is locally integrable, but that the signal  $\frac{1}{t^2}$  is not. **Solution:** For any  $a$  and  $b$

$$\int_a^b t^2 dt = \frac{b^3}{3} - \frac{a^3}{3}$$

exists, and so  $t^2$  is locally integrable. Put  $a = 0$  and  $b > 0$  and

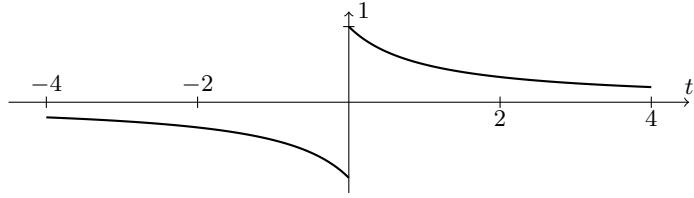
$$\int_0^b \frac{1}{t} dt = \log(b) - \lim_{t \rightarrow -\infty} \log(t) = -\infty$$

where  $\log$  is natural logarithm. The limit above diverges and so  $\frac{1}{t}$  is not locally integrable.

- 1.3. Plot the signal

$$x(t) = \begin{cases} \frac{1}{t+1} & t > 0 \\ \frac{1}{t-1} & t \leq 0. \end{cases}$$

State whether it is: bounded, locally integrable, absolutely integrable, square integrable. **Solution:**



The signal is bounded since  $|x(t)| < M$  for any  $M > 1$ . The signal is locally integrable because it is bounded, i.e., for any finite constants  $a$  and  $b$

$$\int_a^b |x(t)| dt < \int_a^b M dt = (b-a)M < \infty.$$

The signal  $x$  is not absolutely integrable since

$$\begin{aligned} \|x\|_1 &= \int_{-\infty}^{\infty} |x(t)| dt \\ &= 2 \int_0^{\infty} \frac{1}{t+1} dt \\ &= 2 \int_1^{\infty} \frac{1}{t} dt \\ &= \log(1) + \lim_{t \rightarrow \infty} \log(t) \end{aligned}$$

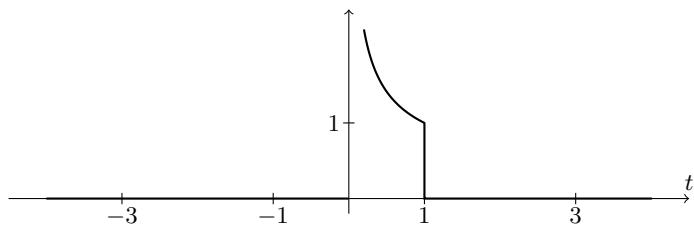
and the limit diverges. The signal is square integrable since

$$\begin{aligned} \|x\|_2 &= \int_{-\infty}^{\infty} |x(t)|^2 dt \\ &= 2 \int_0^{\infty} \frac{1}{(t+1)^2} dt \\ &= 2 \int_1^{\infty} \frac{1}{t^2} dt \\ &= 1 - \lim_{t \rightarrow \infty} \frac{1}{t} = 1. \end{aligned}$$

#### 1.4. Plot the signal

$$x(t) = \begin{cases} \frac{1}{\sqrt{t}} & 0 < t \leq 1 \\ 0 & \text{otherwise.} \end{cases}$$

Show that  $x$  is absolutely integrable, but not square integrable. **Solution:**



The integral

$$\|x\|_1 = \int_{-\infty}^{\infty} |x(t)| dt = \int_0^1 t^{-1/2} dt = [2\sqrt{t}]_0^1 = 2$$

and so  $x$  is absolutely integrable. The integral

$$\|x\|_2 = \int_{-\infty}^{\infty} |x(t)| dt = \int_0^1 t^{-1} dt = [\log(t)]_0^1 = \log(1) - \lim_{t \rightarrow 0} \log(t) = \infty$$

and so  $x$  is not square integrable.

- 1.5. Compute the energy of the signal  $e^{-\alpha^2 t^2}$  (Hint: use equation (1.1.4) on page 3 and a change of variables). **Solution:** From (1.1.4) we see the energy of  $e^{-t^2}$  is  $\sqrt{\pi}$ . Now

$$\int_{-\infty}^{\infty} e^{-\alpha^2 t^2} dt = \frac{1}{\alpha} \int_{-\infty}^{\infty} e^{-\tau^2} d\tau = \frac{\sqrt{\pi}}{\alpha}$$

by the change of variables  $\tau = \alpha t$ .

- 1.6. Show that the integrator  $I_a$  for any  $a \in \mathbb{R}$  is not stable. **Solution:** Put  $M > 1$ . The time-shifted step function  $u(t+a)$  is bounded below  $M$ , i.e.  $|u(t+a)| \leq 1 < M$  for all  $t \in \mathbb{R}$ . However, the response of the integrator  $I_a$  to  $u(t+a)$  is

$$I_a(u(t+a)) = \int_{-a}^t u(\tau+a) d\tau = \begin{cases} \int_{-a}^t d\tau = t + a & t \geq -a \\ 0 & t < -a, \end{cases}$$

and this is not a bounded signal, that is, for every  $K$  we have  $t + a > K$  whenever  $t > K - a$ .

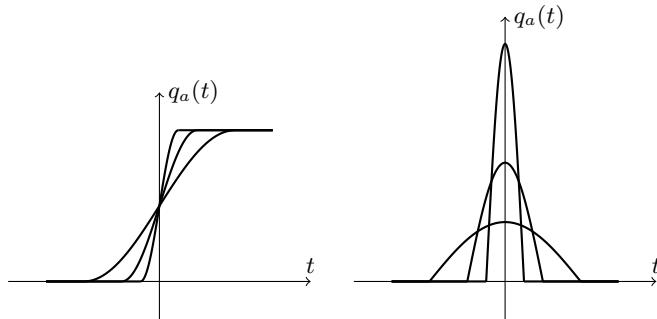
- 1.7. Show that the differentiator system  $D$  is not stable. **Solution:** Put  $M > 2$ . Define the signal

$$q_a(t) = \begin{cases} 0 & 2t < -a \\ 1 + \sin\left(\frac{\pi t}{a}\right) & -a < 2t < a \\ 2 & 2t > a, \end{cases}$$

and observe that  $q_a$  is differentiable and bounded below  $M$ . The response of the differentiator  $D$  to  $q_a$  is

$$D(q_a, t) = \begin{cases} 0 & 2t < -a \\ \frac{\pi}{a} \cos\left(\frac{\pi t}{a}\right) & -a < 2t < a \\ 1 & 2t > a. \end{cases}$$

The signal  $p_a$  and the response  $D(p_a)$  are plotted below for  $a = \frac{1}{2}, 1$  and  $2$ . The response  $D(p_a)$  obtains a maximum amplitude of  $\frac{\pi}{a}$  at  $t = 0$ . So  $D$  is not bounded because for every  $K$  we can choose  $a < \frac{\pi}{K}$  so that  $\frac{\pi}{a} > K$ .



- 1.8. Show that the time-shifter is linear and time-invariant and that the time-scaler is linear, but not time invariant **Solution:** The time shifter  $T_\tau$  is time invariant since

$$T_k(T_\tau(x)) = T_k(x(t - \tau)) = x(t - \tau - k) = x(t - k - \tau) = T_\tau(x(t - k)) = T_\tau(T_k(x))$$

for all signals  $x$ , that is, time shifters commute with time shifters. The time shifter is linear because

$$T_\tau(ax + by) = ax(t - \tau) + by(t - \tau) = aT_\tau(x) + bT_\tau(y).$$

The time-scaler  $H(x) = x(\alpha t)$  is linear because

$$H(ax + by) = ax(\alpha t) + by(\alpha t) = aH(x) + bH(y).$$

The system is not time invariant because

$$H(T_\tau(x)) = H(x(t - \tau)) = x(\alpha t - \tau)$$

but

$$T_\tau(H(x)) = T_\tau(x(\alpha t)) = x(\alpha(t - \tau)) = x(\alpha t - \alpha\tau),$$

and these signals are not equal in general. For example consider the rectangular pulse  $\Pi$ . With time scaling parameter  $\alpha = 2$  and time shift  $\tau = 1$ ,

$$H(T_1(\Pi)) = \Pi(2t - 1) \neq \Pi(2t - 2) = T_1(H(\Pi)).$$

- 1.9. Show that the integrator  $I_c$  with  $c$  finite is linear, but not time-invariant. **Solution:** The system is linear because

$$\begin{aligned} I_c(ax + by) &= \int_{-c}^t ax(\tau) + by(\tau)d\tau \\ &= a \int_{-c}^t x(\tau)d\tau + b \int_{-c}^t y(\tau)d\tau \\ &= aI_c(x) + bI_c(y). \end{aligned}$$

The system is not time invariant because

$$T_k(I_c(x)) = I_c(x, t - k) = \int_{-c}^{t-k} x(\tau)d\tau$$

but

$$I_c(T_k(x)) = \int_{-c}^t x(\tau - k)d\tau.$$

We now need only find some signal  $x$  for which the integrals on the right hand side of the above equations are not equal. Choose the signal  $x = 1$ , i.e., the signal that is equal to 1 for all time. In this case

$$T_k(I_c(1)) = \int_{-c}^{t-k} d\tau = t - k + c \neq t + c = \int_{-c-k}^{t-k} d\tau = I_c(T_k(1)) \quad \text{when } k \neq 0.$$

- 1.10. Show that the integrator  $I_\infty$  is linear and time invariant. **Solution:** The system is linear because

$$\begin{aligned} I_\infty(ax + by) &= \int_{-\infty}^t ax(\tau) + by(\tau)d\tau \\ &= a \int_{-\infty}^t x(\tau)d\tau + b \int_{-\infty}^t y(\tau)d\tau \\ &= aI_\infty(x) + bI_\infty(y). \end{aligned}$$

The system is time invariant because

$$T_k(I_\infty(x)) = I_\infty(x, t - k) = \int_{-\infty}^{t-k} x(\tau) d\tau,$$

and

$$I_\infty(T_k(x)) = \int_{-\infty}^t x(\tau - k) d\tau = \int_{-\infty}^{t-k} x(\tau) d\tau.$$

- 1.11. State whether the system  $H(x) = x + 1$  is linear, time-invariant, stable.

**Solution:** It is not linear because for any signal  $x$  and real number  $a \neq 1$ ,

$$H(ax) = ax + 1 \neq aH(x) = a(x + 1) = ax + a.$$

It is time-invariant because

$$H(T_\tau(x)) = x(t - \tau) + 1 = T_\tau(x + 1) = T_\tau(H(x)).$$

It is stable because for any signal  $x$  with  $x(t) < M$  for all  $t \in \mathbb{R}$ ,

$$H(x, t) = x(t) + 1 < M + 1 \quad \text{for all } t \in \mathbb{R}.$$

- 1.12. State whether the system  $H(x) = 0$  is linear, time-invariant, stable.

**Solution:** It is linear because

$$H(ax + by) = 0 = aH(x) + bH(y) = 0.$$

It is time-invariant because

$$H(T_\tau(x))(t) = 0 = H(x, t - \tau).$$

It is stable because for any  $M > 0$ ,

$$H(x, t) = 0 < M \quad \text{for all } t \in \mathbb{R} \text{ and all signals } x.$$

- 1.13. State whether the system  $H(x) = 1$  is linear, time-invariant, stable.

**Solution:** It is not linear because for any signal  $x$  and real number  $a \neq 1$

$$H(ax) = 1 \neq aH(x) = a.$$

It is time-invariant because

$$H(T_\tau(x)) = 1 = T_\tau(1) = T_\tau(H(x)).$$

It is stable because for any  $K > 1$ ,

$$|H(x)(t)| = 1 < K \quad \text{for all } t \in \mathbb{R} \text{ and all signals } x.$$

- 1.14. Let  $x$  be a signal with period  $T$  that is not equal to zero almost everywhere. Show that  $x$  is not absolutely integrable. **Solution:** This is plain and does not really require further explanation, but I've found some students desire more rigour.

Since  $x$  does not equal to zero almost everywhere there exist some finite real numbers  $a$  and  $b$  such that  $\int_a^b |x(t)| dt = C > 0$ . Let  $k$  be an integer such  $-kT < a$  and  $kT > b$  so that the integral over  $2k + 1$  periods

$$\int_{-kT}^{kT} |x(t)| dt \geq \int_a^b |x(t)| dt = C > 0.$$

Now, since  $x$  has period  $T$

$$\int_{-ckT}^{ckT} |x(t)| dt = (2c+1) \int_{-kT}^{kT} |x(t)| dt \geq (2c+1)C > 0$$

for integers  $c$  and since this integral is increasing monotonically with  $c$  we have  $\int_{-ckT}^{ckT} |x(t)| dt \geq \lfloor 2c+1 \rfloor C$  for all  $c \in \mathbb{R}$  where  $\lfloor 2c+1 \rfloor$  denotes the largest integer less than or equal to  $2c+1$ . Now,

$$\|x\|_1 = \int_{-\infty}^{\infty} |x(t)| dt = \lim_{c \rightarrow \infty} \int_{-ckT}^{ckT} |x(t)| dt \geq \lim_{c \rightarrow \infty} \lfloor 2c+1 \rfloor C = \infty,$$

and so,  $x$  is not absolutely integrable.

## Chapter 2

# Systems modelled by differential equations

Systems of particular interest in this text are those where the input signal  $x$  and output signal  $y$  are related by a linear differential equation with constant coefficients, that is, an equation of the form

$$\sum_{\ell=0}^m a_\ell \frac{d^\ell}{dt^\ell} x(t) = \sum_{\ell=0}^k b_\ell \frac{d^\ell}{dt^\ell} y(t),$$

where  $a_0, \dots, a_m$  and  $b_0, \dots, b_k$  are real or complex numbers. In what follows we use the differentiator system  $D$  rather than the notation  $\frac{d}{dt}$  to represent differentiation. To represent the  $\ell$ th derivative we write  $D^\ell$  instead of  $\frac{d^\ell}{dt^\ell}$ . Using this notation the differential equation above is

$$\sum_{\ell=0}^m a_\ell D^\ell(x) = \sum_{\ell=0}^k b_\ell D^\ell(y). \quad (2.0.1)$$

Equations of this form can be used to model a large number of electrical, mechanical and other real world devices. For example, consider the resistor and capacitor (RC) circuit in Figure 2.1. Let the signal  $v_R$  represent the voltage over the resistor and  $i$  the current through both resistor and capacitor. The voltage signals satisfy

$$x = y + v_R,$$

and the current satisfies both

$$v_R = Ri \quad \text{and} \quad i = CD(y).$$

Combining these equations,

$$x = y + RCD(y) \quad (2.0.2)$$

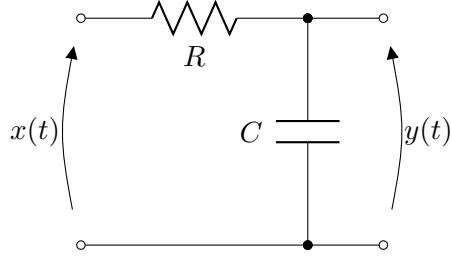


Figure 2.1: An electrical circuit with resistor and capacitor in series, otherwise known as an **RC circuit**.

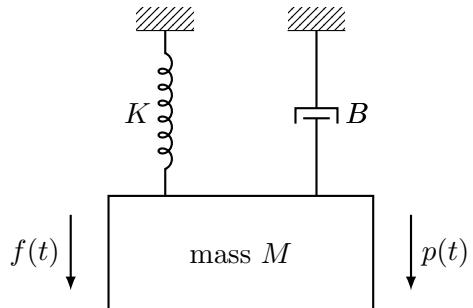


Figure 2.2: A mechanical mass-spring-damper system

that is in the form of (2.0.1).

As another example, consider the mass-spring-damper in Figure 2.2. A force represented by the signal  $f$  is externally applied to the mass, and the position of the mass is represented by the signal  $p$ . The spring exerts force  $-Kp$  that is proportional to the position of the mass, and the damper exerts force  $-BD(p)$  that is proportional to the velocity of the mass. The cumulative force exerted on the mass is

$$f_m = f - Kp - BD(p)$$

and by Newton's law the acceleration of the mass  $D^2(p)$  satisfies

$$MD^2(p) = f_m = f - Kp - BD(p).$$

We obtain the differential equation

$$f = Kp + BD(p) + MD^2(p) \quad (2.0.3)$$

that is in the form of (2.0.1) if we put  $x = f$  and  $y = p$ . Given  $p$  we can readily solve for the corresponding force  $f$ . As a concrete example, let the spring constant, damping constant and mass be  $K = B = M = 1$ . If the position satisfies  $p(t) = e^{-t^2}$ , then the corresponding force satisfies

$$f(t) = e^{-t^2}(4t^2 - 2t - 1).$$

Figure 2.3: A solution to the mass-spring-damper system with  $K = B = M = 1$ . The position is  $p(t) = e^{-t^2}$  with corresponding force  $f(t) = e^{-t^2}(4t^2 - 2t - 1)$ .

Figure 2.3 depicts these signals.

What happens if a particular force signal  $f$  is applied to the mass? For example, say we apply the force

$$f(t) = \Pi(t - \frac{1}{2}) = \begin{cases} 1 & 0 < t \leq 1 \\ 0 & \text{otherwise.} \end{cases}$$

What is the corresponding position signal  $p$ ? We are not yet ready to answer this question, but will be later (Exercise 4.12).

In both the mechanical mass-spring-damper system in Figure 2.2 and the electrical RC circuit in Figure 2.1 we obtain a differential equation relating the input signal  $x$  with the output signal  $y$ . The equations do not specify the output signal  $y$  explicitly in terms of the input signal  $x$ , that is, they do not explicitly define a system  $H$  such  $y = H(x)$ . As they are, the differential equations do not provide as much information about the behaviour of the system as we would like. For example, is the system stable? We will be able to obtain much more information about these systems when the **Laplace transform** is introduced in Chapter 4. The remainder of this chapter details the construction of differential equations that model various mechanical, electrical, and electro-mechanical systems. We will use the systems constructed here as examples throughout the course.

## 2.1 Passive electrical circuits

**Passive electrical circuits** require no sources of power other than the input signal itself. For example, the voltage divider in Figure 1.3 and the RC circuit in Figure 2.1 are passive circuits. Another common passive electrical circuit is the resistor, capacitor and inductor (RLC) circuit depicted in Figure 2.4. In this circuit we let the output signal  $y$  be the voltage over the resistor. Let  $v_C$  represent the voltage over the capacitor and  $v_L$  the voltage over the inductor and let  $i$  be the current. We have

$$y = Ri, \quad i = CD(v_C), \quad v_L = LD(i),$$

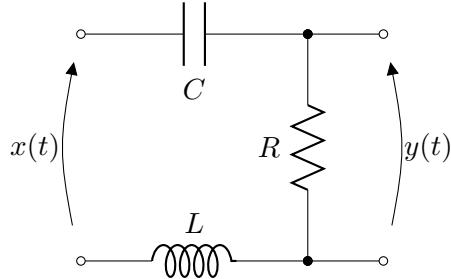


Figure 2.4: An electrical circuit with resistor, capacitor and inductor in series, otherwise known as an **RLC circuit**.

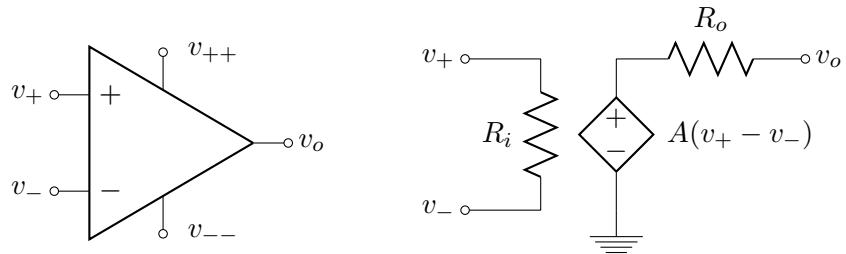


Figure 2.5: Left: triangular component diagram of an **operational amplifier**. The  $v_{++}$  and  $v_{--}$  connectors indicate where an external voltage source can be connected to the amplifier. These connectors will usually be omitted. Right: model for an operational amplifier including input resistance  $R_i$ , output resistance  $R_o$ , and open loop gain  $A$ . The diamond shaped component is a dependent voltage source. This model is usually only useful when the operational amplifier is in a negative feedback circuit.

leading to the following relationships between  $y$ ,  $v_C$  and  $v_L$ ,

$$y = RCD(v_C), \quad Rv_L = LD(y).$$

Kirchhoff's voltage law gives  $x = y + v_C + v_L$  and by differentiating both sides

$$D(x) = D(y) + D(v_C) + D(v_L).$$

Substituting the equations relating  $y$ ,  $v_C$  and  $v_L$  leads to

$$RCD(x) = y + RCD(y) + LCD^2(y). \quad (2.1.1)$$

We can similarly find equations relating the input voltage with  $v_C$  and  $v_L$ .

## 2.2 Active electrical circuits

Unlike passive electrical circuits, an **active electrical circuit** requires a source of power external to the input signal. Active circuits can be modelled

and constructed using **operational amplifiers** as depicted in Figure 2.5. The left hand side of Figure 2.5 shows a triangular circuit diagram for an operational amplifier, and the right hand side of Figure 2.5 shows a circuit that can be used to model the behaviour of the amplifier. The  $v_{++}$  and  $v_{--}$  connectors indicate where an external voltage source can be connected to the amplifier. These connectors will usually be omitted. The diamond shaped component is a dependent voltage source with voltage  $A(v_+ - v_-)$  that depends on the difference between the voltage at the **non-inverting input**  $v_+$  and the voltage at the **inverting input**  $v_-$ . The dimensionless constant  $A$  is called the **open loop gain**. Most operational amplifiers have large open loop gain  $A$ , large **input resistance**  $R_i$  and small **output resistance**  $R_o$ . As we will see, it can be convenient to consider the behaviour as  $A \rightarrow \infty$ ,  $R_i \rightarrow \infty$  and  $R_o \rightarrow 0$ , resulting in an **ideal operational amplifier**.

As an example, an operational amplifier configured as a **multiplier** is depicted in Figure 2.6. This circuit is an example of an operational amplifier configured with **negative feedback**, meaning that the output of the amplifier is connected (in this case by a resistor) to the inverting input  $v_-$ . The horizontal wire at the bottom of the plot is considered to be ground (zero volts) and is connected to the negative terminal of the dependent voltage source of the operational amplifier depicted in Figure 2.5. An equivalent circuit for the multiplier using the model in Figure 2.5 is shown in Figure 2.7. Solving this circuit (Exercise 2.1) yields the following relationship between the input voltage signal  $x$  and the output voltage signal  $y$ ,

$$y = \frac{R_i(R_o - AR_2)}{R_i(R_2 + R_o) + R_1(R_2 + R_i + AR_i + R_o)} x. \quad (2.2.1)$$

For an ideal operational amplifier we let  $A \rightarrow \infty$ ,  $R_i \rightarrow \infty$  and  $R_o \rightarrow 0$ . In this case terms involving the product  $AR_i$  dominate and we are left with the simpler equation

$$y = -\frac{R_2}{R_1} x. \quad (2.2.2)$$

Thus, assuming an ideal operational amplifier, the circuit acts as a multiplier with constant  $-\frac{R_2}{R_1}$ .

The equation relating  $x$  and  $y$  is much simpler for the ideal operational amplifier. Fortunately this equation can be obtained directly using the following two rules:

1. the voltage at the inverting and non-inverting inputs are equal,
2. no current flows through the inverting and non-inverting inputs.

These rules are only useful for analysing circuits with negative feedback. Let us now rederive (2.2.2) using these rules. Because the non-inverting input is connected to ground, the voltage at the inverting input is zero. So, the voltage over resistor  $R_2$  is  $y = R_2 i$ . Because no current flows through the

inverting input the current through  $R_1$  is also  $i$  and  $x = -R_1i$ . Combining these results, the input voltage  $x$  and the output voltage  $y$  are related by

$$y = -\frac{R_2}{R_1}x.$$

In Test 2 the inverting amplifier circuit is constructed and the relationship above is tested using a computer soundcard.

We now consider another circuit consisting of an operational amplifier, two resistors and two capacitors depicted in Figure 2.8. Assuming an ideal operational amplifier, the voltage at the inverting terminal is zero because the non-inverting terminal is connected to ground. Thus, the voltage over capacitor  $C_2$  and resistor  $R_2$  is equal to  $y$  and, by Kirchoff's current law,

$$i = \frac{y}{R_2} + C_2D(y).$$

Similarly, since no current flows through the inverting terminal,

$$i = -\frac{x}{R_1} - C_1D(x).$$

Combining these equations yields

$$-\frac{x}{R_1} - C_1D(x) = \frac{y}{R_2} + C_2D(y). \quad (2.2.3)$$

Observe the similarity between this equation and that for the passive RC circuit (2.0.2) when  $R_1 = R_2$  and  $C_1 = 0$  (an open circuit). In this case

$$x = -y - R_1C_2D(y). \quad (2.2.4)$$

We call this the **active RC circuit**. This circuit is tested in Test 3.

Consider the circuit in Figure 2.9. Assuming an ideal operational amplifier, the input voltage  $x$  satisfies

$$-i = \frac{x}{R_1} + C_1D(x).$$

The voltage over the capacitor  $C_2$  is  $y - R_2i$  and so the current satisfies

$$i = C_2D(y - R_2i).$$

Combining these equations gives

$$-\frac{x}{R_1} - C_1D(x) = C_2D(y) + \frac{R_2C_2}{R_1}D(x) + R_2C_2C_1D^2(x),$$

and after rearranging,

$$D(y) = -\frac{1}{R_1C_1}x - \left(\frac{R_2}{R_1} + \frac{C_1}{C_2}\right)D(x) - R_2C_1D^2(x).$$

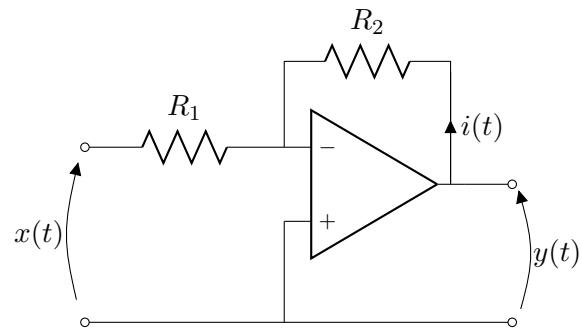


Figure 2.6: Inverting amplifier

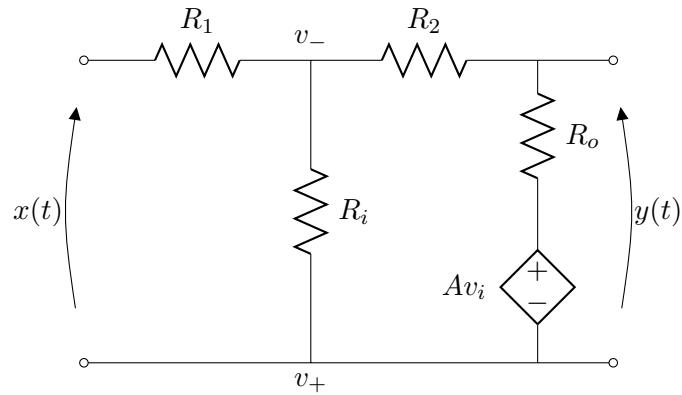


Figure 2.7: An equivalent circuit for the inverting amplifier from Figure 2.6 using the model for an operational amplifier in Figure 2.5. The symbol  $v_i = v_+ - v_-$  is the voltage over resistor  $R_i$ .

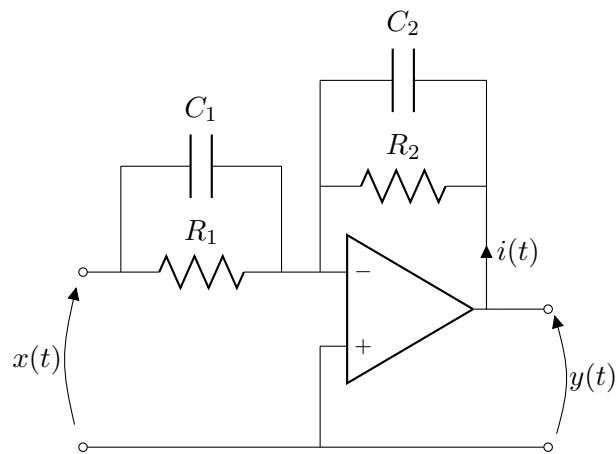


Figure 2.8: Operational amplifier configured with two capacitors and two resistors.

**Test 2 (Inverting amplifier)** In this test we construct the inverting amplifier circuit from Figure 2.6 with  $R_2 \approx 22\text{k}\Omega$  and  $R_1 \approx 12\text{k}\Omega$  that are accurate to within 5% of these values. The operational amplifier used is the Texas Instruments LM358P. Using a computer soundcard (an approximation of) the voltage signal

$$x(t) = \frac{1}{3} \sin(2\pi f_1 t) + \frac{1}{3} \sin(2\pi f_2 t)$$

with  $f_1 = 100$  and  $f_2 = 233$  is passed through the circuit. As in previous tests, the soundcard is used to sample the input signal  $x$  and the output signal  $y$ . Approximate reconstructions of the input signal  $\tilde{x}$  and output signal  $\tilde{y}$  are given according to (1.2.4) and (1.2.2). According to (2.1.1) we expect the approximate relationship

$$\tilde{y} \approx -\frac{R_2}{R_1} \tilde{x} = -\frac{11}{6} \tilde{x}.$$

Each of  $\tilde{y}$ ,  $\tilde{x}$  and  $-\frac{11}{6} \tilde{x}$  are plotted in Figure 2.9.

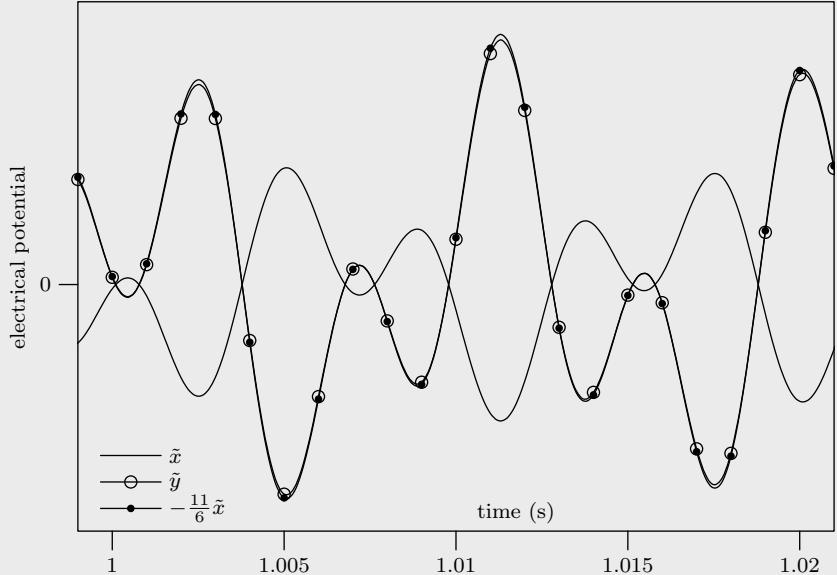


Figure 2.9: Plot of reconstructed input signal  $\tilde{x}$  (solid line), output signal  $\tilde{y}$  (solid line with circle) and hypothesised output signal  $-\frac{11}{6} \tilde{x}$  (solid line with dot).

**Test 3 (Active RC circuit)** In this test we construct the circuit from Figure 2.8 with  $R_1 \approx R_2 \approx 27\text{k}\Omega$  and  $C_2 \approx 10\text{nF}$  accurate to within 5% of these values and  $C_1 = 0$  (an open circuit). The operational amplifier used is a Texas Instruments LM358P. Using a computer soundcard (an approximation of) the voltage signal

$$x(t) = \frac{1}{3} \sin(2\pi f_1 t) + \frac{1}{3} \sin(2\pi f_2 t)$$

with  $f_1 = 500$  and  $f_2 = 1333$  is passed through the circuit. As in previous tests, the soundcard is used to sample the input signal  $x$  and the output signal  $y$  and approximate reconstructions  $\tilde{x}$  and  $\tilde{y}$  are given according to (1.2.4) and (1.2.2). According to (2.2.4) we expect the approximate relationship

$$\tilde{x} \approx -\frac{R_1}{R_2} \tilde{y} - R_1 C D(\tilde{y}) = -\tilde{y} - \frac{27}{10^5} D(\tilde{y}).$$

The derivative of the sinc function is

$$D(\text{sinc}, t) = \frac{1}{\pi t^2} (\pi t \cos(\pi t) - \sin(\pi t)), \quad (2.2.5)$$

and so,

$$D(\tilde{y}, t) = \frac{d}{dt} \left( \sum_{\ell=1}^L y_\ell \text{sinc}(Ft - \ell) \right) = F \sum_{\ell=1}^L y_\ell D(\text{sinc}, Ft - \ell). \quad (2.2.6)$$

Each of  $\tilde{y}$ ,  $\tilde{x}$  and  $-\tilde{y} - \frac{27}{10^5} D(\tilde{y})$  are plotted in Figure 2.9.

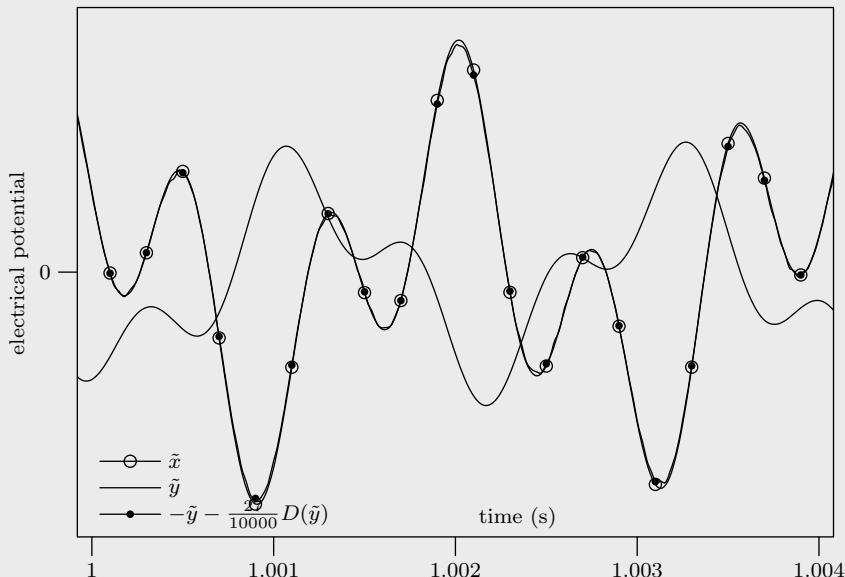


Figure 2.9: Plot of reconstructed input signal  $\tilde{x}$  (solid line with circle), output signal  $\tilde{y}$  (solid line), and hypothesised input signal  $-\tilde{y} - \frac{27}{10^5} D(\tilde{y})$  (solid line with dot).

Put

$$K_i = \frac{1}{R_1 C_2}, \quad K_p = \frac{R_2}{R_1} + \frac{C_1}{C_2}, \quad K_d = R_2 C_1$$

and now

$$D(y) = -K_i x - K_p D(x) - K_d D^2(x). \quad (2.2.7)$$

This equation models what is called a **proportional-integral-derivative controller** or **PID controller**. The coefficients  $K_i$ ,  $K_p$  and  $K_d$  are called the **integral gain**, **proportional gain**, and **derivative gain**.

The final active circuit we consider is called a **Sallen-Key** [Sallen and Key, 1955] and is depicted in Figure 2.10. Observe that the output of the amplifier is connected directly to the inverting input and is also connected to the noninverting input by a capacitor and resistor. This circuit has both negative *and* positive feedback. It is not immediately apparent that we can use the simplifying assumptions for an ideal operational amplifier with negative feedback. However, we will do so, and will find that it works in this case.

Let  $v_{R1}$ ,  $v_{R2}$ ,  $v_{C1}$ , and  $v_{C2}$  be the voltages over the components  $R_1$ ,  $R_2$ ,  $C_1$ , and  $C_2$ . Kirchoff's voltage law leads to the equations

$$x = v_{R1} + v_{R2} + v_{C2}, \quad y = v_{C1} + v_{R2} + v_{C2}.$$

The voltage at the inverting and noninverting terminals is  $y$  and so the voltage over the capacitor  $C_2$  is  $y$ , that is,  $y = v_{C2}$ . Using this, the equations above simplify to

$$x = v_{R1} + v_{R2} + y, \quad v_{C1} = -v_{R2}.$$

The current  $i_2$  through capacitor  $C_2$  satisfies  $i_2 = C_2 D(v_{C2}) = C_2 D(y)$ . Because no current flows into the inverting terminal of the amplifier the current through  $R_2$  is also  $i_2$  and so  $v_{R2} = R_2 i_2 = R_2 C_2 D(y)$ . Substituting this into the equations above gives

$$x = v_{R1} + R_2 C_2 D(y) + y, \quad v_{C1} = -R_2 C_2 D(y). \quad (2.2.8)$$

Kirchoff's current law asserts that  $i + i_1 = i_2$ . The current  $i$  through capacitor  $C_1$  satisfies  $i = C_1 D(v_{C1}) = -R_2 C_1 C_2 D^2(y)$  and the current through resistor  $R_1$  satisfies

$$v_{R1} = R_1 i_1 = R_1 (i_2 - i) = R_1 C_2 D(y) + R_1 R_2 C_1 C_2 D^2(y).$$

Substituting this into the equation on the left of (2.2.8) gives

$$x = y + C_2(R_1 + R_2)D(y) + R_1 R_2 C_1 C_2 D^2(y). \quad (2.2.9)$$

The Sallen-Key will be useful when we consider the design of analogue electrical filters in Section 5.2.

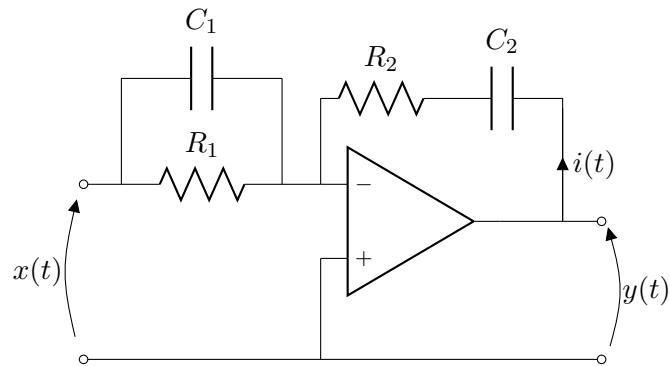


Figure 2.9: Operational amplifier implementing a **proportional-integral-derivative controller**.

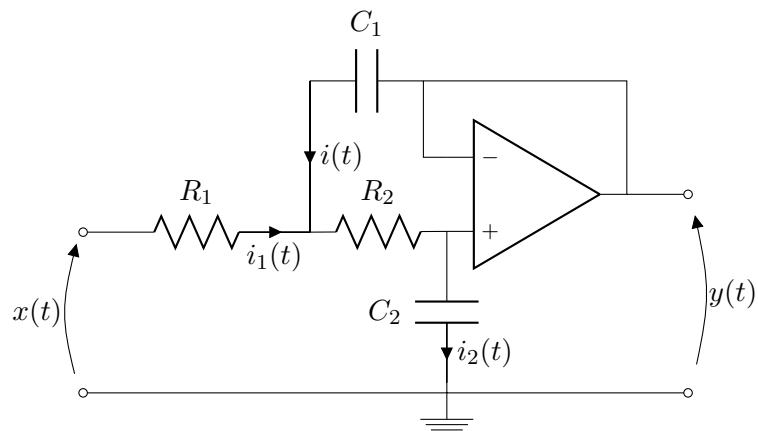


Figure 2.10: Operational amplifier implementing a **Sallen-Key**.

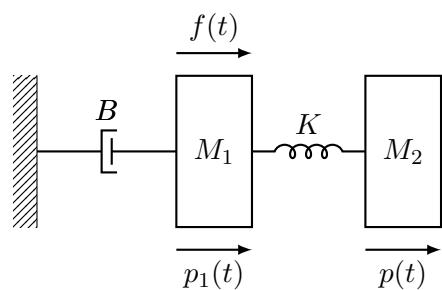


Figure 2.11: Two masses, a spring and a damper

### 2.3 Masses, springs, and dampers

A mechanical mass-spring-damper system was described in Section 2 and Figure 2.2. We now consider another mechanical system involving a different configuration of masses, a spring and a damper depicted in Figure 2.11. A mass  $M_1$  is connected to a wall by a damper with constant  $B$ , and to another mass  $M_2$  by a spring with constant  $K$ . A force represented by the signal  $f$  is applied to the first mass. We will derive a differential equation relating  $f$  with the position  $p$  of the second mass. Assume that the spring applies no force (is in equilibrium) when the masses are distance  $d$  apart. The forces due to the spring satisfy

$$f_{s1} = -f_{s2} = K(p - p_1 - d)$$

where  $f_{s1}$  and  $f_{s2}$  are signals representing the force due to the spring on mass  $M_1$  and  $M_2$  respectively. It is convenient to define the signal  $g(t) = p_1(t) + d$  so that forces due to spring satisfy the simpler equation

$$f_{s1} = -f_{s2} = K(p - g).$$

The only force applied to  $M_2$  is by the spring and so, by Newton's law, the acceleration of  $M_2$  satisfies

$$M_2 D^2(p) = f_{s2}.$$

Substituting this into the previous equation gives a differential equation relating  $g$  and  $p$ ,

$$Kg = Kp + M_2 D^2(p). \quad (2.3.1)$$

The force applied by the damper on mass  $M_1$  is given by the signal

$$f_d = -BD(p_1) = -BD(g)$$

where the replacement of  $p_1$  by  $g$  is justified because differentiation will remove the constant  $d$ . The cumulative force on  $M_1$  is given by the signal

$$\begin{aligned} f_1 &= f + f_d + f_{s1} \\ &= f - Kg + Kp - BD(g), \end{aligned} \quad (2.3.2)$$

and by Newton's law the acceleration of  $M_1$  satisfies

$$M_1 D^2(p_1) = M_1 D^2(g) = f_1.$$

Substituting this into (2.3.2) and using (2.3.1) we obtain a fourth order differential equation relating  $p$  and  $f$ ,

$$f = BD(p) + (M_1 + M_2)D^2(p) + \frac{BM_2}{K}D^3(p) + \frac{M_1 M_2}{K}D^4(p). \quad (2.3.3)$$

Figure 2.12: Solution of the system describing two masses with a spring and damper where  $B = K = 1$  and  $M_1 = M_2 = \frac{1}{2}$  and the position of the second mass is  $p(t) = e^{-t^2}$ .

Given the position of the second mass  $p$  we can readily solve for the corresponding force  $f$  and position of the first mass  $p$ . For example, if the constants  $B = K = 1$  and  $M_1 = M_2 = \frac{1}{2}$  and  $d = \frac{5}{2}$ , and if the position of the second mass satisfies

$$p(t) = e^{-t^2}$$

then, by application of (2.3.3) and (2.3.1),

$$f(t) = e^{-t^2}(1 + 4t - 8t^2 - 4t^3 + 4t^4), \quad \text{and} \quad p_1(t) = 2e^{-t^2}t^2 - \frac{5}{2}.$$

This solution is plotted in Figure 2.12.

## 2.4 Direct current motors

Direct current (DC) motors convert electrical energy, in the form of a voltage, into rotary kinetic energy [Nise, 2007, page 76]. We derive a differential equation relating the input voltage  $v$  to the angular position of the motor  $\theta$ . Figure 2.13 depicts the components of a DC motor.

The voltages over the resistor and inductor satisfy

$$v_R = Ri, \quad v_L = LD(i),$$

and the motion of the motor induces a voltage called the **back electromotive force** (EMF),

$$v_b = K_bD(\theta)$$

that we model as being proportional to the angular velocity of the motor. The input voltage now satisfies

$$v = v_R + v_L + v_b = Ri + LD(i) + K_b D(\theta).$$

The torque  $\tau$  applied by the motor is modelled as being proportional to the current  $i$ ,

$$\tau = K_\tau i.$$

A load with inertia  $J$  is attached to the motor. Two forces are assumed to act on the load, the torque  $\tau$  applied by the current, and a torque  $\tau_d = -BD(\theta)$  modelling a damper that acts proportionally against the angular velocity of the motor. By Newton's law, the angular acceleration of the load satisfies

$$JD^2(\theta) = \tau + \tau_d = K_\tau i - BD(\theta).$$

Combining these equations we obtain the 3rd order differential equation

$$v = \left( \frac{RB}{K_\tau} + K_b \right) D(\theta) + \frac{RJ + LB}{K_\tau} D^2(\theta) + \frac{LJ}{K_\tau} D^3(\theta)$$

relating voltage and motor position. In many DC motors the inductance  $L$  is small and can be ignored, leaving the simpler second order equation

$$v = \left( \frac{RB}{K_\tau} + K_b \right) D(\theta) + \frac{RJ}{K_\tau} D^2(\theta). \quad (2.4.1)$$

Given the position signal  $\theta$  we can find the corresponding voltage signal  $v$ . For example, put the constants  $K_b = K_\tau = B = R = J = 1$  and assume that

$$\theta(t) = 2\pi(1 + \text{erf}(t))$$

where  $\text{erf}(t) = \frac{2}{\sqrt{\pi}} \int_{-\infty}^t e^{-\tau^2} d\tau$  is the **error function**. The corresponding angular velocity  $D(\theta)$  and voltage  $v$  satisfy

$$D(\theta, t) = 4\sqrt{\pi}e^{-t^2}, \quad v(t) = 8\sqrt{\pi}e^{-t^2}(1 - t).$$

These signals are depicted in Figure 2.14. This voltage signal is sufficient to make the motor perform two revolutions and then come to rest.

## 2.5 Exercises

- 2.1. Analyse the inverting amplifier circuit in Figure 2.7 to obtain the relationship between input voltage  $x$  and output voltage  $y$  given by (2.2.1). You may wish to use a symbolic programming language (for example Sage, Mathematica, or Maple). **Solution:** We provide two solutions. Let  $v_i$ ,  $v_o$ ,  $v_1$  and  $v_2$  be the voltages over the input resistor  $R_i$ , the output resistor  $R_o$ , and

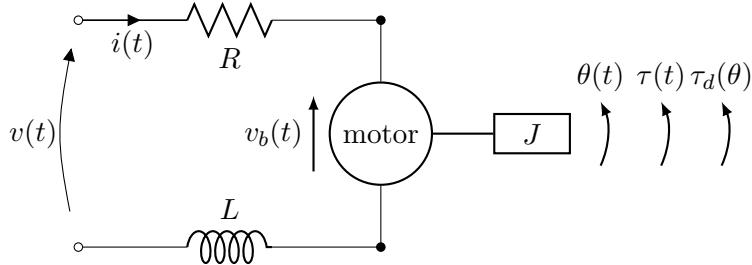


Figure 2.13: Diagram for a rotary direct current (DC) motor

Figure 2.14: Voltage and corresponding angle for a DC motor with constants  $K_b = K_\tau = B = R = J = 1$ .

resistors  $R_1$  and  $R_2$  respectively. Observe that  $v_+ - v_i = v_i$  and so the voltage over the dependent source is  $Av_i$ . The voltages satisfy,

$$\begin{aligned} x &= v_1 - v_i \\ y &= -v_i - v_2 \\ y &= v_o + Av_i \end{aligned}$$

The currents into the 3 way connection between  $R_i$ ,  $R_1$  and  $R_2$  sum to zero, and so

$$\frac{v_1}{R_1} + \frac{v_i}{R_i} = \frac{v_2}{R_2}$$

by Ohm's law, the direction of current moving from positive to negative voltage. Finally the currents through  $R_o$  and  $R_2$  are the same, and so

$$\frac{v_o}{R_o} = \frac{v_2}{R_2}.$$

We now have 5 linearly independent equations for the six unknowns  $v_1, v_2, v_o, v_i, x, y$ . We can use these to find an equation that describes  $y$  in terms of  $x$ . The Mathematica command

```
Simplify[Solve[{x == v1 - vi,
y == vo + A*vi,
y == -vi - v2,
v1/r1 + vi/ri == v2/r2,
vo/ro == v2/r2,
r1 > 0, r2 > 0, ro > 0, ri > 0, A > 0},
{y, vi, vo, v2, v1}, Reals]]
```

readily obtains

$$y = \frac{R_i(R_o - AR_2)}{R_i(R_2 + R_o) + R_1(R_2 + R_i + AR_i + R_o)} x.$$

The second solution is thanks to Badri Vellambi. Consider the operational amplifier circuit with feedback presented in Fig. 2.15. Suppose that the voltage signal fed into the circuit is  $x(t)$  and the voltage signal measured at the output of the opamp is  $y(t)$ .

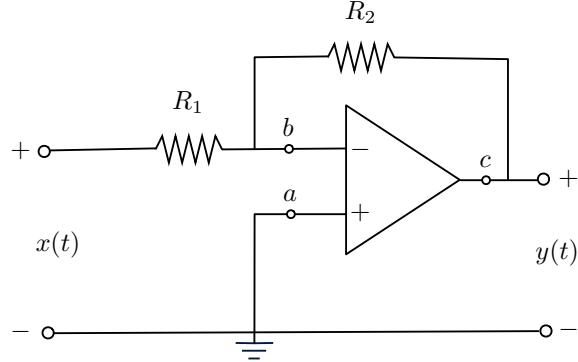


Figure 2.15: The circuit

To simplify the circuit, one has to use the model for the opamp given in Fig. 2.16 which involves the voltage-controlled voltage-source (VCVS) at the output side (indicated in green). While replacing the operational amplifier with its model, it must be noted that the positive terminal of the operational amplifier is connected to the ground.

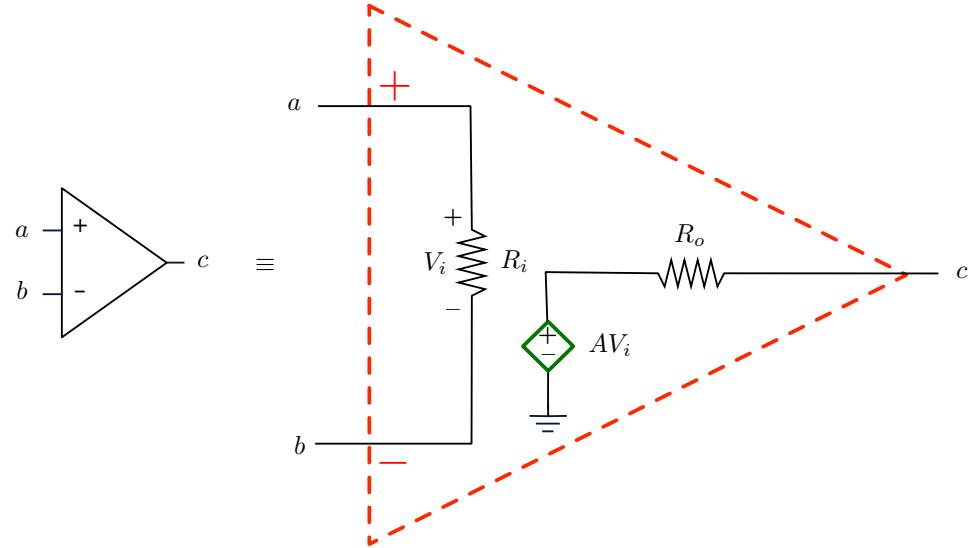


Figure 2.16: The model for an operational amplifier

Upon replacement, we obtain the following equivalent circuit. Again notice that since the positive terminal of the opamp was connected to the ground, the voltage output by the VCVS is  $AV_i$  where  $V_i$  is the voltage between the ground and the top of the resistance  $R_i$ , and is measured against the flow of the current  $i - i_1$  as

is indicated in the figure.

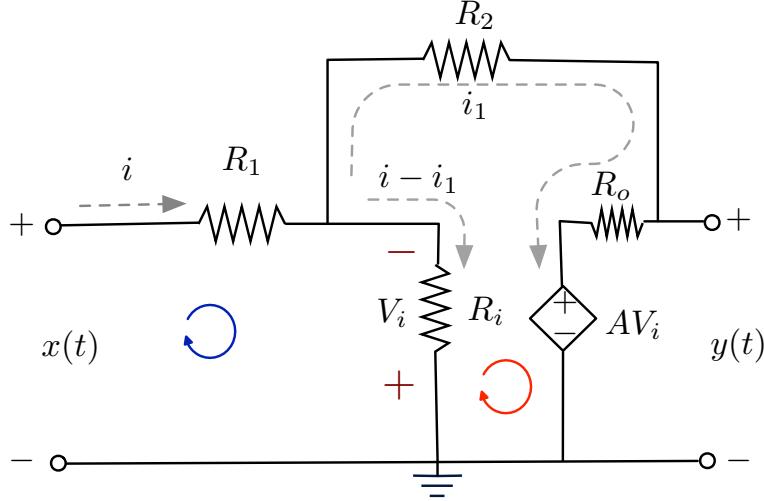


Figure 2.17: The operational amplifier circuit with the model

Applying Kirchoff's law to the outer loop indicated in blue in Fig. 2.17, we obtain the following equation.

$$x(t) = iR_1 + (i - i_1)R_i = i(R_1 + R_i) - i_1R_i \quad (2.5.1)$$

Note that by definition, the voltage  $V_i$  that controls the VCVS is the voltage across  $R_i$  measured against the indicated direction of the current  $i - i_1$ , and is given by

$$V_i = -(i - i_1)R_i. \quad (2.5.2)$$

Next, writing out the Kirchoff's law for the inner loop indicated in red, we obtain the following.

$$0 = i_1R_2 + i_2R_0 + AV_i - (i - i_1)R_i \quad (2.5.3)$$

Substituting  $V_i$  in the above equation with the RHS of (2.5.2), we obtain the following.

$$0 = i_1(R_2 + R_0) - A(i - i_1)R_i - (i - i_1)R_i \quad (2.5.4)$$

$$= -i(1 + A)R_i + i_1((1 + A)R_i + R_0 + R_2) \quad (2.5.5)$$

Combining (2.5.5) and (2.5.1), we obtain the following linear system of equations governing the electrical circuit.

$$\begin{bmatrix} R_1 + R_i & -R_i \\ -(1 + A)R_i & (1 + A)R_i + R_0 + R_2 \end{bmatrix} \begin{bmatrix} i \\ i_1 \end{bmatrix} = \begin{bmatrix} x(t) \\ 0 \end{bmatrix} \quad (2.5.6)$$

Solving the above linear system, we identify the current in the different branches to be

$$\begin{bmatrix} i \\ i_1 \end{bmatrix} = x(t) \begin{bmatrix} \frac{(1+A)R_i+R_0+R_2}{(1+A)R_iR_1+R_0R_1+R_2R_1+R_0R_i+R_2R_i} \\ \frac{(1+A)R_i}{(1+A)R_iR_1+R_0R_1+R_2R_1+R_0R_i+R_2R_i} \end{bmatrix}. \quad (2.5.7)$$

Lastly, notice that

$$y(t) = i_1R_0 + AV_i \quad (2.5.8)$$

$$= i_1R_0 - (i - i_1)R_i. \quad (2.5.9)$$

Substituting the solutions for  $i$  and  $i_1$  in terms of  $x(t)$ , we obtain the following.

$$y(t) = \left( \frac{R_i R_0 - R_2 R_i A}{(1 + A) R_i R_1 + R_0 R_1 + R_2 R_1 + R_0 R_i + R_2 R_i} \right) x(t) \quad (2.5.10)$$

## Chapter 3

# Linear time invariant systems

In the previous section we derived differential equations that model mechanical, electrical, and electro-mechanical systems. The equations themselves often do not provide as much information about these system as we require. For example, we were able to find a signal  $p$  representing the position of the mass-spring-damper in Figure 2.2 given a particular force signal  $f$  is applied to the mass. However, it is not immediately obvious how to find the force signal  $f$  given a particular position signal  $p$ . We will be able to solve this problem and, more generally, to describe properties of systems modelled by linear differential equations with constant coefficient, if we make the added assumptions that the systems are **linear** and **time invariant**. We study linear time invariant systems in this chapter. Throughout this chapter  $H$  will denote a linear time invariant system.

### 3.1 Convolution, regular systems and the delta “function”

A large number of linear time invariant systems can be represented by a signal called the **impulse response**. The impulse response of a system  $H$  is a signal  $h$  such that

$$H(x, t) = \int_{-\infty}^{\infty} h(\tau)x(t - \tau)d\tau,$$

that is, the response of  $H$  to input signal  $x$  can be represented as an integral equation involving  $x$  and the impulse response  $h$ . The integral is called a **convolution** and appears so often that a special notation is used for it. We write  $h * x$  to indicate the signal that results from convolution of signals  $h$  and  $x$ , that is,  $h * x$  is the signal satisfying

$$h * x = \int_{-\infty}^{\infty} h(\tau)x(t - \tau)d\tau.$$

Those systems that have an impulse response we call **regular systems**<sup>1</sup>. Observe that regular systems are linear because

$$\begin{aligned}
 H(ax + by) &= h * (ax + by) \\
 &= \int_{-\infty}^{\infty} h(\tau)(ax(t - \tau) + by(t - \tau))d\tau \\
 &= a \int_{-\infty}^{\infty} h(\tau)x(t - \tau)d\tau + b \int_{-\infty}^{\infty} h(\tau)y(t - \tau)d\tau \\
 &= a(h * x) + b(h * y) \\
 &= aH(x) + bH(y).
 \end{aligned} \tag{3.1.1}$$

The above equations show that convolution commutes with scalar multiplication and distributes with addition, that is,

$$h * (ax + by) = a(h * x) + b(h * y).$$

Regular systems are also time invariant because

$$\begin{aligned}
 T_{\kappa}(H(x)) &= T_{\kappa}(h * x) \\
 &= \int_{-\infty}^{\infty} h(\tau)x(t - \kappa - \tau)d\tau \\
 &= \int_{-\infty}^{\infty} h(\tau)T_{\kappa}(x, t - \tau)d\tau \\
 &= h * T_{\kappa}(x) \\
 &= H(T_{\kappa}(x)).
 \end{aligned}$$

We can define the impulse response of a regular system  $H$  in the following way. First define the signal

$$p_{\gamma}(t) = \begin{cases} \gamma, & 0 < t \leq \frac{1}{\gamma} \\ 0, & \text{otherwise,} \end{cases}$$

that is, a rectangular shaped pulse of height  $\gamma$  and width  $\frac{1}{\gamma}$ . The signal  $p_{\gamma}$  is plotted in Figure 3.1 for  $\gamma = \frac{1}{2}, 1, 2, 5$ . As  $\gamma$  increases the pulse gets thinner and higher so as to keep the area under  $p_{\gamma}$  equal to one. Consider the response of the regular system  $H$  to the signal  $p_{\gamma}$ ,

$$\begin{aligned}
 H(p_{\gamma}) &= h * p_{\gamma} \\
 &= \int_{-\infty}^{\infty} h(\tau)p_{\gamma}(t - \tau)d\tau \\
 &= \gamma \int_{t-1/\gamma}^t h(\tau)d\tau.
 \end{aligned}$$

---

<sup>1</sup>The name **regular system** is motivated by the term **regular distribution** [Zemanian, 1965]

Taking limits as  $\gamma \rightarrow \infty$ ,

$$\lim_{\gamma \rightarrow \infty} H(p_\gamma) = \lim_{\gamma \rightarrow \infty} \gamma \int_{t-1/\gamma}^t h(\tau) d\tau = h(t) \text{ a.e.}$$

Thus, we define the impulse response of a regular system  $H$  as the limit

$$h = \lim_{\gamma \rightarrow \infty} H(p_\gamma). \quad (3.1.2)$$

The limit exists when  $H$  is regular. If this limit does not exist, the system is not regular and does not have an impulse response.

As an example, consider the integrator system

$$I_\infty(x) = \int_{-\infty}^t x(\tau) d\tau \quad (3.1.3)$$

described in Section 1.3. This system's response to  $p_\gamma$  is

$$I_\infty(p_\gamma, t) = \int_{-\infty}^t p_\gamma(\tau) d\tau = \begin{cases} 0, & t \leq 0 \\ \gamma t, & 0 < t \leq \frac{1}{\gamma} \\ 1, & t > \frac{1}{\gamma} \end{cases}$$

The response is plotted in Figure 3.1. Taking the limit as  $\gamma \rightarrow \infty$  we find that the impulse response of the integrator is the step function

$$u(t) = \lim_{\gamma \rightarrow \infty} H(p_\gamma) = \begin{cases} 0 & t \leq 0 \\ 1 & t > 0. \end{cases} \quad \text{a.e.} \quad (3.1.4)$$

Some important systems do not have an impulse response. For example, the identity system  $T_0$  does not because

$$\lim_{\gamma \rightarrow \infty} T_0(p_\gamma) = \lim_{\gamma \rightarrow \infty} p_\gamma$$

does not exist. Similarly, all the time shifters  $T_\tau$  do not have impulse responses. However, it can be notationally useful to pretend that  $T_0$  *does* have an impulse response and we denote it by the symbol  $\delta$  called the **delta function**. The idea is to assign  $\delta$  the property

$$\int_{-\infty}^{\infty} x(t)\delta(t) dt = x(0)$$

so that convolution of  $x$  and  $\delta$  satisfies

$$\delta * x = \int_{-\infty}^{\infty} \delta(\tau)x(t-\tau) d\tau = x(t) = T_0(x).$$

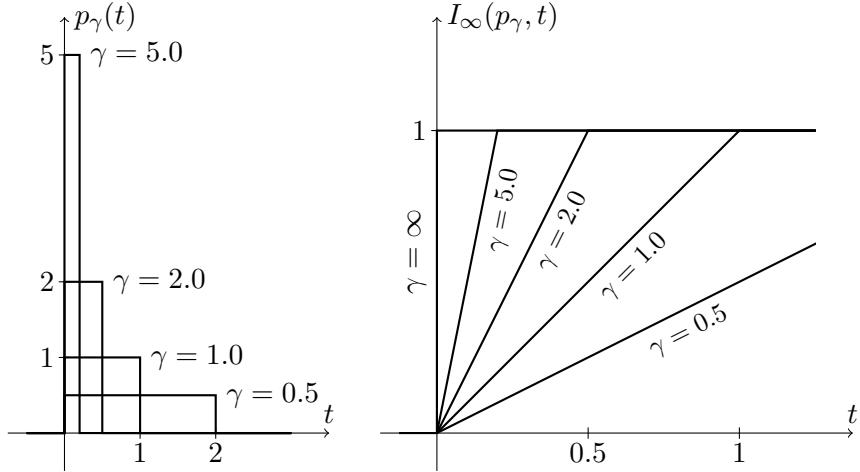


Figure 3.1: The rectangular shaped pulse  $p_\gamma$  for  $\gamma = 0.5, 1, 2, 5$  and the response of the integrator (3.1.3) to  $p_\gamma$  for  $\gamma = 0.5, 1, 2, 5, \infty$ .

We now treat  $\delta$  as if it were a signal. So  $\delta(t - \tau)$  will represent the impulse response of the time shifter  $T_\tau$  because

$$\begin{aligned} T_\tau(x) &= \delta(t - \tau) * x \\ &= \int_{-\infty}^{\infty} \delta(\kappa - \tau)x(t - \kappa)d\kappa \\ &= \int_{-\infty}^{\infty} \delta(k)x(t - \tau - k)dk \quad (\text{change variable } k = \kappa - \tau) \\ &= x(t - \tau). \end{aligned}$$

For  $a \in \mathbb{R}$  it is common to plot  $a\delta(t - \tau)$  using an arrow of height  $a$  at  $t = \tau$  as indicated in Figure 3.2. It is important to realise that  $\delta$  is not actually a signal. It is not a function. However, it can be convenient to treat  $\delta$  as if it were a function. The manipulations in the last set of equations, such as the change of variables, are not formally justified, but they do lead to the desired result  $T_\tau(x) = x(t - \tau)$  in this case. In general, there is no guarantee that mechanical mathematical manipulations involving  $\delta$  will lead to sensible results.

The only other non regular systems that we have use of are differentiators  $D^k$ , and it is convenient to define a similar notation for pretending that these systems have an impulse response. In this case we use the symbol  $\delta^k$  and assign it the property

$$\int_{-\infty}^{\infty} x(t)\delta^k(t)dt = D^k(x, 0),$$

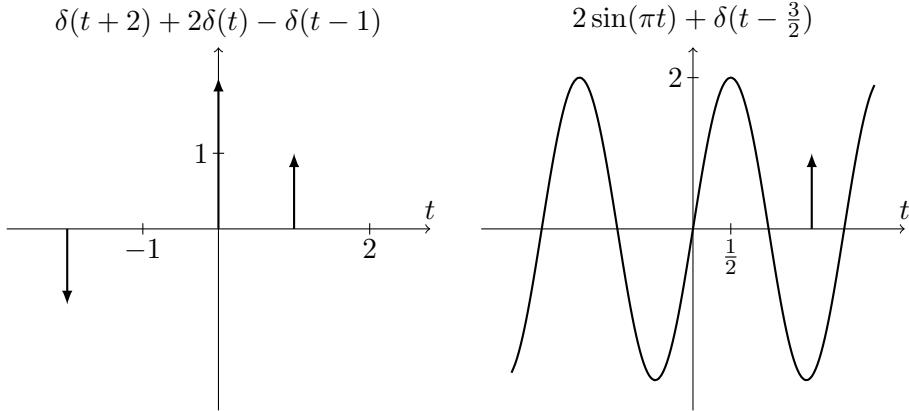


Figure 3.2: Plot of the “signal”  $\delta(t + 2) + 2\delta(t) - \delta(t - 1)$  (left) and the “signal”  $2 \sin(\pi t) + \delta(t - \frac{3}{2})$  (right).

so that convolution of  $x$  and  $\delta$  is

$$\delta^k * x = \int_{-\infty}^{\infty} \delta^k(\tau) x(t - \tau) d\tau = D^k(x, t).$$

As with the delta function the symbol  $\delta^k$  must be treated with care. This notation can be useful, but purely formal manipulations with  $\delta^k$  may not lead to sensible results in general.

The impulse response  $h$  immediately yields some properties of the corresponding system  $H$ . For example, if  $h(t) = 0$  for all  $t < 0$ , then  $H$  is causal because

$$H(x) = h * x = \int_{-\infty}^{\infty} h(\tau) x(t - \tau) d\tau = \int_0^{\infty} h(\tau) x(t - \tau) d\tau$$

only depends on values of the input signal  $x$  at times less than  $t$ , i.e., only times  $t - \tau$  with  $\tau > 0$ . The system  $H$  is stable if and only if  $h$  is absolutely integrable (Exercise 3.3).

Another important signal is the **step response** of a system that is defined as the response of the system to the step function  $u(t)$ . For example, the step response of the time shifter  $T_\tau$  is the time shifted step function  $T_\tau(u, t) = u(t - \tau)$ . The step response of the integrator  $I_\infty$  is

$$I_\infty(u) = \int_{-\infty}^t u(\tau) d\tau = \begin{cases} \int_0^t d\tau = t & t > 0 \\ 0 & t \leq 0. \end{cases}$$

This signal is often called the **ramp function**. Not all systems have a step response. For example, the system with impulse response  $u(-t)$  does not because the convolution of the step  $u(t)$  and its reflection  $u(-t)$  does not exist. If a system  $H$  has both an impulse response  $h$  and a step response

$H(u)$ , then these two signals are related. To see this, observe that the step response is

$$H(u) = h * u = \int_{-\infty}^{\infty} h(\tau)u(t - \tau)d\tau = \int_{-\infty}^t h(\tau)d\tau = I_{\infty}(h). \quad (3.1.5)$$

Thus, the step response can be obtained by applying the integrator  $I_{\infty}$  to the impulse response.

### 3.2 Properties of convolution

The convolution  $x * y$  of two signals  $x$  and  $y$  does not always exist. For example, if  $x(t) = u(t)$  and  $y(t) = 1$ , then

$$x * y = \int_{-\infty}^{\infty} x(\tau)y(t - \tau)d\tau = \int_{-\infty}^{\infty} u(\tau)d\tau = \int_0^{\infty} d\tau$$

is not finite for any  $t$ . We cannot convolve the step function  $u$  and the signal that is equal to 1 for all time. On the other hand, if  $x(t) = y(t) = u(t)$ , then

$$x * y = \int_{-\infty}^{\infty} u(\tau)u(t - \tau)d\tau = \begin{cases} \int_0^t d\tau = \tau & t > 0 \\ 0 & t \leq 0, \end{cases}$$

if finite for all  $t$ .

We have already shown in (3.1.1) that convolution commutes with scalar multiplication and is distributive with addition, that is, for signals  $x, y, w$  and complex numbers  $a, b$ ,

$$a(x * w) + b(y * w) = (ax + by) * w.$$

Convolution is commutative, that is,  $x * y = y * x$  whenever these convolutions exist. To see this, write

$$\begin{aligned} x * y &= \int_{-\infty}^{\infty} x(\tau)y(t - \tau)d\tau \\ &= \int_{-\infty}^{\infty} x(t - \kappa)y(\kappa)d\kappa \quad (\text{change variable } \kappa = t - \tau) \\ &= y * x. \end{aligned}$$

Convolution is also associative, that is, for signals  $x, y, z$ ,

$$(x * y) * z = x * (y * z). \quad (\text{see Exercise 3.2})$$

By combining the associative and commutative properties we find that the order in which the convolutions in  $x * y * z$  are performed does not matter, that is

$$x * y * z = y * z * x = z * x * y = y * x * z = x * z * y = z * y * x$$

provided that all the convolutions involved exist. More generally, the order in which any sequence of convolutions is performed does not change the final result.

### 3.3 Linear combining and composition

Let  $H_1$  and  $H_2$  be linear time invariant systems and let  $H$  be the system

$$H(x) = cH_1(x) + dH_2(x), \quad c, d \in \mathbb{C}$$

formed by a linear combination of  $H_1$  and  $H_2$ . The system  $H$  is linear because for signals  $x, y$  and complex numbers  $a, b$ ,

$$\begin{aligned} H(ax + by) &= cH_1(ax + by) + dH_2(ax + by) \\ &= acH_1(x) + bcH_1(y) + adH_2(x) + bdH_2(y) \quad (\text{linearity } H_1, H_2) \\ &= a(cH_1(x) + dH_2(x)) + b(cH_1(y) + dH_2(y)) \\ &= aH(x) + bH(y). \end{aligned}$$

The system is also time invariant because

$$\begin{aligned} H(T_\tau(x)) &= cH_1(T_\tau(x)) + dH_2(T_\tau(x)) \\ &= cT_\tau(H_1(x)) + dT_\tau(H_2(x)) \quad (\text{time-invariance } H_1, H_2) \\ &= T_\tau(cH_1(x) + dH_2(x)) \quad (\text{linearity } T_\tau) \\ &= T_\tau(H(x)). \end{aligned}$$

So, we can construct linear time invariant systems by **linearly combining** (adding and multiplying by constants) other linear time invariant systems. If  $H_1$  and  $H_2$  are regular systems this linear combining property can be expressed using their impulse responses  $h_1$  and  $h_2$ . We have

$$\begin{aligned} H(x) &= aH_1(x) + bH_2(x) \\ &= ah_1 * x + bh_2 * x \\ &= (ah_1 + bh_2) * x \quad (\text{distributivity of convolution}) \\ &= h * x, \end{aligned}$$

and so,  $H$  is a regular system with impulse response  $h = ah_1 + bh_2$ .

Another way to construct linear time invariant systems is by **composition**. Let  $H_1$  and  $H_2$  be linear time invariant systems and let

$$H(x) = H_2(H_1(x)),$$

that is,  $H$  first applies the system  $H_1$  and then applies the system  $H_2$ . The composition  $H_2(H_1(x))$  only applies to those signals  $x$  in the domain of  $H_1$  and such that the signal  $H_1(x)$  is in the domain of  $H_2$ . The system  $H$  is linear because, for signals  $x, y$  and complex numbers  $a, b$ ,

$$\begin{aligned} H(ax + by) &= H_2(H_1(ax + by)) \\ &= H_2(aH_1(x) + bH_1(y)) \quad (\text{linearity } H_1) \\ &= aH_2(H_1(x)) + bH_2(H_1(y)) \quad (\text{linearity } H_2) \\ &= aH(x) + bH(y). \end{aligned}$$

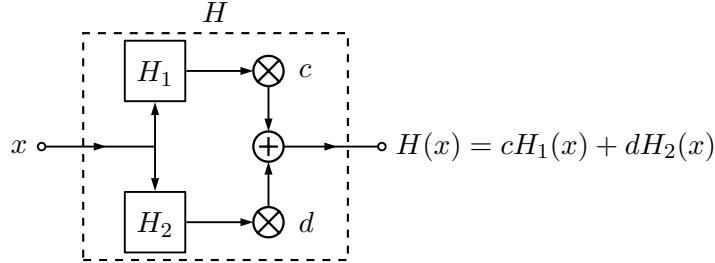


Figure 3.3: Block diagram depicting the linear combining property of linear time invariant systems. The system  $cH_1(x) + dH_2(x)$  can be expressed as a single linear time invariant system  $H(x)$ .

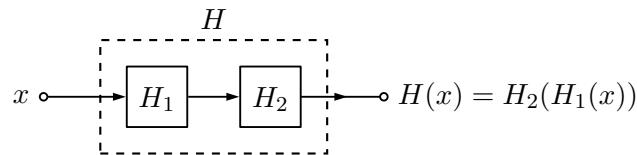


Figure 3.4: Block diagram depicting the composition property of linear time invariant systems. The system  $H_2(H_1(x))$  can be expressed as a single linear time invariant system  $H(x)$ .

The system is also time invariant because

$$\begin{aligned} H(T_\tau(x)) &= H_2(H_1(T_\tau(x))) \\ &= H_2(T_\tau(H_1(x))) \quad (\text{time-invariance } H_1) \\ &= T_\tau(H_2(H_1(x))) \quad (\text{time-invariance } H_2) \\ &= T_\tau(H(x)). \end{aligned}$$

If  $H_1$  and  $H_2$  are regular systems the composition property can be expressed using their impulse responses  $h_1$  and  $h_2$ . It follows that

$$\begin{aligned} H(x) &= H_2(H_1(x)) \\ &= h_2 * (h_1 * x) \\ &= (h_2 * h_1) * x \quad (\text{associativity of convolution}) \\ &= h * x, \end{aligned}$$

and so,  $H$  is a regular system with impulse response  $h = h_2 * h_1$ .

A wide variety of linear time invariant systems can now be constructed by linearly combining and composing simpler systems.

### 3.4 Eigenfunctions and the transfer function

Let  $s = \sigma + j\omega \in \mathbb{C}$ . Complex exponential signals of the form

$$e^{st} = e^{\sigma t} e^{j\omega t} = e^{\sigma t} (\cos(\omega t) + j \sin(\omega t))$$

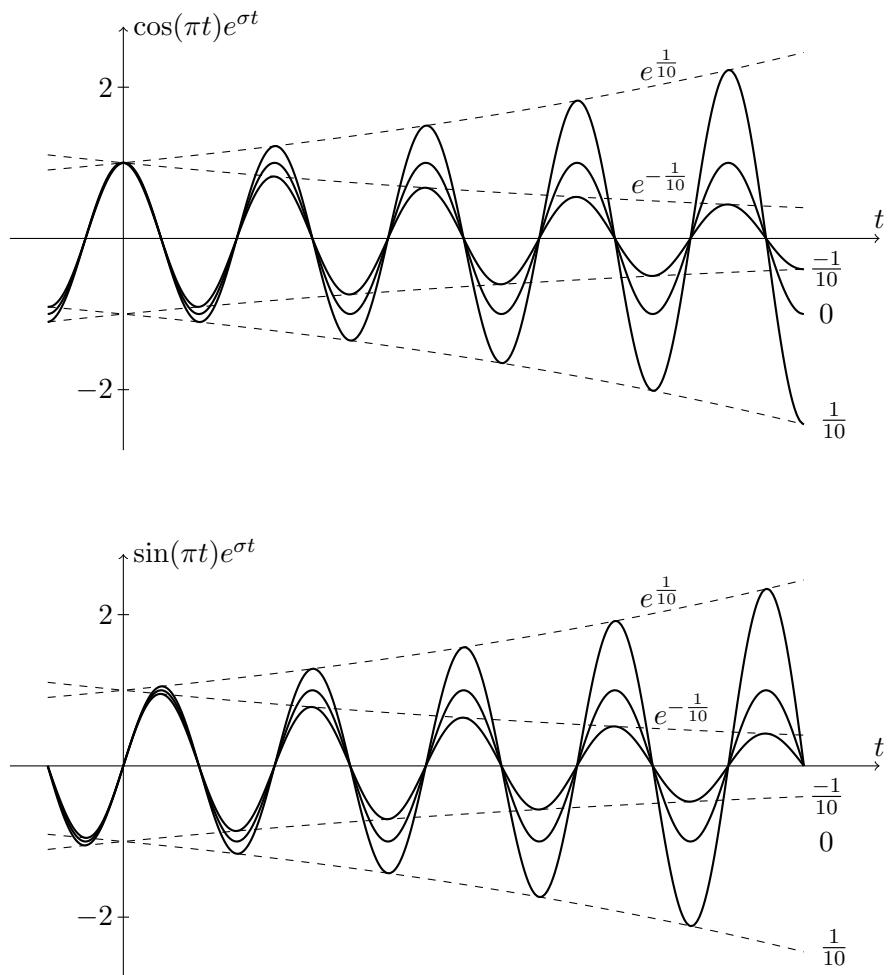


Figure 3.5: The function  $\cos(\pi t)e^{\sigma t}$  (top) and  $\sin(\pi t)e^{\sigma t}$  (bottom) for  $\sigma = -\frac{1}{10}, 0, \frac{1}{10}$ .

play an important role in the study of linear time invariant systems. The real and imaginary parts of the signal  $e^{(\sigma+j\pi)t}$  with  $\sigma = -\frac{1}{10}, 0, \frac{1}{10}$  are plotted in Figure 3.5. The signal is oscillatory when  $\omega \neq 0$ . The signal converges to zero as  $t \rightarrow \infty$  when  $\sigma < 0$  and diverges as  $t \rightarrow \infty$  when  $\sigma > 0$ .

Let  $H$  be a linear time invariant system and let  $y = H(e^{st})$  be the response of  $H$  to the exponential signal  $e^{st}$ . Consider the response of  $H$  to the time-shifted signal  $e^{s(t+\tau)}$  for  $\tau \in \mathbb{R}$ . By time-invariance

$$H(e^{s(t+\tau)}, t) = H(e^{st}, t + \tau) = y(t + \tau) \quad \text{for all } t, \tau \in \mathbb{R},$$

and by linearity

$$H(e^{s(t+\tau)}, t) = e^{s\tau} H(e^{st}, t) = e^{s\tau} y(t) \quad \text{for all } t, \tau \in \mathbb{R}.$$

Combining these equations we obtain

$$y(t + \tau) = e^{s\tau} y(t) \quad \text{for all } t, \tau \in \mathbb{R}.$$

This equation is satisfied by signals of the form  $y(t) = \lambda e^{st}$  where  $\lambda$  is a complex number. That is, the response of a linear time invariant system  $H$  to an exponential signal  $e^{st}$  is the same signal  $e^{st}$  multiplied by some constant complex number  $\lambda$ . Due to this property exponential signals are called **eigenfunctions** of linear time invariant systems. The constant  $\lambda$  does not depend on  $t$ , but it does usually depend on the complex number  $s$  and the system  $H$ . To highlight this dependence on  $H$  and  $s$  we write  $\lambda(H)(s)$  or  $\lambda(H, s)$ . Considered as a function of  $s$ ,  $\lambda(H, s)$  is called the **transfer function** of the system  $H$ . Thus, the transfer function satisfies

$$H(e^{st}) = \lambda(H, s) e^{s\tau}. \tag{3.4.1}$$

We can use these eigenfunctions to better understand the properties of systems modelled by differential equations, such as those in Section 2. As an example, consider the active electrical circuit from Figure 2.8. In the case that the resistors  $R_1 = R_2$ , and the capacitor  $C_1 = 0$  (an open circuit) the differential equation relating the input voltage  $x$  and output voltage  $y$  is

$$x = -y - R_1 C_2 D(y).$$

We call this the **active RC** circuit. To simplify notation put  $R = R_1$  and  $C = C_2$  so that  $x = -y - RCD(y)$ . Observe what occurs when  $y = ce^{st}$  is a complex exponential signal with  $c \in \mathbb{C}$ . We have

$$x = -ce^{st} - cRCse^{st} = -(1 + RCs)ce^{st} = -(1 + RCs)y,$$

and so,  $x$  is also a complex exponential signal. We immediately obtain the relationship

$$y = -\frac{1}{1 + RCs}x,$$

that holds whenever  $y$  (or equivalently  $x$ ) is of the form  $ce^{st}$  with  $c \in \mathbb{C}$ . Let  $H$  be a system that maps the input voltage  $x$  to the output voltage  $y$ , i.e.,  $H$  is a system that describes the active RC circuit. Putting  $x = e^{st}$  in the equation above, we find that

$$y = H(x) = H(e^{st}) = -\frac{1}{1 + RCs}e^{st},$$

and so, the transfer function of the system  $H$  describing the active RC circuit is

$$\lambda(H, s) = -\frac{1}{1 + RCs}. \quad (3.4.2)$$

### 3.5 The spectrum

It is often of interest to focus on the transfer function when  $s$  is purely imaginary, that is, when  $s = j\omega$ . In this case the complex exponential signal takes the form

$$e^{j\omega t} = \cos(\omega t) + j \sin(\omega t).$$

This signal is oscillatory when  $\omega \neq 0$  and does not decay or explode as  $|t| \rightarrow \infty$ . The function

$$\Lambda(H, f) = \lambda(H, j2\pi f)$$

is called the **spectrum** of the system  $H$ . It follows from (3.4.1) that the response of the system to the complex exponential signal  $e^{j2\pi ft}$  satisfies

$$H(e^{j2\pi ft}) = \lambda(H, j2\pi f)e^{j2\pi ft} = \Lambda(H, f)e^{j2\pi ft}, \quad f \in \mathbb{R}.$$

It is of interest to consider the **magnitude spectrum**  $|\Lambda(H)|$  and the **phase spectrum**  $\angle \Lambda(H)$  separately. The notation  $\angle$  denotes the **argument** (or **phase**) of a complex number. We have,

$$\Lambda(H, f) = |\Lambda(H, f)| e^{j\angle \Lambda(H, f)}$$

and correspondingly

$$H(e^{j2\pi ft}) = |\Lambda(H, f)| e^{j(2\pi ft + \angle \Lambda(H, f))}.$$

By taking real and imaginary parts we obtain the pair of real valued solutions

$$\begin{aligned} H(\cos(2\pi ft)) &= |\Lambda(H, f)| \cos(2\pi ft + \angle \Lambda(H, f)), \\ H(\sin(2\pi ft)) &= |\Lambda(H, f)| \sin(2\pi ft + \angle \Lambda(H, f)). \end{aligned} \quad (3.5.1)$$

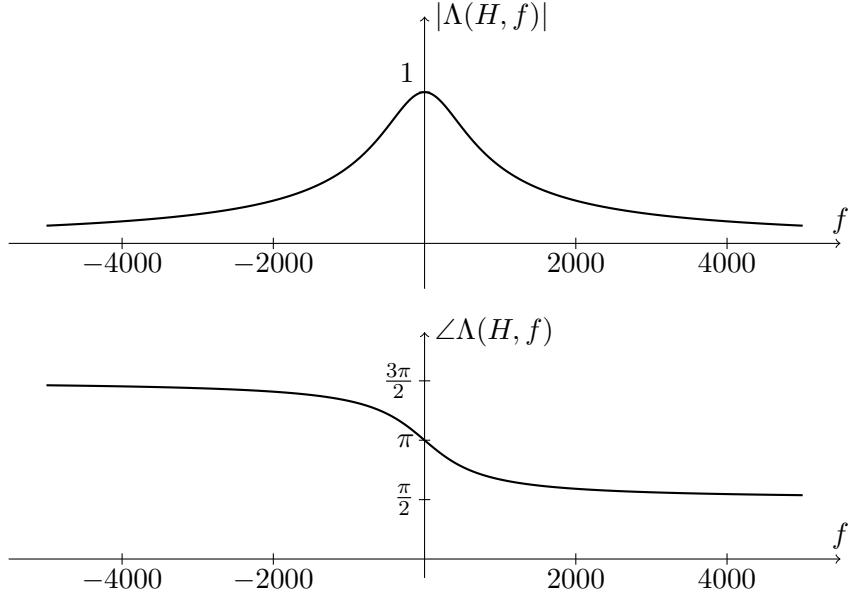


Figure 3.6: Magnitude spectrum (top) and phase spectrum (bottom) of the active RC circuit with  $R = 27 \times 10^3$  and  $C = 10 \times 10^{-9}$ .

Consider again the active RC circuit with  $H$  the system mapping the input voltage  $x$  to the output voltage  $y$ . According to (3.4.2) the spectrum of  $H$  is

$$\Lambda(H, f) = -\frac{1}{1 + 2\pi RCfj}. \quad (3.5.2)$$

The magnitude and phase spectrum is

$$|\Lambda(H, f)| = (1 + 4\pi^2 R^2 C^2 f^2)^{-\frac{1}{2}}, \quad \angle \Lambda(H, f) = \pi - \text{atan}(2\pi RCf).$$

The magnitude and phase spectrum are plotted in Figure 3.6 when  $R = 27 \times 10^3$  and  $C = 10 \times 10^{-9}$ . Observe from the plot of the magnitude spectrum that a low frequency sinusoidal signal, say 100Hz or less, input to the active RC circuit results in a sinusoidal output signal with the same frequency and approximately the same amplitude. However, a high frequency sinusoidal signal, say greater than 1000Hz, input to the circuit results in a sinusoidal output signal with the same frequency, but smaller amplitude. For this reason RC circuits are called **low pass filters**.

**Test 4 (Spectrum of the active RC circuit)** We test the hypothesis that the active RC circuit satisfies (3.5.1). To do this sinusoidal signals at varying frequencies of the form

$$x_k(t) = \sin(2\pi f_k t), \quad f_k = \lceil 110 \times 2^{k/2} \rceil, \quad k = 0, 1, \dots, 12$$

are input to the active RC circuit constructed as in Test 3 with  $R = R_1 = 27\text{k}\Omega$  and  $C = C_2 = 10\text{nF}$ . The notation  $\lceil \cdot \rceil$  denotes rounding to the nearest integer with half integers rounded up. In view of (3.5.1) the expected output signals are of the form

$$y_k(t) = |\Lambda(H, f_k)| \sin(2\pi f_k t + \angle \Lambda(H, f_k)), \quad k = 0, 1, \dots, 12.$$

This equality can also be shown directly using the differential equation for the active RC circuit.

Using the soundcard each signal  $x_k$  is played for a period of approximately 1 second and approximately  $F = 44100$  samples are obtained. On the soundcard hardware used for this test samples near the beginning and end of playback are distorted. This appears to be an unavoidable feature of the soundcard. To alleviate this we discard the first  $10^5$  samples and use only the  $L = 8820$  samples that follow (corresponding to 200ms of signal). After this process we have samples  $x_{k,1}, \dots, x_{k,L}$  and  $y_{k,1}, \dots, y_{k,L}$  of the input and output signals corresponding with the  $k$ th signal  $x_k$ . The samples are expected to take the form

$$x_{k,\ell} \approx x_k(P\ell - \tau) = \rho \sin(2\pi f_k P\ell - \theta)$$

and

$$y_{k,\ell} \approx y_k(\ell P - \tau) = |\Lambda(H, f_k)| \rho \sin(2\pi f_k P\ell - \theta + \angle \Lambda(H, f_k))$$

where  $P = \frac{1}{F}$  is the sample period, the positive real number  $\rho$  corresponds with the gain on the input and output of the soundcard, and  $\theta = 2\pi f_k \tau$  corresponds with delays caused by discarding the first  $10^5$  samples and also unavoidable delays that occur when starting soundcard playback and recording.

We will not measure the gain  $\rho$  nor the delay  $\theta$ , but will be able to test the properties of the circuit without knowledge of these. To simplify notation put  $\gamma = 2\pi f_k P$ . From the samples of the input signal  $x_{k,1}, \dots, x_{k,L}$  compute the complex number

$$\begin{aligned} A &= \frac{2j}{L} \sum_{\ell=1}^L x_{k,\ell} e^{-j\gamma\ell} \\ &\approx \frac{2j}{L} \sum_{\ell=1}^L \rho \sin(\gamma\ell - \theta) e^{-j\gamma\ell} \\ &= \alpha + \alpha^* C \end{aligned}$$

where  $\alpha = \rho e^{-j\theta}$  and  $\alpha^*$  denotes the complex conjugate of  $\alpha$  and

$$C = e^{-\gamma(L+1)} \frac{\sin(\gamma L)}{L \sin(\gamma)} \quad (\text{Excercise 3.6}).$$

Similarly, from the samples of the output signal  $y_{k,1}, \dots, y_{k,L}$  we compute the complex number

$$B = \frac{2j}{L} \sum_{\ell=1}^L y_{k,\ell} e^{-j\gamma\ell} \approx \beta + \beta^* C$$

where  $\beta = \rho e^{-j\theta} \Lambda(H, f_k) = \alpha \Lambda(H, f_k)$ . Now compute the quotient

$$Q_k = \frac{B - B^* C}{A - A^* C} \approx \frac{\beta(1 + |C|^2)}{\alpha(1 + |C|^2)} = \frac{\beta}{\alpha} = \Lambda(H, f_k).$$

Thus, we expect the quotient  $Q_k$  to be close to the spectrum of the active RC circuit evaluated at frequency  $f_k$ . We will test this hypothesis by observing the magnitude and phase of  $Q_k$  individually, that is, we will test the expected relationships

$$|Q_k| \approx |\Lambda(H, f_k)| = \frac{1}{1 + 4\pi^2 R^2 C^2 f_k^2}$$

and

$$\angle Q_k \approx \angle \Lambda(H, f_k) = \pi - \text{atan}(2\pi R C f)$$

for each  $k = 0, \dots, 12$ .

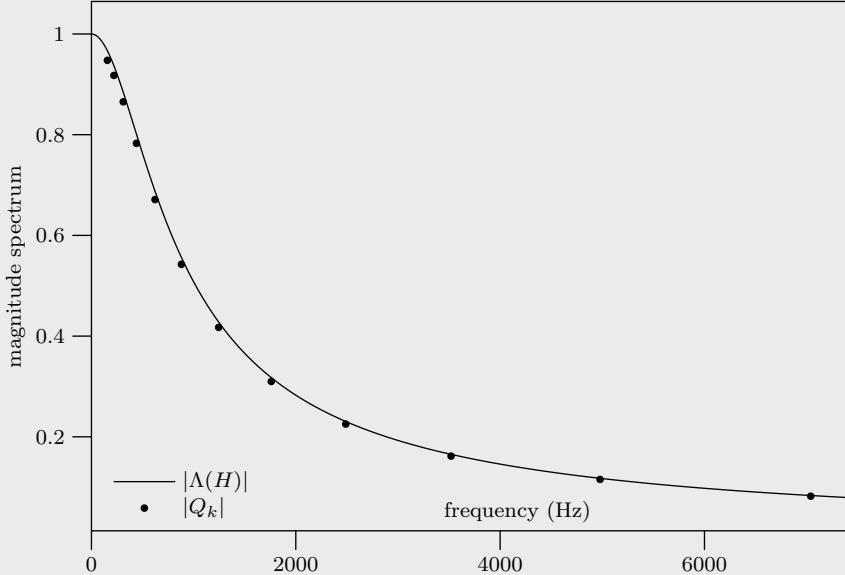


Figure 3.7: Plot of the hypothesised magnitude spectrum  $|\Lambda(H, f)|$  (solid line) and the measured magnitude spectrum  $|Q_k|$  for  $k = 0, \dots, 12$  (dots).

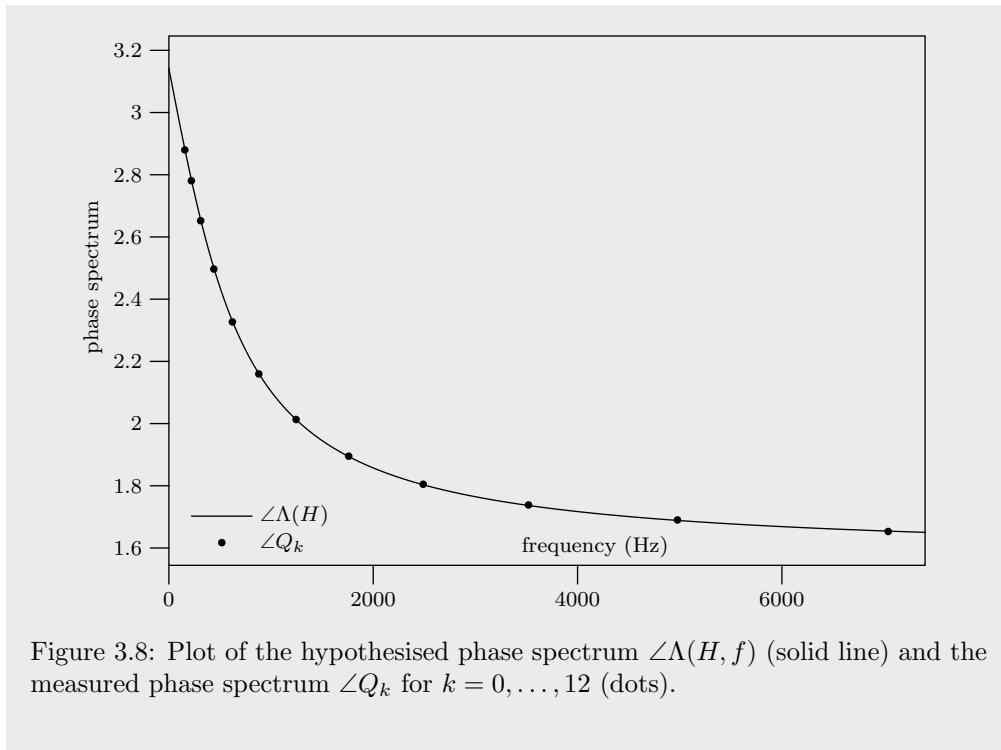


Figure 3.8: Plot of the hypothesised phase spectrum  $\angle \Lambda(H, f)$  (solid line) and the measured phase spectrum  $\angle Q_k$  for  $k = 0, \dots, 12$ .

### 3.6 Exercises

- 3.1. Show that convolution distributes with addition and commutes with scalar multiplication, that is, show that  $a(x*w)+b(y*w) = (ax+by)*w$ .

**Solution:**

$$\begin{aligned}
 a(x * w) + b(y * w) &= a \int_{-\infty}^{\infty} x(\tau) w(t - \tau) d\tau + b \int_{-\infty}^{\infty} y(\tau) w(t - \tau) d\tau \\
 &= \int_{-\infty}^{\infty} (ax(\tau) + by(\tau)) w(t - \tau) d\tau \\
 &= (ax + by) * w.
 \end{aligned}$$

- 3.2. Show that convolution is associative. That is, if  $x, y, z$  are signals then

$$x * (y * z) = (x * y) * z.$$

**Solution:**

$$\begin{aligned}
(x * y) * z &= \int_{-\infty}^{\infty} (x * y)(\tau)z(t - \tau)d\tau \\
&= \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} x(\kappa)y(\tau - \kappa)z(t - \tau)d\kappa d\tau \\
&= \int_{-\infty}^{\infty} x(\kappa) \int_{-\infty}^{\infty} y(\tau - \kappa)z(t - \tau)d\tau d\kappa \quad (\text{swap order of integration}) \\
&= \int_{-\infty}^{\infty} x(\kappa) \int_{-\infty}^{\infty} y(\nu)z(t - \kappa - \nu)d\tau d\kappa \quad (\text{change variable } \nu = \tau - \kappa) \\
&= \int_{-\infty}^{\infty} x(\kappa)(y * z)(t - \kappa)d\kappa \\
&= x * (y * z).
\end{aligned}$$

The exchange of integration order can be justified using Fubini's theorem whenever the all of the convolutions involved in  $x * (y * z) = (x * y) * z$  exist.

- 3.3. Show that a regular system is stable if and only if its impulse response is absolutely integrable. **Solution:** Let  $H$  be a regular system and  $h$  its impulse response. If  $h$  is absolutely integrable then for all signals  $x$  such that  $|x(t)| < M$  for all  $t$ ,

$$\begin{aligned}
H(x, t) &= h * x \\
&= \int_{-\infty}^{\infty} h(\tau)x(t - \tau)d\tau \\
&\leq \int_{-\infty}^{\infty} |h(\tau)x(t - \tau)|d\tau \\
&\leq \int_{-\infty}^{\infty} M|h(\tau)|d\tau \\
&= M\|h\|_1
\end{aligned}$$

for all  $t$ , and so  $H(x, t)$  is bounded. On the other hand if  $h$  is not absolutely integrable then the bounded signal

$$s(t) = \begin{cases} 1 & h(-t) > 0 \\ -1 & h(-t) \leq 0 \end{cases}$$

is such that

$$H(x, 0) = \int_{-\infty}^{\infty} h(\tau)s(-\tau)d\tau = \int_{-\infty}^{\infty} |h(\tau)|d\tau = \infty,$$

and so the signal  $H(x)$  is not bounded at  $t = 0$ .

- 3.4. Show that the system  $H(x) = \int_{-1}^1 \sin(\pi\tau)x(t + \tau)d\tau$  is linear time invariant and regular. Find and sketch the impulse response. **Solution:**

The easy way is to spot the impulse response directly. Observe that

$$\begin{aligned}
 H(x)(t) &= \int_{-1}^1 \sin(\pi\tau)x(t+\tau)d\tau \\
 &= \int_{-\infty}^{\infty} \Pi(\tau/2)\sin(\pi\tau)x(t+\tau)d\tau \\
 &= -\int_{\infty}^{-\infty} \Pi(-\tau/2)\sin(-\pi\tau)x(t-\tau)d\tau \quad (\text{ch. var. } \tau \rightarrow -\tau) \\
 &= -\int_{-\infty}^{\infty} \Pi(\tau/2)\sin(\pi\tau)x(t-\tau)d\tau \quad (\text{ch. var. } \tau \rightarrow -\tau) \\
 &= (h * x)(t),
 \end{aligned}$$

where we put  $h(t) = -\Pi(t/2)\sin(\pi t)$ . It follows that  $h$  is the impulse response of  $H$ . Since  $h$  has an impulse response it is regular, and since it is regular its also linear and time invariant.

The hard way is to first show linear, then show time invariance, and then find this impulse response using (3.1.2). We have

$$\begin{aligned}
 H(ax + by) &= \int_{-1}^1 \sin(\pi\tau)(ax(t+\tau) + by(t+\tau))d\tau \\
 &= a \int_{-1}^1 \sin(\pi\tau)x(t+\tau)d\tau + b \int_{-1}^1 \sin(\pi\tau)y(t+\tau)d\tau \\
 &= aH(x) + bH(y),
 \end{aligned}$$

and so,  $H$  is linear. We also have

$$\begin{aligned}
 H(T_k(x)) &= \int_{-1}^1 \sin(\pi\tau)T_k(x)(t+\tau)d\tau \\
 &= \int_{-1}^1 \sin(\pi\tau)x(t+\tau-k)d\tau \\
 &= T_k \left( \int_{-1}^1 \sin(\pi\tau)x(t+\tau)d\tau \right) \\
 &= T_k(H(x)),
 \end{aligned}$$

and so,  $H$  is time invariant. Now, if  $H$  is regular then its impulse response is  $h = \lim_{\gamma \rightarrow \infty} H(p_\gamma)$ . Let  $h_\gamma$  be the signal

$$h_\gamma(t) = \int_{-1}^1 \sin(\pi\tau)p_\gamma(t+\tau)d\tau.$$

The impulse response exists if  $h_\gamma$  converges for each fixed  $t$  as  $\gamma \rightarrow \infty$ . Now,  $p_\gamma(t+\tau) = \gamma$  for  $t+\tau \in [0, \frac{1}{\gamma}]$ , i.e.  $\tau \in [-t, \frac{1}{\gamma} - t]$ , and zero otherwise. The integral ranges from  $-1$  to  $1$  so we are also interested in those  $\tau \in [-1, 1]$ . When  $t > \frac{1}{\gamma} + 1$  or  $t < -1$  the intervals  $[-1, 1]$  and  $[-t, \frac{1}{\gamma} - t]$  are disjoint and we obtain  $h(t) = 0$ . Otherwise, when  $[-t, \frac{1}{\gamma} - t] \subset [-1, 1]$ , i.e.,  $-t > -1$  and  $\frac{1}{\gamma} - t < 1$  we obtain

$$\begin{aligned}
 h_\gamma(t) &= \int_{-1}^1 \sin(\pi\tau)p_\gamma(t+\tau)d\tau \\
 &= \gamma \int_{-t}^{1/\gamma-t} \sin(\pi\tau)d\tau \\
 &= -\frac{\gamma}{\pi} (\cos(\pi(1/\gamma - t)) - \cos(-\pi t)) \\
 &= -\frac{\gamma}{\pi} (\cos(\pi(t - \frac{1}{\gamma})) - \cos(\pi t)).
 \end{aligned}$$

Put  $\Delta = -\frac{1}{\gamma}$  and

$$h_\gamma(t) = \frac{1}{\pi} \frac{\cos(\pi(t + \delta)) - \cos(\pi t)}{\delta}.$$

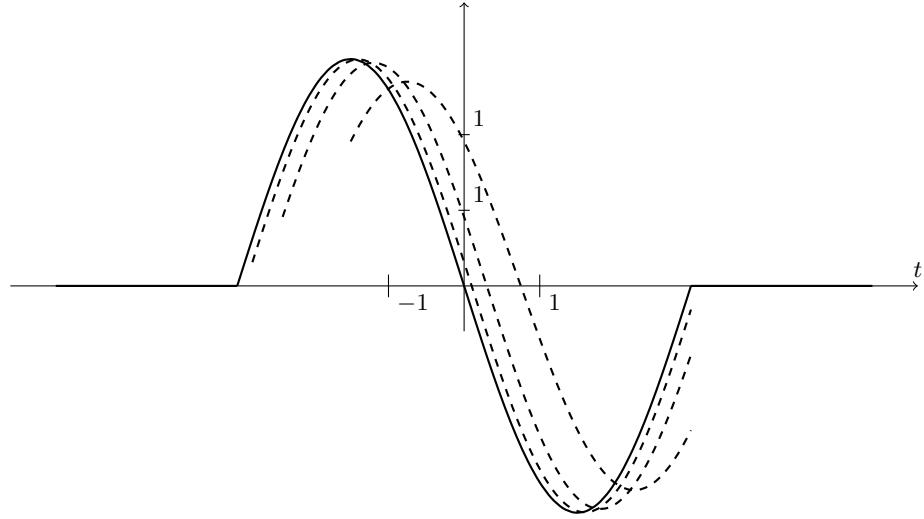
Recognising the limit as  $\gamma \rightarrow \infty$ , or equivalently as  $\Delta \rightarrow 0$  as

$$\lim_{\delta \rightarrow 0} \frac{\cos(\pi(t + \delta)) - \cos(\pi t)}{\delta} = \frac{d}{dt} \cos(\pi t)$$

we immediately have

$$\lim_{\gamma \rightarrow \infty} h_\gamma(t) = h(t) = \frac{1}{\pi} \frac{d}{dt} \cos(\pi t) = -\sin(\pi t).$$

on the interval  $t \in [\frac{1}{\gamma} - 1, 1]$ . It remains to show what happens on the interval  $[-1, \frac{1}{\gamma} - 1]$  that shrinks as  $\gamma \rightarrow \infty$ .



- 3.5. Show that  $\sum_{\ell=1}^L e^{\beta\ell} = \frac{e^{\beta(L+1)} - e^\beta}{e^\beta - 1}$  (Hint: sum a geometric progression).

**Solution:** Put  $r = e^\beta$  and put

$$S_L = \sum_{\ell=1}^L e^{\beta\ell} = \sum_{\ell=1}^L r^\ell.$$

This is the sum of the first  $L$  terms of a geometric progression. We have

$$rS_L - S_L = r^{L+1} - r$$

and so

$$S_L = \frac{r^{L+1} - r}{r - 1} = \frac{e^{\beta(L+1)} - e^\beta}{e^\beta - 1}$$

as required.

- 3.6. Show that

$$\frac{2j}{L} \sum_{\ell=1}^L \sin(\gamma\ell - \theta) e^{-j\gamma\ell} = \alpha + \alpha^* C$$

where  $\alpha = e^{-j\theta}$  and  $C = e^{-j\gamma(L+1)} \frac{\sin(\gamma L)}{L \sin(\gamma)}$ . (Hint: solve Exercise 3.5 first and then use the formula  $2j \sin(x) = e^{jx} - e^{-jx}$ ). **Solution:** We have

$$2j \sin(\gamma\ell - \theta) = e^{j(\gamma\ell - \theta)} - e^{-j(\gamma\ell - \theta)}$$

and so the sum becomes

$$\begin{aligned} \frac{1}{L} \sum_{\ell=1}^L (e^{j(\gamma\ell - \theta)} - e^{-j(\gamma\ell - \theta)}) e^{-j\gamma\ell} &= \frac{1}{L} \sum_{\ell=1}^L e^{-j\theta} - \frac{1}{L} \sum_{\ell=1}^L e^{-2j\gamma} e^{j\theta} \\ &= \alpha - \frac{\alpha^*}{L} \sum_{\ell=1}^L e^{-2j\gamma}. \end{aligned}$$

The sum is a geometric progression and, using the answer to Exercise 3.5, we have

$$\sum_{\ell=1}^L e^{-2j\gamma} = \frac{e^{-2j\gamma(L+1)} - e^{-2j\gamma}}{e^{-2j\gamma} - 1}.$$

The denominator satisfies

$$e^{-2j\gamma} - 1 = e^{-j\gamma}(e^{-j\gamma} - e^{j\gamma}) = -2je^{-j\gamma} \sin(\gamma).$$

The numerator satisfies

$$\begin{aligned} e^{-2j\gamma(L+1)} - e^{-2j\gamma} &= e^{-2j\gamma}(e^{-2j\gamma L} - 1) \\ &= e^{-2j\gamma} e^{-j\gamma L} (e^{-j\gamma L} - e^{j\gamma L}) \\ &= -2je^{-j\gamma(L+2)} \sin(\gamma L). \end{aligned}$$

Thus

$$\sum_{\ell=1}^L e^{-2j\gamma} = \frac{-2je^{-j\gamma(L+2)} \sin(\gamma L)}{-2je^{-j\gamma} \sin(\gamma)} = \frac{e^{-j\gamma(L+1)} \sin(\gamma L)}{\sin(\gamma)} = LC$$

where  $C$  is defined in the question statement. Now

$$\frac{2j}{L} \sum_{\ell=1}^L \sin(\gamma\ell - \theta) e^{-j\gamma\ell} = \alpha - \frac{\alpha^*}{L} LC = \alpha - \alpha^* C$$

as required.



## Chapter 4

# The Laplace transform

Let  $x: \mathbb{R} \rightarrow \mathbb{C}$  be a complex valued function of the real line (a signal). The function

$$\mathcal{L}(x) = \int_{-\infty}^{\infty} x(t)e^{-st}dt \quad (4.0.1)$$

is called the **Laplace transform** of  $x$ . The Laplace transform is a function of the complex parameter  $s$ , and if we need to indicate this we write  $\mathcal{L}(x)(s)$  or  $\mathcal{L}(x, s)$ . The Laplace transform  $\mathcal{L}(x)$  is not necessarily defined for all values of  $s \in \mathbb{C}$ . Let  $R$  be the set of real numbers such that  $x(t)e^{-\sigma t}$  is absolutely integrable if and only if  $\sigma \in R$ , that is

$$\int_{-\infty}^{\infty} |x(t)| e^{-\sigma t} dt < \infty \quad \text{if and only if } \sigma \in R.$$

In this case,  $\mathcal{L}(x, s)$  is finite for all  $s$  with real part satisfying  $\operatorname{Re}(s) \in R$  because

$$|\mathcal{L}(x, s)| = \left| \int_{-\infty}^{\infty} x(t)e^{-st} dt \right| \leq \int_{-\infty}^{\infty} |x(t)| e^{-\operatorname{Re}(s)t} dt < \infty.$$

The subset of the complex plane with real part from  $R$  is called the **region of convergence** (ROC) of the signal  $x$ .

For example, the Laplace transform of the right sided signal  $e^{\alpha t}u(t)$  is

$$\begin{aligned} \mathcal{L}(e^{\alpha t}u(t)) &= \int_{-\infty}^{\infty} e^{\alpha t}e^{-st}u(t)dt \\ &= \int_0^{\infty} e^{(\alpha-s)t}dt \\ &= \lim_{t \rightarrow \infty} \frac{e^{(\alpha-s)t}}{\alpha - s} - \frac{1}{\alpha - s}. \end{aligned}$$

The limit converges for all  $s$  with  $\operatorname{Re}(\alpha - s) < 0$ . Thus, the Laplace transform of  $e^{\alpha t}u(t)$  is

$$\mathcal{L}(e^{\alpha t}u(t)) = \frac{1}{s - \alpha} \quad \operatorname{Re}(s) > \operatorname{Re}(\alpha)$$

The region of convergence of  $e^{\alpha t}u(t)$  is the subset of the complex plane with real part greater than  $\text{Re}(\alpha)$ . Figure 4.1 shows the region of convergence when  $\text{Re}(\alpha) = -2$ . Now consider the left sided signal  $e^{\beta t}u(-t)$  with Laplace transform

$$\mathcal{L}(e^{\beta t}u(-t)) = \lim_{t \rightarrow -\infty} \frac{e^{(\beta-s)t}}{\beta - s} + \frac{1}{\beta - s}.$$

The limit converges when  $\text{Re}(\beta - s) > 0$ , and so,

$$\mathcal{L}(e^{\beta t}u(-t)) = \frac{1}{\beta - s} \quad \text{Re}(s) < \text{Re}(\beta).$$

The region of convergence of  $e^{\beta t}u(-t)$  is those  $s \in \mathbb{C}$  such that  $\text{Re}(s) < \text{Re}(\beta)$ . The signal  $ae^{\alpha t}u(t) + be^{\beta t}u(-t)$  has Laplace transform

$$\begin{aligned} \mathcal{L}(ae^{\alpha t}u(t) + be^{\beta t}u(-t)) &= \int_{-\infty}^{\infty} (ae^{\alpha t}u(t) + be^{\beta t}u(-t))e^{-st}dt \\ &= a \int_{-\infty}^{\infty} e^{\alpha t}u(t)e^{-st}dt + b \int_{-\infty}^{\infty} e^{\beta t}u(-t)e^{-st}dt \\ &= a\mathcal{L}(e^{\alpha t}u(t)) + b\mathcal{L}(e^{\beta t}u(-t)) \end{aligned}$$

that is finite only when  $\text{Re}(\alpha) < \text{Re}(s) < \text{Re}(\beta)$ . The corresponding ROC is shown in Figure 4.1 when  $\text{Re}(\alpha) = -2$  and  $\text{Re}(\beta) = 3$ . In the previous equation we have discovered that the Laplace transform is **linear**, that is, for signals  $x$  and  $y$  and constants  $a$  and  $b$ ,

$$\mathcal{L}(ax + by) = a\mathcal{L}(x) + b\mathcal{L}(y). \quad (4.0.2)$$

In words: the Laplace transform of a linear combination of signals is the same linear combination of the Laplace transforms of those signals.

In the previous example the Laplace transform is guaranteed to be finite for any  $s$  if  $\text{Re}(\alpha) \geq \text{Re}(\beta)$ , and the region of convergence is correspondingly the empty set. Other signals also have this property. For example, the signal  $x(t) = 1$  because

$$\mathcal{L}(1) = \int_{\infty}^{\infty} e^{-st}dt = \lim_{t \rightarrow -\infty} \frac{e^{-st}}{s} - \lim_{t \rightarrow \infty} \frac{e^{-st}}{s}$$

and the limit as  $t \rightarrow -\infty$  converges only when  $\text{Re}(s) < 0$  while the limit as  $t \rightarrow \infty$  converges only when  $\text{Re}(s) > 0$ .

As a final example, consider the rectangular pulse

$$\Pi(t) = \begin{cases} 1 & -\frac{1}{2} < t \leq \frac{1}{2} \\ 0 & \text{otherwise.} \end{cases}$$

Its Laplace transform is

$$\mathcal{L}(\Pi) = \int_{-\infty}^{\infty} \Pi(t)e^{-st}dt = \int_{-1/2}^{1/2} e^{-st}dt = \frac{e^{s/2} - e^{-s/2}}{s} \quad (4.0.3)$$

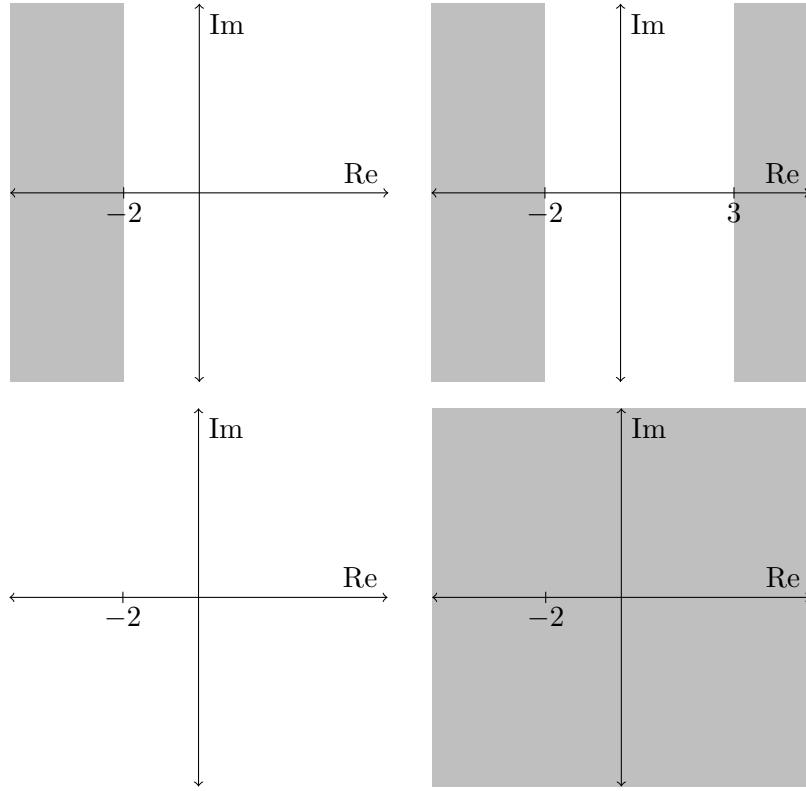


Figure 4.1: Regions of convergence (unshaded) for the signal  $e^{-2t}u(t)$  (top left), the signal  $e^{-2t}u(t) + e^{3t}u(-t)$  (top right), the rectangular pulse  $\Pi$  (bottom left), and the constant signal  $x(t) = 1$  (bottom right).

and is finite for all  $s \in \mathbb{C}$ . The region of convergence of the rectangular pulse  $\Pi$  is the entire complex plane. The examples just given exhibit all the possible types of regions of convergence. The region of convergence is either the entire complex plane, a left or right half plane, a vertical strip, or the empty set.

Given the Laplace transform  $\mathcal{L}(x)$  the signal  $x$  can be recovered by the **inverse Laplace transform**

$$x(t) = \mathcal{L}^{-1}(x) = \frac{1}{2\pi j} \lim_{\omega \rightarrow \infty} \int_{\sigma-j\omega}^{\sigma+j\omega} \mathcal{L}(x, s) e^{st} ds,$$

where  $\sigma$  is a real number that is inside the region of convergence of  $x$ . Solving the integral above typically requires a special type of integration called **contour integration** that we will not consider here [Stewart and Tall, 2004]. For our purposes, and for many engineering purposes, it suffices to remember only the following Laplace transform pair

$$\mathcal{L}(t^n u(t)) = \frac{n!}{s^{n+1}} \quad \operatorname{Re}(s) > 0, \quad (4.0.4)$$

where  $n \geq 0$  is an integer (Exercise 4.2). Let  $x(t)$  be a signal with region of convergence  $R$ . The Laplace transforms of the signal  $x(t)$  and the signal  $e^{\alpha t}x(t)$  are related. To see this write

$$\begin{aligned}\mathcal{L}(e^{\alpha t}x(t), s) &= \int_{-\infty}^{\infty} e^{\alpha t}x(t)e^{-st}dt \\ &= \int_{-\infty}^{\infty} x(t)e^{-(s-\alpha)t}dt \\ &= \mathcal{L}(x, s - \alpha) \quad \text{Re}(s - \alpha) \in R.\end{aligned}\quad (4.0.5)$$

This is called the **frequency shift rule**. Combining the frequency shift rule with (4.0.4) we obtain the transform pair

$$\mathcal{L}(t^n e^{\alpha t} u(t)) = \mathcal{L}(t^n u(t), s - \alpha) = \frac{n!}{(s - \alpha)^{n+1}} \quad \text{Re}(s) > \text{Re}(\alpha), \quad (4.0.6)$$

where  $n \geq 0$  is an integer. This is the only Laplace transform pair we require here.

A useful relationship exists between the Laplace transform of a signal  $x$  and its time scaled version  $x(\alpha t)$  where  $\alpha \neq 0$ . If  $x$  is a signal with region of convergence  $R$  then the time scaled signal  $x(\alpha t)$  with  $\alpha \neq 0$  has Laplace transform

$$\mathcal{L}(x(\alpha t), s) = \frac{1}{|\alpha|} \mathcal{L}(x, s/\alpha), \quad \text{Re}(s/\alpha) \in R. \quad (4.0.7)$$

This is called the **time scaling property** (Excercise 4.10).

## 4.1 The transfer function and the Laplace transform

Our purpose for introducing the Laplace transform is to study the response of a linear time invariant system  $H$  to exponential signals of the form  $e^{st}$ . Recall from Section 3.4 that exponential signals are **eigenfunctions** of linear time invariant systems. That is, if  $s \in \mathbb{C}$  such that the complex exponential signal  $e^{st}$  is in the domain of  $H$ , then response of  $H$  to  $e^{st}$  is  $\lambda e^{st}$  where  $\lambda \in \mathbb{C}$  is a constant that does not depend on  $t$ , but may depend on  $s$  and the system  $H$ . To highlight this dependence on  $H$  and  $s$  we write  $\lambda(H, s)$  or  $\lambda(H)(s)$  and do not distinguish between these notations. Considered as a function of  $s$ ,  $\lambda(H)$  is called the **transfer function** of the system  $H$ . For a given system  $H$ , we would like to understand how  $\lambda(H, s)$  behaves as  $s$  changes. In what follows we regularly drop the argument “ $(s)$ ” and simply write  $\lambda(H)$  as the transfer function of  $H$ .

Assume that  $H$  is a regular system with impulse response  $h$ . In this case,

$$\begin{aligned} H(e^{st}) &= e^{st}\lambda(H, s) = h * e^{st} \\ &= \int_{-\infty}^{\infty} h(\tau)e^{s(t-\tau)}d\tau \\ &= e^{st} \int_{-\infty}^{\infty} h(\tau)e^{-s\tau}d\tau \\ &= e^{st}\mathcal{L}(h, s), \end{aligned}$$

and so,  $\lambda(H) = \mathcal{L}(h)$ . That is, the transfer function of a regular system is precisely the Laplace transform of its impulse response. The region of convergence of the impulse response describes a set of complex exponential signals  $e^{st}$  in the domain of the system and we refer to this as the region of convergence of the *system*. In this way, both signals and systems have regions of convergence.

The transfer functions of the time-shifter and differentiator can be obtained by inspection. For the time-shifter

$$T_\tau(e^{st}) = e^{s(t-\tau)} = e^{-s\tau}e^{st} \quad \text{and so} \quad \lambda(T_\tau) = e^{-s\tau}. \quad (4.1.1)$$

The region of convergence is the whole complex plane  $s \in \mathbb{C}$ . For the special case of the identity system  $T_0$  we obtain  $\lambda(T_0, s) = 1$ . For the differentiator

$$D(e^{st}) = \frac{d}{dt}e^{st} = se^{st} \quad \text{and so} \quad \lambda(D) = s.$$

The region of convergence is the whole complex plane  $s \in \mathbb{C}$ . More generally, for the  $k$ th differentiator

$$D^k(e^{st}) = \frac{d^k}{dt^k}e^{st} = s^k e^{st} \quad \text{and so} \quad \lambda(D^k) = s^k. \quad (4.1.2)$$

The region of convergence is again the whole complex plane. These results motivate assigning the following Laplace transforms to the delta “function” and its derivatives

$$\mathcal{L}(\delta) = 1, \quad \mathcal{L}(\delta^k) = s^k.$$

These conventions are common in the literature [Oppenheim et al., 1996].

#### 4.1.1 The transfer function of a linear combination of systems

Let  $H = aH_1 + bH_2$  be a linear combination of systems  $H_1$  and  $H_2$ . Let  $R_1 \subseteq \mathbb{C}$  and  $R_2 \subseteq \mathbb{C}$  be the regions of convergence of  $H_1$  and  $H_2$ . We have,

$$\begin{aligned} H(e^{st}) &= aH_1(e^{st}) + bH_2(e^{st}) \\ &= a\lambda(H_1)e^{st} + b\lambda(H_2)e^{st} \quad s \in R_1 \cap R_2, \\ &= (a\lambda(H_1) + b\lambda(H_2))e^{st} \quad s \in R_1 \cap R_2, \\ &= \lambda(H)e^{st} \quad s \in R_1 \cap R_2, \end{aligned}$$

and so,

$$\lambda(H) = a\lambda(H_1) + b\lambda(H_2) \quad s \in R_1 \cap R_2.$$

That is, the transfer function of a linear combination of systems is the same linear combination of the transfer functions. The region of convergence of the linear combination is the intersection of the regions of convergence of the systems being combined.

#### 4.1.2 The transfer function of a composition of systems

Let  $H$  be the system constructed by composing two systems  $H_1$  and  $H_2$  with regions of convergence  $R_1$  and  $R_2$ , that is,  $H(x) = H_1(H_2(x))$ . The response of  $H$  to the signal  $e^{st}$  is

$$\begin{aligned} H(e^{st}) &= H_1(H_2(e^{st})) \\ &= H_1(\lambda(H_2)e^{st}) \quad s \in R_2 \\ &= \lambda(H_2)H_1(e^{st}) \quad s \in R_2 \\ &= \lambda(H_2)\lambda(H_1)e^{st} \quad s \in R_1 \cap R_2 \\ &= \lambda(H)e^{st} \quad s \in R_1 \cap R_2, \end{aligned}$$

and so,

$$\lambda(H) = \lambda(H_1)\lambda(H_2) \quad s \in R_1 \cap R_2. \quad (4.1.3)$$

That is, the transfer function of a composition of linear time invariant systems is the multiplication of the transfer functions of those systems. The region of convergence of the composition is the intersection of the regions of convergence of the systems being composed.

#### 4.1.3 The convolution theorem

We showed in Section 3.3 that if  $H_1$  and  $H_2$  are regular systems with impulse responses  $h_1$  and  $h_2$ , then the impulse of the system  $H(x) = H_1(H_2(x))$  is given by the convolution  $h = h_1 * h_2$ . Because,

$$\lambda(H) = \mathcal{L}(h) \quad \lambda(H_1) = \mathcal{L}(h_1) \quad \lambda(H_2) = \mathcal{L}(h_2),$$

and using (4.1.3), we obtain,

$$\mathcal{L}(h_1 * h_2) = \mathcal{L}(h) = \lambda(H) = \lambda(H_1)\lambda(H_2) = \mathcal{L}(h_1)\mathcal{L}(h_2), \quad s \in R_1 \cap R_2.$$

Putting  $x = h_1$ ,  $y = h_2$ ,  $R_x = R_1$ , and  $R_y = R_2$  we obtain the **convolution theorem**,

$$\mathcal{L}(x * y) = \mathcal{L}(x)\mathcal{L}(y), \quad s \in R_x \cap R_y. \quad (4.1.4)$$

In words: the Laplace transform of a convolution of signals is the multiplication of their Laplace transforms.

#### 4.1.4 The Laplace transform of an output signal

Let  $H$  be a regular system with impulse response  $h$  and let  $y = H(x) = h * x$  be the response of  $H$  to input signal  $x$ . Using the convolution theorem, the Laplace transform of the output signal  $y$  is

$$\mathcal{L}(y) = \mathcal{L}(h)\mathcal{L}(x) = \lambda(H)\mathcal{L}(x), \quad s \in R \cap R_x, \quad (4.1.5)$$

where  $R$  is the region of convergence of the system  $H$  and  $R_x$  is the region of convergence of the input signal  $x$ . Thus, the Laplace transform of the output signal  $y = H(x)$  is the transfer function of the system  $H$  multiplied by the Laplace transform of the input signal  $x$ . This result also holds when  $H$  is a time-shifter or a differentiator (Exercise 4.1.5).

## 4.2 Solving differential equations

Assume we have a system modelled by a differential equation of the form

$$\sum_{\ell=0}^m a_\ell D^\ell(x) = \sum_{\ell=0}^k b_\ell D^\ell(y), \quad (4.2.1)$$

where  $x$  and  $y$  are signals. Taking Laplace transforms of both sides of this equation,

$$\begin{aligned} \mathcal{L}\left(\sum_{\ell=0}^m a_\ell D^\ell(x)\right) &= \mathcal{L}\left(\sum_{\ell=0}^k b_\ell D^\ell(y)\right) \\ \sum_{\ell=0}^m a_\ell \mathcal{L}(D^\ell(x)) &= \sum_{\ell=0}^k b_\ell \mathcal{L}(D^\ell(y)) \quad (\text{linearity (4.0.2)}) \\ \sum_{\ell=0}^m a_\ell \lambda(D^\ell) \mathcal{L}(x) &= \sum_{\ell=0}^k b_\ell \lambda(D^\ell) \mathcal{L}(y) \quad (\text{using (4.1.5)}) \\ \sum_{\ell=0}^m a_\ell s^\ell \mathcal{L}(x) &= \sum_{\ell=0}^k b_\ell s^\ell \mathcal{L}(y). \quad (\text{since } \lambda(D^\ell) = s^\ell \text{ by (4.1.2)}) \end{aligned}$$

We have obtained an equation relating the Laplace transforms of  $x$  and  $y$ ,

$$\mathcal{L}(x)(a_0 + a_1 s + \dots + a_m s^m) = \mathcal{L}(y)(b_0 + b_1 s + \dots + b_k s^k).$$

Rearranging this equation we obtain

$$\mathcal{L}(y) = \frac{a_0 + a_1 s + \dots + a_m s^m}{b_0 + b_1 s + \dots + b_k s^k} \mathcal{L}(x).$$

Let  $H$  be a linear time invariant system such that  $y = H(x)$  whenever  $x$  and  $y$  satisfy the differential equation (4.2.1). According to (4.1.5) the transfer function of  $H$  is

$$\lambda(H) = \frac{\mathcal{L}(y)}{\mathcal{L}(x)} = \frac{a_0 + a_1 s + \dots + a_m s^m}{b_0 + b_1 s + \dots + b_k s^k}.$$

Properties of  $H$  can be obtained by inspecting this transfer function. For example, the impulse response of  $H$  (if it exists) can be obtained by applying the inverse Laplace transform.

We now apply these results to the differential equations that model the RC electrical circuit from Figure 2.1 and the mass spring damper from Figure 2.2. The RC circuit is an example of what is called a **first order system** and the mass spring damper is an example of what is called a **second order system**.

### 4.3 First order systems

Recall the passive electrical RC circuit from Figure 2.1. The differential equation modelling this circuit is (2.0.1),

$$x = y + RCD(y),$$

where  $x$  is the input voltage signal,  $y$  is the voltage over the capacitor, and  $R$  and  $C$  are the resistance and capacitance. The RC circuit is an example of a **first order system**. Let  $H$  be a system mapping the input voltage signal  $x$  to the output voltage signal  $y$ . We will discover the impulse response of  $H$ . Taking the Laplace transform on both sides of the differential equation gives

$$\mathcal{L}(x) = (1 + RCs)\mathcal{L}(y),$$

and it follows that the transfer function of  $H$  is

$$\lambda(H) = \frac{\mathcal{L}(y)}{\mathcal{L}(x)} = \frac{1}{1 + RCs} = \frac{r}{r + s},$$

where  $r = \frac{1}{RC}$ . The value  $\frac{1}{r} = RC$  is called the **time constant**. The impulse response of  $H$  is given by the inverse of this Laplace transform. There are two signals with Laplace transform  $\frac{r}{r+s}$ : the right sided signal  $re^{-rt}u(t)$  with region of convergence  $\text{Re}(s) > -r$ , and the left sided signal  $-re^{-rt}u(-t)$  with region of convergence  $\text{Re}(s) < -r$ . The RC circuit (and in fact all physically realisable systems) are expected to be causal. For this reason, the left sided signal  $-re^{-rt}u(-t)$  cannot be the impulse response of  $H$ . The impulse response is the right sided signal

$$h(t) = re^{-rt}u(t).$$

Given an input voltage signal  $x$  we can now find the corresponding output signal  $y = H(x)$  by convolving  $x$  with the impulse response  $h$ . That is,

$$y = H(x) = h * x = \int_{-\infty}^{\infty} re^{-r\tau} u(\tau) x(t - \tau) d\tau = r \int_0^{\infty} e^{-rt} x(t - \tau) d\tau.$$

If  $r \geq 0$  the impulse response is absolutely integrable, that is,

$$\begin{aligned} \|h\|_1 &= \int_{-\infty}^{\infty} |re^{-rt} u(t)| dt \\ &= r \int_0^{\infty} e^{-rt} dt \\ &= 1 - \lim_{t \rightarrow \infty} e^{-rt} = 1, \end{aligned}$$

and the system is stable (Exercise 3.3). However, if  $r < 0$  the impulse response is not absolutely integrable, and the system is not stable. Figure 4.3 shows the impulse response when  $r = -\frac{1}{5}, -\frac{1}{3}, -\frac{1}{2}, 1, 2$ . In a passive electrical RC circuit the resistance  $R$  and capacitance  $C$  are always positive and  $r = \frac{1}{RC}$  is positive. For this reason, passive electrical RC circuits are always stable.

From (3.1.5), the step response  $H(u)$  is given by applying the integrator  $I_{\infty}$  to the impulse response, that is,

$$H(u) = I_{\infty}(h) = \int_{-\infty}^t re^{-r\tau} u(\tau) d\tau = \begin{cases} r \int_0^t e^{-r\tau} d\tau & t > 0 \\ 0 & \text{otherwise} \end{cases}$$

or more simply

$$H(u) = (1 - e^{-rt})u(t). \quad (4.3.1)$$

This step response is plotted in Figure 4.3.

**Test 5 (The active RC circuit again)** In this test we again use the active RC circuit from Test 3 with resistors  $R = R_1 = R_2 = 27\text{k}\Omega$  and capacitors  $C = C_2 = 10\text{nF}$ . In Test 3 we applied the differential equation (2.2.4) to the reconstructed output signal  $\tilde{y}$  and asserted that the resulting signal was close to the reconstructed input signal  $\tilde{x}$ . In this test we instead convolve the input signal  $\tilde{x}$  with the impulse response

$$h = -\frac{1}{RC}e^{-t/RC} = -re^{-rt}, \quad r = \frac{1}{RC} = \frac{10^5}{27}$$

and assert that the resulting signal is close to the output signal  $\tilde{y}$ . That is, we test the expected relationship

$$\tilde{y} \approx h * \tilde{x} = - \int_{-\infty}^{\infty} h(\tau) u(\tau) \tilde{x}(t - \tau) d\tau = \int_0^{\infty} h(\tau) \tilde{x}(t - \tau) d\tau.$$

From (1.2.4),

$$\begin{aligned}\tilde{y}(t) &\approx \int_0^\infty h(\tau) \sum_{\ell=1}^L x_\ell \operatorname{sinc}(Ft - F\tau - \ell) d\tau \\ &= \sum_{\ell=1}^L x_\ell \int_0^\infty h(\tau) \operatorname{sinc}(Ft - F\tau - \ell) d\tau \\ &= \sum_{\ell=1}^L x_\ell f(Ft - \ell),\end{aligned}$$

where the function

$$f(t) = \int_0^\infty h(\tau) \operatorname{sinc}(t - F\tau) d\tau = -r \int_0^\infty e^{-r\tau} \operatorname{sinc}(t - F\tau) d\tau.$$

An approximation of  $f(t)$  is made using the trapezoidal sum

$$f(t) \approx \frac{K}{2N} \left( g(0) + g(K) + 2 \sum_{n=1}^{N-1} g(\Delta n) \right),$$

where  $g(\tau) = h(\tau) \operatorname{sinc}(t - F\tau)$ , and

$$K = -RC \log(10^{-3}), \quad N = \lceil 10FK \rceil, \quad \Delta = K/N.$$

Figure 4.2 plots the input signal  $\tilde{x}$ , output signal  $\tilde{y}$ , and hypothesised output signal  $h * \tilde{x}$  over a 4ms window.

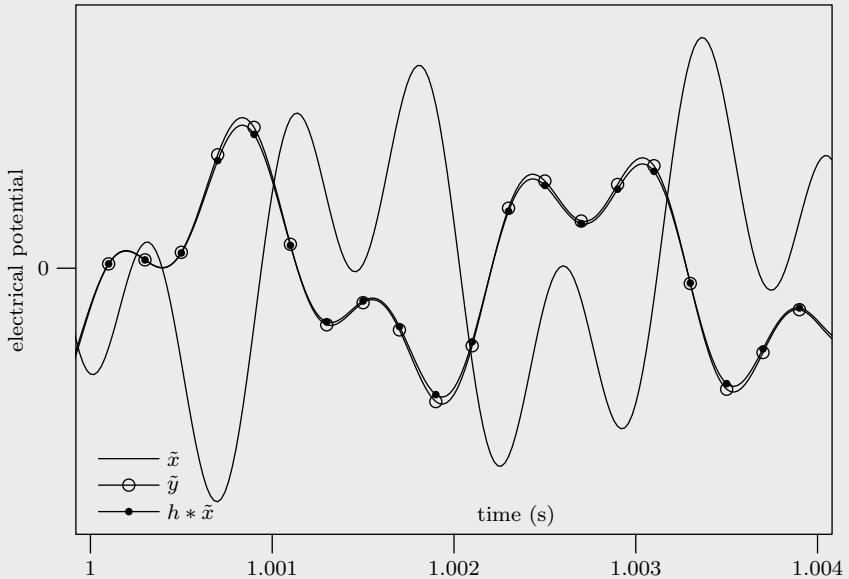


Figure 4.2: Plot of reconstructed input signal  $\tilde{x}$  (solid line), output signal  $\tilde{y}$  (solid line with circle), and hypothesised output signal  $h * \tilde{x}$  (solid line with dot).

## 4.4 Second order systems

Consider the mass spring damper system from Figure 2.2 that is described by the equation

$$f = Kp + BD(p) + MD^2(p) \quad (4.4.1)$$

where  $f$  is the force applied to the mass  $M$  and  $p$  is the position of the mass and  $K$  and  $B$  are the spring and damping coefficients. The mass spring damper is an example of a **second order system**. Another example of a second order system is the Sallen-Key active electrical circuit depicted in Figure 2.10. In Section 2 we were able to find the force  $f$  corresponding with a given position signal  $p$ . Let  $H$  be a system mapping  $f$  to  $p$ , that is, such that  $p = H(f)$ . We will find the impulse response of  $H$ . Taking Laplace transforms on both sides of the differential equation gives

$$\mathcal{L}(f) = (K + Bs + Ms^2)\mathcal{L}(p).$$

Rearranging gives the transfer function of  $H$ ,

$$\lambda(H) = \frac{\mathcal{L}(p)}{\mathcal{L}(f)} = \frac{1}{K + Bs + Ms^2}.$$

We can invert this Laplace transform to obtain the impulse response. There are three cases to consider depending on whether the quadratic  $K + Bs + Ms^2$  has two distinct real roots, is irreducible (does not have real roots), or has two identical real roots.

**Case 1: (Distinct real roots)** In this case, the roots are

$$\beta - \alpha, \quad -\beta - \alpha,$$

where

$$\alpha = \frac{B}{2M}, \quad \beta = \frac{\sqrt{B^2 - 4KM}}{2M}$$

and  $B^2 - 4KM > 0$ . By a partial fraction expansion (Exercise 4.7),

$$\begin{aligned} \lambda(H) &= \frac{1}{M(s - \beta + \alpha)(s + \beta + \alpha)} \\ &= \frac{1}{2\beta M} \left( \frac{1}{s - \beta + \alpha} - \frac{1}{s + \beta + \alpha} \right). \end{aligned}$$

From (4.0.6) we obtain the transform pairs

$$\mathcal{L}(e^{(\beta-\alpha)t}u(t)) = \frac{1}{s - \beta + \alpha}, \quad \mathcal{L}(e^{-(\beta+\alpha)t}u(t)) = \frac{1}{s + \beta + \alpha}.$$

As in Section 4.3, other signals with these Laplace transforms are discarded because they do not lead to an impulse response that is zero for  $t < 0$ . That

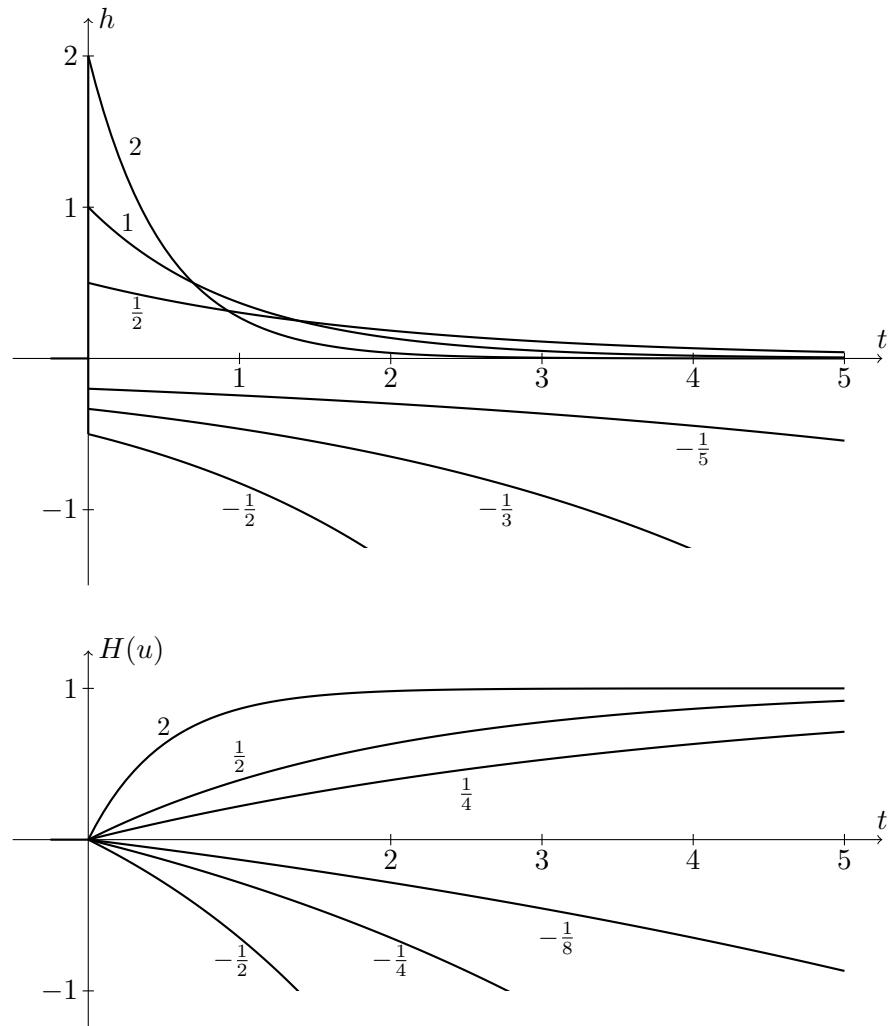


Figure 4.3: Top: impulse response of a first order system with  $r = -\frac{1}{2}, -\frac{1}{3}, -\frac{1}{5}, \frac{1}{2}, 1, 2$ . Bottom: step response of a first order system with  $r = -\frac{1}{2}, -\frac{1}{4}, -\frac{1}{8}, \frac{1}{4}, \frac{1}{2}, 2$ .

is, they do not lead to a causal system  $H$ . The impulse response of  $H$  is thus

$$h(t) = \frac{1}{2\beta M} u(t) e^{-\alpha t} (e^{\beta t} - e^{-\beta t}).$$

This is a sum of the impulse responses of two first order systems.

**Case 2: (Distinct imaginary roots)** The solution is as in the previous case, but now  $4KM - B^2 > 0$  and  $\beta$  is imaginary. Put  $\theta = \beta/j$  so that

$$e^{\beta t} - e^{-\beta t} = e^{j\theta t} - e^{-j\theta t} = 2j \sin(\theta t).$$

The impulse response of  $H$  is

$$h(t) = \frac{1}{\theta M} u(t) e^{-\alpha t} \sin(\theta t).$$

**Case 3: (Identical roots)** In this case, the two roots are equal to  $-\alpha$  and

$$\lambda(H) = \frac{1}{M(s + \alpha)^2}.$$

From (4.0.6) we obtain the transform pair

$$\mathcal{L}(te^{-\alpha t} u(t)) = \frac{1}{(s + \alpha)^2}$$

and this is the only signal with this Laplace transform that leads to a causal impulse response. The impulse response of  $H$  is thus

$$h(t) = \frac{1}{M} te^{-\alpha t} u(t).$$

A second order system is called **overdamped** when there are two distinct real roots, **underdamped** when their are two distinct imaginary roots, and **critically damped** when the roots are identical. The different types of impulse responses for are plotted in Figure 4.4.

With no damping (i.e. damping coefficient  $B = 0$ ) the roots are of the form  $\pm\beta$  and have no real part. In this case, the impulse response is

$$h(t) = \frac{1}{\theta M} u(t) \sin(\theta t),$$

where  $\theta = \beta/j = \sqrt{KM}$  is called the **natural frequency** of the second order system. This impulse response oscillates for all  $t > 0$  without decay or explosion. Two identical roots occur when the damping coefficient  $B = \sqrt{4KM}$  and this is sometimes called the **critical damping coefficient**.

The impulse response of a second order system is absolutely integrable when  $\alpha = \frac{B}{2M} > 0$ , but not when  $\alpha \leq 0$ . Thus, the system is stable when

$\alpha > 0$  and not stable when  $\alpha \leq 0$ . For the mass spring damper both the mass  $M$  and damping coefficient  $B$  are positive and so mass spring dampers are always stable.

From (3.1.5) the step response  $H(u)$  is given by applying the integrator  $I_\infty$  to the impulse response. There are three cases to consider depending on whether the system is overdamped, underdamped, or critically damped. When the system is overdamped the step response is

$$\begin{aligned} H(u) = I_\infty(h) &= \frac{1}{2\beta M} \int_{-\infty}^t e^{-\alpha\tau} (e^{\beta\tau} - e^{-\beta\tau}) u(\tau) d\tau \\ &= \frac{1}{2\beta M} \int_0^t e^{-\alpha\tau} (e^{\beta\tau} - e^{-\beta\tau}) d\tau \\ &= \frac{1}{2\beta M} u(t) \left( \frac{e^{(\beta-\alpha)t} - 1}{\beta - \alpha} + \frac{e^{-(\beta+\alpha)t} - 1}{\beta + \alpha} \right). \end{aligned}$$

When the system is underdamped the step response is

$$\begin{aligned} H(u) = I_\infty(h) &= \frac{1}{\theta M} \int_0^t e^{-\alpha\tau} \sin(\theta\tau) dt \\ &= u(t) \left( \frac{\theta - e^{-t\alpha} (\theta \cos(t\theta) + \alpha \sin(t\theta))}{M\theta(\alpha^2 + \theta^2)} \right). \end{aligned}$$

When the system is critically damped the step response is

$$\begin{aligned} H(u) = I_\infty(h) &= \frac{1}{\theta M} \int_0^t \frac{1}{M} t e^{-\alpha t} dt \\ &= \frac{1}{M\alpha^2} u(t) (1 - e^{-t\alpha s} (1 + t\alpha)). \end{aligned}$$

These step responses are plotted in Figure 4.5.

## 4.5 Poles, zeros, and stability

As discussed in Section 4.2 the transfer function of a system described by a linear differential equation with constant coefficients is of the form

$$\lambda(H) = \frac{\mathcal{L}(y)}{\mathcal{L}(x)} = \frac{a_0 + a_1 s + \dots + a_m s^m}{b_0 + b_1 s + \dots + b_k s^k}.$$

Factorising the polynomials on the numerator and denominator we obtain

$$\lambda(H) = C \frac{(s - \alpha_0)(s - \alpha_1) \cdots (s - \alpha_m)}{(s - \beta_0)(s - \beta_1) \cdots (s - \beta_k)},$$

where  $\alpha_0, \dots, \alpha_m$  are the roots of the numerator polynomial  $a_0 + a_1 s + \dots + a_m s^m$ , and  $\beta_0, \dots, \beta_k$  are the roots of the denominator polynomial

Figure 4.4: Impulse response of the mass spring damper with  $M = 1$ ,  $K = \frac{\pi^2}{4}$  and damping constant  $B = \frac{\pi}{3}$  (underdamped),  $B = \sqrt{4KM} = \pi$  (critically damped), and  $B = 2\pi$  (overdamped).

Figure 4.5: Step response of the mass spring damper with  $M = 1$ ,  $K = \frac{\pi^2}{4}$  and damping constant  $B = \frac{\pi}{3}$  (underdamped),  $B = \sqrt{4KM} = \pi$  (critically damped), and  $B = 2\pi$  (overdamped).

$b_0 + b_1 s + \dots + b_k s^k$ , and  $C = \frac{a_m}{b_m}$ . That such a factorisation is always possible is called the **fundamental theorem of algebra** [Fine and Rosenberger, 1997]. If the numerator and denominator polynomials share one or more roots, then these roots cancel leaving the simpler expression

$$\lambda(H) = C \frac{(s - \alpha_d)(s - \alpha_1) \cdots (s - \alpha_m)}{(s - \beta_d)(s - \beta_1) \cdots (s - \beta_k)}, \quad (4.5.1)$$

where  $d$  is the number of shared roots, these shared roots being

$$\alpha_0 = \beta_0, \quad \alpha_1 = \beta_1, \quad \dots, \quad \alpha_{d-1} = \beta_{d-1}.$$

The roots from the numerator  $\alpha_d, \dots, \alpha_m$  are called the **zeros** and the roots from the denominator  $\beta_d, \dots, \beta_m$  are called the **poles**. A **pole-zero plot** is constructed by marking the complex plane with a cross at the location of each pole and a circle at the location of each zero. Pole-zero plots for the first order system from Section 4.3, the second order system from Section 4.4, and the system describing the PID controller (2.2.7) are shown in Figure 4.6.

It is always possible to apply partial fractions and write (4.5.1) in the form

$$\lambda(H) = p(s) + \sum_{\ell \in K} \frac{A_\ell}{(s - \beta_\ell)^{r_\ell}},$$

where  $r_\ell$  are positive integers,  $A_\ell$  are complex constants,  $K$  is a subset of the indices from  $\{d, d+1, \dots, k\}$ , and  $p(s)$  is a polynomial of degree  $m-k$ . If  $k > m$  then  $p(s) = 0$ . The integer  $r_\ell$  is called the **multiplicity** of the pole  $\beta_\ell$ . We now restrict attention to the common case when the coefficients of the numerator polynomial  $a_0, \dots, a_m$  and the coefficients of the denominator polynomial  $b_0, \dots, b_k$  are real. In this case, the coefficients of the polynomial  $p(s)$  are real, and the constant  $A_\ell$  is real whenever the corresponding pole  $\beta_\ell$  is real. If the pole  $\beta_\ell$  has nonzero imaginary part there always exists another pole  $\beta_i$  such that  $\beta_\ell = \beta_i^*$ , where  $\beta_i^*$  is the **complex conjugate** of  $\beta_i$ . These poles have the same multiplicity, that is,  $r_\ell = r_i$ , and also the constants  $A_\ell = A_i^*$ . Stated another way: the complex poles occur in **conjugate pairs**.

We see that the transfer function contains the summation of two parts: the polynomial  $p(s)$ , and a sum of terms of the form  $\frac{A}{(s-\beta)^r}$ . Let  $p(s) = \gamma_0 + \gamma_1 s + \dots + \gamma_{m-k} s^{m-k}$ . This polynomial is the transfer function of the nonregular system

$$H_1 = \gamma_0 T_0 + \gamma_1 D + \gamma_2 D^2 + \dots + \gamma_{m-k} D^{m-k}.$$

This system is a linear combination of the identity system  $T_0$  and differentiators of order at most  $m-k$ . From (4.0.6),

$$\mathcal{L}\left(\frac{A}{r!} t^{r-1} e^{\beta t} u(t)\right) = \frac{A}{(s - \beta)^r} \quad \text{Re}(s) > \text{Re}(\beta)$$

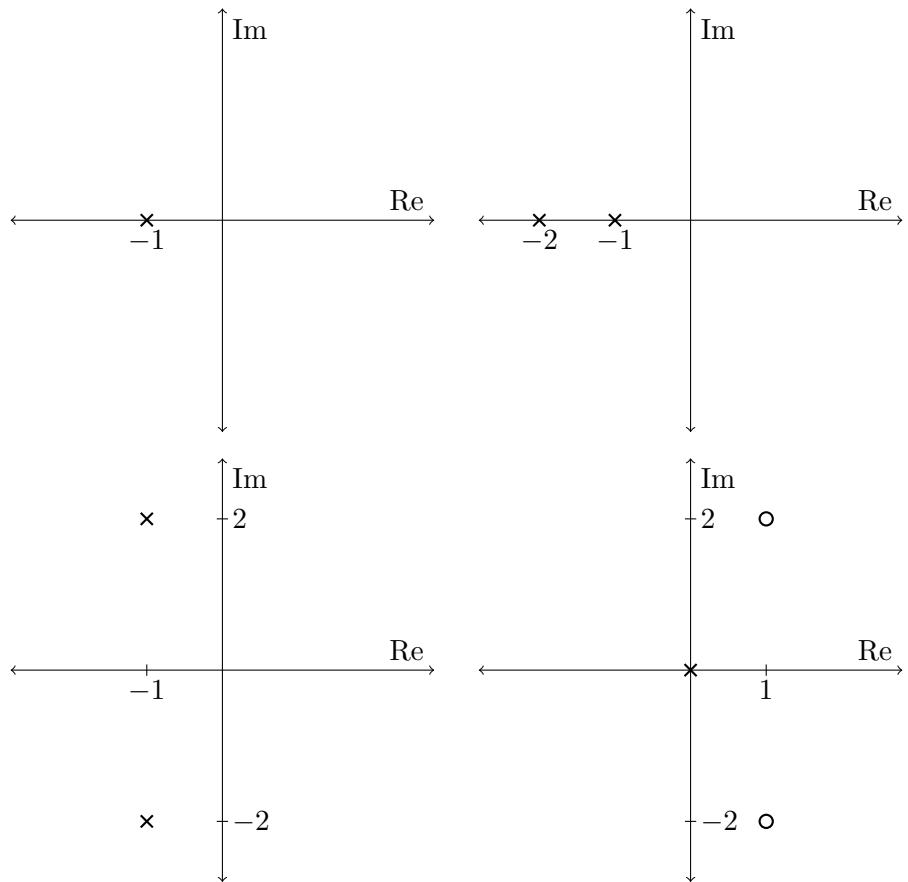


Figure 4.6: Top left: pole zero plot for the first order system  $x = y + D(y)$ . There is a single pole at  $-1$ . Top right: pole zero plot for the overdamped second order system  $x = 2y + 3D(y) + D^2(y)$  that has two real poles at  $-1$  and  $-2$ . Bottom left: pole zero plot for the underdamped second order system  $x = 5y + 2D(y) + D^2(y)$  that has two imaginary poles at  $-1 + 2j$  and  $-1 - 2j$ . The poles form a conjugate pair. Bottom right: pole zero plot for the equation  $D(y) = 5x - 2D(x) + D^2(x)$  that models a PID controller (2.2.7). The system has a single pole at the origin and two zeros at  $1 + 2j$  and  $1 - 2j$ .

and so the terms of the form  $\frac{A}{(s-\beta)^r}$  correspond with the transfer function of a regular system with impulse response  $\frac{A}{r!}t^{r-1}e^{\beta t}u(t)$ . Other signals with Laplace transform  $\frac{A}{(s-\beta)^r}$  are discarded because they do not correspond with the impulse response of a causal system. Thus,  $\sum_{\ell \in K} \frac{A_\ell}{(s-\beta_\ell)^{r_\ell}}$  is the transfer function of the regular system  $H_2$  with impulse response

$$h_2(t) = u(t) \sum_{\ell \in K} \frac{A_\ell}{r_\ell!} t^{r_\ell-1} e^{\beta_\ell t}.$$

Let  $K_r = \{\ell \in K ; \operatorname{Im} \beta_\ell = 0\}$  be the indices from  $K$  corresponding with the real poles, and let  $K_i = \{\ell \in K ; \operatorname{Im} \beta_\ell > 0\}$  be the indices corresponding with those poles with positive imaginary part. Because the imaginary poles occur in conjugate pairs the impulse response  $h_2$  can be written as

$$h_2(t) = u(t) \sum_{\ell \in K_r} \frac{A_\ell}{r_\ell!} t^{r_\ell-1} e^{\beta_\ell t} + u(t) \sum_{\ell \in K_i} \frac{t^{r_\ell-1}}{r_\ell!} (A_\ell e^{\beta_\ell t} + A_\ell^* e^{\beta_\ell^* t}).$$

The terms

$$\begin{aligned} A_\ell e^{\beta_\ell t} + A_\ell^* e^{\beta_\ell^* t} &= |A_\ell| e^{\operatorname{Re} \beta_\ell t} (e^{\operatorname{Im} \beta_\ell t + \angle A_\ell} + e^{-\operatorname{Im} \beta_\ell t - \angle A_\ell}) \\ &= 2 |A_\ell| e^{\operatorname{Re} \beta_\ell t} \cos(\operatorname{Im} \beta_\ell t + \angle A_\ell), \end{aligned}$$

and so, the impulse response is

$$h_2(t) = u(t) \sum_{\ell \in K_r} \frac{A_\ell}{r_\ell!} t^{r_\ell-1} e^{\beta_\ell t} + u(t) \sum_{\ell \in K_i} \frac{2 |A_\ell|}{r_\ell!} t^{r_\ell-1} e^{\operatorname{Re} \beta_\ell t} \cos(\operatorname{Im} \beta_\ell t + \angle A_\ell).$$

This expression can be simplified by putting

$$B_\ell = \begin{cases} \frac{A_\ell}{r_\ell!} & \operatorname{Im} \beta_\ell = 0 \\ 2 \frac{|A_\ell|}{r_\ell!} & \operatorname{Im} \beta_\ell > 0 \end{cases}$$

so that

$$h_2(t) = u(t) \sum_{\ell \in K_r \cup K_i} B_\ell t^{r_\ell-1} e^{\operatorname{Re} \beta_\ell t} \cos(\operatorname{Im} \beta_\ell t + \angle B_\ell). \quad (4.5.2)$$

Observe that the impulse response is a real valued signal (as expected).

The system  $H$  mapping  $x$  to  $y$  is the sum of the regular system  $H_2$  and nonregular system  $H_1$ , that is,

$$y = H(x) = H_1(x) + H_2(x).$$

Observe that  $H$  is regular only if the system  $H_1 = 0$ , that is, only if  $H_1$  maps all input signals to the signal  $x(t) = 0$  for all  $t \in \mathbb{R}$ . This occurs only when the polynomial  $p(s) = 0$ , that is, only when the number of poles exceeds

the number of zeros. The system  $H$  will be stable if both  $H_1$  and  $H_2$  are stable. Because the differentiator  $D^\ell$  is not stable (Exercise 1.7) the system  $H_1$  is stable if and only if the order of the polynomial  $p(s)$  is zero, that is, if  $p(s) = \gamma_0$  is a constant (potentially  $\gamma_0 = 0$ ). In this case  $H_1(x) = \gamma_0 T_0(x)$  is the identity system multiplied by a constant. The polynomial  $p(s)$  is a constant only when the order of the denominator polynomial is greater than or equal to the order of the numerator polynomial, that is, when the number of poles is greater than or equal to the number of zeros. The regular system  $H_2$  is stable if and only if its impulse response  $h_2$  is absolutely integrable. This occurs only when the terms  $e^{\operatorname{Re} \beta_\ell t}$  inside the sum (4.5.2) are decreasing as  $t \rightarrow \infty$ , that is, only if the real part of the poles  $\operatorname{Re} \beta_\ell$  are negative. Thus, the system  $H_2$  is stable if and only if the real part of the poles are strictly negative.

The stability of the system  $H$  can be immediately determined from its pole-zero plot. The system is stable if and only if:

1. the number of poles is greater than or equal to the number of zeros (there are at least as many crosses on the pole-zero plot as circles),
2. all of the poles (crosses) lie strictly in the left half plane.

The pole-zero plots in Figure 4.6 all represent stable systems with the exception of the plot on the bottom right (a PID controller). This system has two zeros and only one pole. The single pole is contained on the imaginary axis. It is not strictly in the left half plane.

#### 4.5.1 Two masses, a spring, and a damper

Consider the system involving two masses, a spring, and a damper in Figure 2.11. From (2.3.3), the equation relating the force applied to the first mass  $f$  and the position of the second mass  $p$  is

$$f = BD(p) + (M_1 + M_2)D^2(p) + \frac{BM_2}{K}D^3(p) + \frac{M_1 M_2}{K}D^4(p),$$

where  $B$  is the damping coefficient,  $K$  is the spring constant, and  $M_1$  and  $M_2$  are the masses. Taking Laplace transforms

$$\mathcal{L}(f) = s \left( B + (M_1 + M_2)s + \frac{BM_2}{K}s^2 + \frac{M_1 M_2}{K}s^3 \right) \mathcal{L}(p),$$

from which, we obtain the transfer function of a system  $H$  that maps  $f$  to  $p$ ,

$$\lambda(H) = \frac{\mathcal{L}(p)}{\mathcal{L}(f)} = \frac{1}{s \left( B + (M_1 + M_2)s + \frac{BM_2}{K}s^2 + \frac{M_1 M_2}{K}s^3 \right)}.$$

The system has no zeros and 4 poles. One of these poles always exists at the origin. The system is not stable because this pole is not strictly in the left half of the complex plane.

Consider the specific case when  $B = K = M_1 = M_2 = 1$ . Factorising the denominator polynomial gives

$$\lambda(H) = \frac{1}{s(s - \beta_1)(s - \beta_2)(s - \beta_2^*)},$$

where

$$\beta_1 = \frac{1}{3} \left( \gamma - \frac{5}{\gamma} - 1 \right) \approx -0.56984,$$

$$\beta_2 = \frac{1}{6} \left( \frac{5(1 + j\sqrt{3})}{\gamma} - (1 - j\sqrt{3})\gamma - \frac{1}{2} \right) \approx -0.21508 + 1.30714j,$$

and  $\gamma = \left(\frac{3\sqrt{69}-11}{2}\right)^{1/3}$ . Applying partial fractions (Exercise 4.8) gives

$$\lambda(H) = \frac{1}{s(s - \beta_1)(s - \beta_2)(s - \beta_2^*)} = \frac{A_0}{s} + \frac{A_1}{s - \beta_1} + \frac{A_2}{s - \beta_2} + \frac{A_2^*}{s - \beta_2^*},$$

where

$$A_0 = -\frac{1}{\beta_1|\beta_2|^2} = 1, \quad A_1 = \frac{1}{\beta_1|\beta_1 - \beta_2|^2} \approx -0.956611,$$

$$A_2 = \frac{1}{\beta_2(\beta_2 - \beta_1)(\beta_2 - \beta_2^*)} \approx -0.0216944 + 0.212084j.$$

From (4.5.2), the impulse response of  $H$  is

$$h(t) = u(t) \left( A_0 + A_1 e^{\beta_1 t} + 2|A_2| e^{\operatorname{Re} \beta_2 t} \cos(\operatorname{Im} \beta_2 t + \angle A_2) \right).$$

This impulse response is plotted in Figure 4.7. Observe that  $h$  is not absolutely integrable and the system is not stable. The impulse response  $h(t)$  does not converge to zero as  $t \rightarrow \infty$  and correspondingly the mass  $M_2$  does not come to rest at position zero in Figure 4.7. In the figure it is assumed that the spring is at equilibrium when the two masses are  $d = 1$  apart. From (2.3.1), the position of mass  $M_1$  is given by the signal  $p_1 = g - d$  where  $g = h + M_2 D^2(h)$ .

### 4.5.2 Direct current motors

Recall the direct current (DC) motor from Figure 2.13 described by the differential equation from (2.4.1),

$$v = \left( \frac{RB}{K_\tau} + K_b \right) D(\theta) + \frac{RJ}{K_\tau} D^2(\theta),$$

where  $v$  is the input voltage signal and  $\theta$  is a signal representing the angle of the motor. The constants  $R, B, K_\tau, K_b$ , and  $J$  are related to components of

Figure 4.7: Impulse response of the system with two masses, a spring, and a damper, where  $B = K = M_1 = M_2 = 1$ .

the motor as described in Section 2.4. To simplify the differential equation put  $a = \frac{RB}{K\tau} + K_b$  and  $b = \frac{RJ}{K\tau}$  and the equation becomes

$$v = aD(\theta) + bD^2(\theta).$$

Taking Laplace transforms on both sides of this equation gives the transfer function of a system  $H$  that maps input voltage  $v$  to motor angle  $\theta$ ,

$$\lambda(H) = \frac{1}{s(a + bs)}.$$

This system has no zeros and two poles. One pole is at  $-\frac{a}{b}$  and the other is at the origin. The system is not stable because the pole at the origin is not strictly in the left half of the complex plane.

Applying partial fractions we find that

$$\lambda(H) = \frac{1}{as} - \frac{1}{a(s - \beta)}, \quad (4.5.3)$$

where  $\beta = -\frac{a}{b}$ . Using (4.0.6), the impulse response of  $H$  is

$$h(t) = \frac{1}{a}u(t)(1 - e^{\beta t}). \quad (4.5.4)$$

Figure 4.8: Impulse response (top) and step response (bottom) of a DC motor with constants  $K_b = \frac{1}{4}$ ,  $K_\tau = 8$  and  $B = R = J = 1$ .

Other signals with Laplace transform (4.5.3) are discarded because they do not lead to a causal system. The step response  $H(u)$  is obtained by applying the integrator system  $I_\infty$  to the impulse response, that is

$$H(u) = I_\infty(h) = \frac{1}{a\beta} u(t)(\beta t + e^{\beta t} - 1).$$

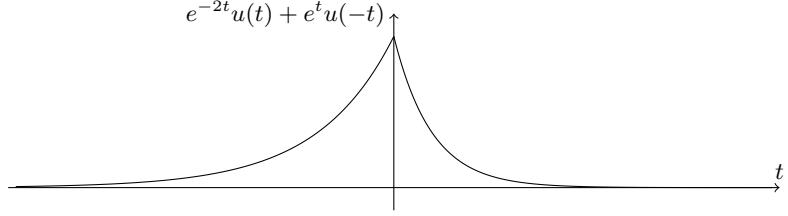
The impulse response and step response are plotted in Figure 4.8 when  $K_b = \frac{1}{8}$ ,  $K_\tau = 8$  and  $B = R = 1$  and  $J = 2$  so that  $a = \frac{1}{4}$ ,  $b = \frac{1}{4}$  and  $\beta = -1$ .

## 4.6 Exercises

### 4.1. Sketch the signal

$$x(t) = e^{-2t}u(t) + e^tu(-t)$$

where  $u(t)$  is the step function. Find the Laplace transform of  $x(t)$  and the corresponding region of convergence (ROC). Sketch the region of convergence on the complex plane. **Solution:**



The Laplace transform of  $e^{-2t}u(t)$  is

$$\begin{aligned}\mathcal{L}(e^{-2t}u(t), s) &= \int_{-\infty}^{\infty} e^{-2t}u(t)e^{-st}dt \\ &= \int_0^{\infty} e^{-(s+2)t}dt \\ &= -\frac{1}{s+2}, \quad \text{Re}(s) > -2\end{aligned}$$

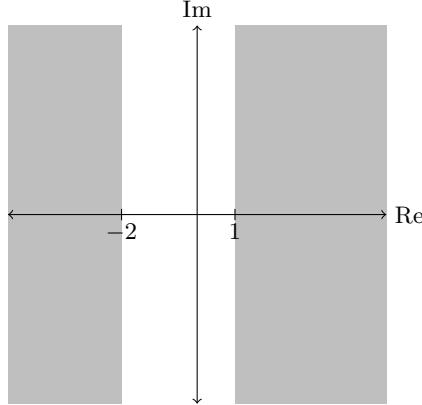
and the Laplace transform of  $e^tu(-t)$  is

$$\begin{aligned}\mathcal{L}(e^tu(-t), s) &= \int_{-\infty}^0 e^{-(s-1)t}dt \\ &= -\frac{1}{s-1}, \quad \text{Re}(s) < 1.\end{aligned}$$

Thus, the Laplace transform of  $e^{-2t}u(t) + e^tu(-t)$  is

$$\mathcal{L}(e^{-2t}u(t) + e^tu(-t), s) = -\frac{1}{s+2} - \frac{1}{s-1}, \quad -2 < \text{Re}(s) < 1$$

and the region of convergence (ROC) is the subset of the complex plane satisfying  $-2 < \text{Re}(s) < 1$ . The unshaded region in the plot below depicts the ROC.



- 4.2. Find the Laplace transform of the signal  $t^n u(t)$  where  $n \geq 0$  is an integer. **Solution:** We have

$$\mathcal{L}(t^n u(t)) = \int_{-\infty}^{\infty} t^n u(t) e^{-st} dt = \int_0^{\infty} t^n e^{-st} dt.$$

Integration by parts gives the indefinite integral

$$\int t^n e^{-st} dt = \frac{t^n}{s} e^{-st} + \frac{n}{s} \int t^{n-1} e^{-st} dt.$$

So, when  $\operatorname{Re}(s) > 0$ ,

$$\begin{aligned}\mathcal{L}(t^n u(t)) &= \lim_{t \rightarrow 0} \frac{t^n}{s} e^{-st} - \lim_{t \rightarrow \infty} \frac{t^n}{s} e^{-st} + \frac{n}{s} \int_0^\infty t^{n-1} e^{-st} dt \\ &= \frac{n}{s} \mathcal{L}(t^{n-1} u(t)),\end{aligned}$$

since both limits converge to zero. Unravelling the above recursive equation gives

$$\mathcal{L}(t^n u(t)) = \frac{n}{s} \times \frac{n-1}{s} \times \cdots \times \frac{1}{s} \times \mathcal{L}(u(t)) = \frac{n!}{s^{n+1}}, \quad \operatorname{Re}(s) > 0,$$

since  $\mathcal{L}(u(t)) = \frac{1}{s}$  when  $\operatorname{Re}(s) > 0$ .

- 4.3. Let  $n \geq 0$  be an integer. Show that the Laplace transform of the signal  $t^n u(-t)$  is the same as the Laplace transform of the signal  $t^n u(t)$ , but with a different region of convergence. **Solution:** The working is identical to question 4.2, but now the limits converge only when  $\operatorname{Re}(s) < 0$ , and this is the region of convergence.
- 4.4. Show that equation (4.1.5) on page 61 holds when the system  $H$  is the differentiator  $D^k$  or the time shifter  $T_\tau$ . **Solution:** Put  $y = T_\tau(x) = x(t-\tau)$ . Taking Laplace transforms

$$\begin{aligned}\mathcal{L}(y) &= \mathcal{L}(T_\tau(x)) \\ &= \int_{-\infty}^{\infty} T_\tau(x, t) e^{-st} dt \\ &= \int_{-\infty}^{\infty} x(t-\tau) e^{-st} dt \\ &= \int_{-\infty}^{\infty} x(\kappa) e^{-s(\kappa+\tau)} d\kappa \quad (\text{ch. vars. } \kappa = t - \tau) \\ &= e^{-s\tau} \int_{-\infty}^{\infty} x(\kappa) e^{-s\kappa} d\kappa \\ &= e^{-s\tau} \mathcal{L}(x) \\ &= \lambda(T_\tau) \mathcal{L}(x)\end{aligned}$$

as required.

We now show the result for the differentiator  $D$ . The argument we use is based on that of Rudin [1986, page 179]. We have just shown that the Laplace transform of the time shifted  $T_{-\tau}(x, t) = x(t+\tau)$  is  $e^{\tau s} \mathcal{L}(x, s)$ . Since the Laplace transform is linear

$$\mathcal{L}\left(\frac{T_{-\tau}(x) - x}{\tau}\right) = \frac{e^{\tau s} - 1}{\tau} \mathcal{L}(x).$$

We now consider what happens to both side of this equation as  $\tau \rightarrow 0$ . Application of L'Hopital's rule shows that

$$\lim_{\tau \rightarrow 0} \frac{e^{\tau s} - 1}{\tau} = s,$$

and so, the right hand side satisfies

$$\lim_{\tau \rightarrow 0} \frac{e^{\tau s} - 1}{\tau} \mathcal{L}(x) = s \mathcal{L}(x).$$

On the left hand side we have

$$\lim_{\tau \rightarrow 0} \mathcal{L}\left(\frac{T_{-\tau}(x) - x}{\tau}\right) = \lim_{\tau \rightarrow 0} \int_{-\infty}^{\infty} \frac{x(t+\tau) - x(t)}{\tau} e^{-st} dt.$$

Observe that

$$\lim_{\tau \rightarrow 0} \frac{x(t + \tau) - x(t)}{\tau} = D(x)$$

by definition of the differentiator  $D$ . Lebesgue's dominated convergence theorem [Rudin, 1986, page 26] can be used to justify exchanging limits and integration so that

$$\begin{aligned} \lim_{\tau \rightarrow 0} \int_{-\infty}^{\infty} \frac{x(t + \tau) - x(t)}{\tau} e^{-st} dt &= \int_{-\infty}^{\infty} \lim_{\tau \rightarrow 0} \frac{x(t + \tau) - x(t)}{\tau} e^{-st} dt \\ &= \int_{-\infty}^{\infty} D(x, t) e^{-st} dt = \mathcal{L}(D(x)). \end{aligned}$$

We now have  $\mathcal{L}(D(x)) = s\mathcal{L}(x) = \lambda(D)\mathcal{L}(x)$  as required.

The result follows for the  $k$ th differentiator  $D^k$  because

$$\mathcal{L}(D^k(y)) = \mathcal{L}(D(D^{k-1}(y))) = s\mathcal{L}(D^{k-1}(y))$$

and unravelling this recursion gives

$$\mathcal{L}(D^k(y)) = \underbrace{s \times s \times \cdots \times s}_{k-1 \text{ times}} \times \mathcal{L}(D(y)) = s^k \mathcal{L}(y) = \lambda(D^k)\mathcal{L}(y)$$

as required.

- 4.5. What is the transfer function of the integrator system  $I_{\infty}$  and what is its region of convergence? **Solution:** We have

$$I_{\infty}(e^{st}, t) = \int_{-\infty}^t e^{s\tau} d\tau = \frac{e^{st}}{s} - \lim_{a \rightarrow -\infty} \frac{e^{sa}}{s}$$

and the limit exists only when  $\operatorname{Re} s > 0$  and in this case it is zero. So

$$I_{\infty}(e^{st}, t) = \frac{1}{s} e^{st} \quad \operatorname{Re} s > 0$$

and  $\lambda(I_{\infty}, s) = \frac{1}{s}$ .

- 4.6. By partial fractions, or otherwise, assert that

$$\frac{as}{s+b} = a - \frac{ab}{s+b}$$

**Solution:** Adding and subtracting  $ab$  from the numerator

$$\frac{as + ab - ab}{s+b} = \frac{a(s+b) - ab}{s+b} = \frac{a(s+b)}{s+b} - \frac{ab}{s+b} = a - \frac{ab}{s+b}$$

- 4.7. By partial fractions, or otherwise, assert that

$$\frac{s+c}{(s+a)(s+b)} = \frac{a-c}{(a-b)(s+a)} + \frac{c-b}{(a-b)(s+b)}$$

**Solution:** Hypothesise the solution

$$\frac{s+c}{(s+a)(s+b)} = \frac{A}{s+a} + \frac{B}{s+b}.$$

Multiplying both sides by  $(s + a)(s + b)$ ,

$$s + c = A(s + b) + B(s + a).$$

Putting  $s = -a$  gives  $c - a = A(b - a)$ , and putting  $s = -b$  gives  $c - b = B(a - b)$ , and so,

$$\frac{s + c}{(s + a)(s + b)} = \frac{a - c}{(a - b)(s + a)} + \frac{c - b}{(a - b)(s + b)}.$$

4.8. By partial fractions, or otherwise, assert that

$$\frac{1}{s(s - a)(s - b)(s - b^*)} = \frac{A_0}{s} + \frac{A_1}{s - a} + \frac{A_2}{s - b} + \frac{A_2^*}{s - b^*}$$

where  $a, b \in \mathbb{C}$  and  $\operatorname{Re}(b) \neq 0$  and

$$A_0 = -\frac{1}{a|b|^2}, \quad A_1 = \frac{1}{a|a - b|^2}, \quad A_2 = \frac{1}{b(b - a)(b - b^*)}.$$

You might wish to check your solution using a symbolic programming language (for example Sage, Mathematica, or Maple). **Solution:** The mathematica command

```
Apart[1/s/(s - a)/(s - b)/(s - c), s]
```

returns the equation

$$\frac{1}{s(s - a)(s - b)(s - c)} = \frac{A_0}{s} + \frac{A_1}{s - a} + \frac{A_2}{s - b} + \frac{A_3}{s - b^*} - \frac{1}{abcs} + \frac{1}{abc} + \frac{1}{a} + \frac{1}{b} + \frac{1}{c} + \frac{1}{(s - c)}$$

where

$$A_0 = -\frac{1}{abc}, \quad A_1 = \frac{1}{a(a - b)(a - c)}, \\ A_2 = \frac{1}{b(b - a)(b - c)}, \quad A_3 = \frac{1}{c(c - a)(c - b)}.$$

Setting  $c = b^*$  gives

$$A_0 = -\frac{1}{a|b|^2}, \quad A_1 = \frac{1}{a|a - b|^2}, \\ A_2 = \frac{1}{b(b - a)(b - b^*)}, \quad A_3 = \frac{1}{b^*(b^* - a)(b^* - b)} = A_2^*$$

as required.

4.9. Let

$$\mathcal{L}(y) = \frac{2s + 1}{s^2 + s - 2}$$

be the Laplace transform of a signal  $y$ . By partial fractions, or otherwise, find all possible signals  $y$  and their regions of convergence.

**Solution:** Factorise the polynomial on the denominator

$$\frac{2s + 1}{(s + 2)(s - 1)}.$$

Adding and subtracting  $s - 1$  on the numerator

$$\begin{aligned}\frac{2s+1+(s-1)-(s-1)}{(s+2)(s-1)} &= \frac{s-1}{(s-1)(s+2)} + \frac{s+2}{(s-1)(s+2)} \\ &= \frac{1}{s+2} + \frac{1}{s-1}.\end{aligned}$$

There are two time domain signals with Laplace transform  $\frac{1}{s+2}$ ,

$$e^{-2t}u(t), \operatorname{Re}(s) > -2 \quad \text{and} \quad -e^{-2t}u(-t), \operatorname{Re}(s) < -2,$$

and two time domain signals with Laplace transform  $-\frac{1}{s-1}$ ,

$$e^tu(t), \operatorname{Re}(s) > 1 \quad \text{and} \quad -e^tu(-t), \operatorname{Re}(s) < 1.$$

There are three possible signals with nonempty regions of convergence

$$y(t) = e^{-2t}u(t) - e^tu(-t) \quad -2 < \operatorname{Re}(s) < 1,$$

$$y(t) = e^{-2t}u(t) + e^tu(t) \quad 1 < \operatorname{Re}(s),$$

$$y(t) = -e^{-2t}u(-t) - e^tu(-t) \quad \operatorname{Re}(s) < -2.$$

- 4.10. Let  $x$  be a signal with region of convergence  $R$ . Show that the time scaled signal  $x(\alpha t)$  with  $\alpha \neq 0$  satisfies equation (4.0.7) on page 58.

**Solution:** First consider when  $\alpha = -1$  so that  $x(-t)$  is the reflection of the signal  $x$  in time (see Section 1.3). We have

$$\begin{aligned}\mathcal{L}(x(-t), s) &= \int_{-\infty}^{\infty} x(-t)e^{-st}dt \\ &= - \int_{\infty}^{-\infty} x(\tau)e^{s\tau}d\tau \quad (\text{change variable } \tau = -t) \\ &= \int_{-\infty}^{\infty} x(\tau)e^{s\tau}d\tau = \mathcal{L}(x, -s) \quad \operatorname{Re}(-s) \in R.\end{aligned}$$

This special case is called the **time reversal property**. Now, when  $\alpha > 0$ ,

$$\begin{aligned}\mathcal{L}(x(\alpha t), s) &= \int_{-\infty}^{\infty} x(\alpha t)e^{-st}dt \\ &= \frac{1}{\alpha} \int_{\infty}^{\infty} x(\tau)e^{-s\tau/\alpha}d\tau \quad (\text{change variable } \tau = \alpha t) \\ &= \frac{1}{\alpha} \mathcal{L}(x, s/\alpha) \quad \operatorname{Re}(s/\alpha) \in R.\end{aligned}$$

Combining this with the time reversal property we obtain

$$\mathcal{L}(x(\alpha t), s) = \frac{1}{|\alpha|} \mathcal{L}(x, s/\alpha), \quad a \neq 0, \operatorname{Re}(s/\alpha) \in R.$$

as required.

- 4.11. Consider the active electrical circuit from Figure 2.8 described by the differential equation from (2.2.3). Derive the transfer function of this system. Find an explicit system  $H$  that maps the input voltage  $x$  to the output voltage  $y$ . State whether this system is stable and/or regular. **Solution:** The differential equation modelling the circuit is

$$-\frac{x}{R_1} - C_1 D(x) = \frac{y}{R_2} + C_2 D(y),$$

and taking Laplace transforms on both sides of this equation

$$\mathcal{L}y = -\frac{\frac{1}{R_1} + C_1 s}{\frac{1}{R_2} + C_2 s} \mathcal{L}(x) = -\frac{\alpha + \gamma s}{\beta + s}$$

where  $\alpha = \frac{1}{R_1 C_2}$ ,  $\beta = \frac{1}{R_2 C_2}$ , and  $\gamma = \frac{C_1}{C_2}$ . The transfer function of the system mapping  $x$  to  $y$  is correspondingly

$$\lambda(H) = -\frac{\alpha + \gamma s}{\beta + s} = -\frac{\alpha}{\beta + s} - \frac{\gamma s}{\beta + s}$$

Applying partial fraction (as in Exercise 4.6) to the second term gives

$$\lambda(H) = -\frac{\alpha + \gamma \beta}{\beta + s} - \gamma$$

The first term  $-\frac{\alpha + \gamma \beta}{\beta + s}$  corresponds with a regular system, say  $H_2$ , having impulse response

$$h_2 = -(\alpha + \gamma \beta)u(t)e^{-\beta t}$$

by using the Laplace transform pair from (4.0.6) with the integer  $n = 0$ . The term  $-\gamma$  correspond with the system  $H_1 = \gamma T_0$ , i.e., the identity system multiplied by  $-\gamma$ . The system  $H$  that describes the mapping between input voltage  $x$  and output voltage  $y$  is thus

$$H(x) = H_1(x) + H_2(x) = -\gamma x + h_2 * x.$$

The system is not regular because the  $H_1$  is not regular. The system is stable because  $H_1$  is stable and  $H_2$  is stable because the impulse response  $h_2$  is absolutely integrable since  $\beta = \frac{1}{R_2 C_2} > 0$ . Equivalently the system is not regular because the transfer function does not have more poles than zero, and the system is stable because the transfer function has at least as many poles as zeros (equal in this case), and because all the poles lie strictly in the left half plane.

- 4.12. Given the mass spring damper system described by (4.4.1), find the position signal  $p$  given that the force signal

$$f(t) = \Pi(t - \frac{1}{2}) = \begin{cases} 1 & 0 < t \leq 1 \\ 0 & \text{otherwise} \end{cases}$$

is the rectangular function time shifted by  $\frac{1}{2}$ . Consider three cases:

- (a)  $M = 1$ ,  $K = \frac{\pi^2}{4}$  and  $B = \frac{\pi}{3}$ ,
- (b)  $M = 1$ ,  $K = \frac{\pi^2}{4}$  and  $B = \pi$ ,
- (c)  $M = 1$ ,  $K = \frac{\pi^2}{4}$  and  $B = 2\pi$ ,

Plot the solution in each case, and comment on whether the system is underdamped, overdamped, or critically damped. **Solution:** Observe that the input force signal can be written as the sum of the step function  $u$  and its negated time-shift, that is,

$$f(t) = u(t) - u(t - 1) = u(t) - T_1(u, t)$$

and so, the response of the linear, time invariant system  $H$  modelling the mass spring damper to input force signal  $f$  is

$$H(f) = H(u - T_1(u)) = H(u) - T_1(H(u)),$$

and so,  $H(f, t) = H(u, t) - H(u, t - 1)$ , where  $H(u)$  is the step response of the system. The step responses are described in Section 4.4. As described in Section 4.4, the system is underdamped when  $B = \frac{\pi}{3}$ , critically damped when  $B = \pi$  and overdamped when  $B = 2\pi$ .

## Chapter 5

# The Fourier transform

The **Fourier transform** of an absolutely integrable signal  $x$  is defined as

$$\mathcal{F}(x) = \int_{-\infty}^{\infty} x(t)e^{-j2\pi ft}dt. \quad (5.0.1)$$

The Fourier transform is a function of the real number  $f$ . We indicate its value at  $f$  by  $\mathcal{F}(x)(f)$  or  $\mathcal{F}(x, f)$ . For example, the rectangular pulse  $\Pi(t)$  from (1.1.2) is absolutely integrable and has Fourier transform

$$\begin{aligned} \mathcal{F}(\Pi) &= \int_{-\infty}^{\infty} \Pi(t)e^{-j2\pi ft}dt \\ &= \int_{-1/2}^{1/2} e^{-j2\pi ft}dt \\ &= \frac{e^{j\pi f} - e^{-j\pi f}}{j2\pi f} \\ &= \frac{\sin(\pi f)}{\pi f} = \text{sinc}(f). \end{aligned} \quad (5.0.2)$$

The sinc function is plotted in Figure 5.1.

The Fourier transform is closely related to the Laplace transform because

$$\mathcal{F}(x, f) = \mathcal{L}(x, j2\pi f)$$

for those signals  $x$  with region of convergence containing the imaginary axis, that is, for absolutely integrable  $x$ . The Fourier transform inherits the properties of the Laplace transform that were described in Section 4.1. For example, if  $H$  is a stable regular system with absolutely integrable impulse response  $h$  having Fourier transform  $\mathcal{F}(h)$ , then the spectrum of  $H$  satisfies

$$\Lambda(H, f) = \lambda(H, j2\pi f) = \mathcal{L}(h, j2\pi f) = \mathcal{F}(h, f)$$

That is, the spectrum of a stable regular system is given by the Fourier transform of its impulse response. Like the Laplace transform, the Fourier transform obeys the **convolution theorem** (4.1.4), that is,

$$\mathcal{F}(x * y) = \mathcal{F}(x)\mathcal{F}(y) \quad (5.0.3)$$

when each of the signals  $x$ ,  $y$ , and  $x * y$  are absolutely integrable. In words: the Fourier transform of a convolution of signals is given by the multiplication of the Fourier transforms of those signals.

It follows from (4.1.5) that if  $H$  is a regular system with spectrum  $\Lambda(H)$  and if  $x$  is a signal with Fourier transform  $\mathcal{F}(x)$ , then the signal  $y = H(x)$  has Fourier transform

$$\mathcal{F}(y) = \Lambda(H)\mathcal{F}(x).$$

This property also holds for the differentiator system  $D$  and the time shifter system  $T_\tau$  (Exercise 4.1.5). From (4.1.1) and (4.1.2) the spectrum of  $T_\tau$  and the  $k$ th differentiator  $D^k$  satisfy

$$\Lambda(T_\tau) = e^{-j2\pi f\tau}, \quad \Lambda(D^k) = (j2\pi f)^k$$

from which we obtain the **time shift property**,

$$\mathcal{F}(T_\tau(x)) = \Lambda(T_\tau)\mathcal{F}(x) = e^{-j2\pi f\tau}\mathcal{F}(x),$$

and the **differentiation property**,

$$\mathcal{F}(D^k(x)) = \Lambda(D^k)\mathcal{F}(x) = (j2\pi f)^k\mathcal{F}(x),$$

of the Fourier transform. These results motivate assigning the following Fourier transforms to the delta “function” and its derivatives

$$\mathcal{F}(\delta) = 1, \quad \mathcal{L}(\delta^k) = (j2\pi f)^k. \quad (5.0.4)$$

These conventions are common in the literature [Oppenheim et al., 1996].

Similarly to the Laplace transform (4.0.5), the Fourier transform obeys a **frequency shift rule** that relates the transform of a signal  $x(t)$  to that of the signal  $e^{2\pi j\gamma ft}x(t)$  where  $\gamma \in \mathbb{R}$ . From (4.0.5), the frequency shift rule asserts that

$$\mathcal{F}(e^{2\pi j\gamma t}x(t), f) = \mathcal{F}(x, f - \gamma). \quad (5.0.5)$$

Since  $\cos(2\pi\gamma t) = \frac{1}{2}e^{2\pi j\gamma t} + \frac{1}{2}e^{-2\pi j\gamma t}$  we also have

$$\mathcal{F}(\cos(2\pi\gamma t)x(t), f) = \frac{1}{2}\mathcal{F}(x, f - \gamma) + \frac{1}{2}\mathcal{F}(x, f + \gamma). \quad (5.0.6)$$

This is sometimes called the **modulation property** of the Fourier transform [Papoulis, 1977, page 61]. This property is of particular importance in communications engineering [Proakis, 2007].

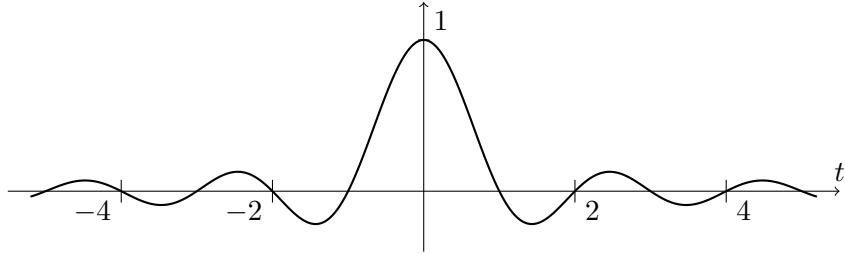


Figure 5.1: The **sinc function**  $\text{sinc}(t) = \frac{\sin(\pi t)}{\pi t}$ .

Like the Laplace transform (4.0.7), the Fourier obeys a **time scaling property**. If  $x$  is an absolutely integrable signal then the time scaled signal  $x(\alpha t)$  with  $\alpha \neq 0$  has Fourier transform

$$\mathcal{F}(x(\alpha t), f) = \frac{1}{|\alpha|} \mathcal{F}(x, f/\alpha). \quad (5.0.7)$$

## 5.1 The inverse transform and the Plancherel theorem

Given a signal  $x$  we will often denote its Fourier transform by  $\hat{x} = \mathcal{F}(x)$ . Observe that  $\hat{x}$ , like  $x$ , is a function that maps a real number to a complex number. Thus, the Fourier transform  $\hat{x}$  is a **signal** with independent variable representing **frequency**. It is usual to call  $x$  the **time domain** representation of the signal and  $\hat{x}$  the **frequency domain** representation. If  $\hat{x}$  is absolutely integrable, then  $x$  can be recovered using the **inverse Fourier transform**

$$x(t) = \mathcal{F}^{-1}(\hat{x}) = \int_{-\infty}^{\infty} \hat{x}(f) e^{j2\pi ft} df. \quad (5.1.1)$$

For example, let  $\hat{x} = \mathcal{F}(x) = \Pi$  be the rectangular pulse. By working analogous to that from (5.0.2),

$$x(t) = \int_{-\infty}^{\infty} \Pi(f) e^{j2\pi ft} df = \text{sinc}(-t) = \text{sinc}(t).$$

We are lead to the conclusion that the Fourier transform of sinc is the rectangular pulse  $\Pi$ .

The rectangular pulse  $\Pi$  is finite in time and absolutely integrable. The sinc function is not absolutely integrable (Exercise 5.3). Because of this the integral equation that we have used to define the Fourier transform (5.0.1) cannot be directly applied to the sinc function. Although sinc is not absolutely integrable, it is square integrable (Exercise 5.3). It happens that all square integrable signals can be assigned a Fourier transform by interpreting

the integral in (5.0.1) as what is called its **Cauchy principal value**. That is, for  $x$  a square integrable signal, we assign the Fourier transform

$$\hat{x} = \mathcal{F}(x) = \lim_{T \rightarrow \infty} \int_{-T}^T x(t)e^{-j2\pi ft} dt.$$

This Fourier transform  $\hat{x}$  is itself a square integrable signal and the original time domain signal  $x$  can be recovered almost everywhere by taking the Cauchy principal value of the integral in (5.1.1), that is,

$$x = \mathcal{F}^{-1}(\hat{x}) = \lim_{T \rightarrow \infty} \int_{-T}^T \hat{x}(f)e^{j2\pi ft} df \quad \text{a.e..}$$

Infact, the energy of  $x$  and its Fourier transform  $\hat{x}$  are the same, that is,

$$\|x\|_2^2 = \int_{-\infty}^{\infty} |x(t)|^2 dt = \int_{-\infty}^{\infty} |\hat{x}(f)|^2 df = \|\hat{x}\|_2^2. \quad (5.1.2)$$

These results are known as the **Plancherel theorem** [Rudin, 1986, Th. 9.13]. The equality of energies in (5.1.2) is often called **Parseval's identity**. For our purposes it will suffice to remember only that the Fourier transform of the sinc function is the rectangular pulse  $\Pi$ .

A result more general than (5.1.2) holds. If  $x$  and  $y$  are square integrable signals with Fourier transforms  $\hat{x}$  and  $\hat{y}$ , then

$$\int_{-\infty}^{\infty} x(t)y^*(t)dt = \int_{-\infty}^{\infty} \hat{x}(f)\hat{y}^*(f)df \quad (5.1.3)$$

where the superscript \* denotes the complex conjugate. One obtains (5.1.2) by putting  $y = x$  in (5.1.3). This more general result often also goes by the name of Parseval's identity.

Let  $x$  be a signal with Fourier transform

$$\mathcal{F}(x, f) = \int_{-\infty}^{\infty} x(\tau)e^{-j2\pi f\tau} d\tau.$$

Evaluating the Fourier transform at  $-t$  we find that

$$\mathcal{F}(x, -t) = \int_{-\infty}^{\infty} x(\tau)e^{j2\pi t\tau} d\tau = \mathcal{F}^{-1}(x, t). \quad (5.1.4)$$

This is the called the **duality** property of the Fourier transform. In words, if  $\hat{x}$  is the Fourier transform of  $x$ , then  $x$  is the Fourier transform of  $\hat{x}$  reflected in time.

## 5.2 Analogue filters

For many engineering purposes it is desirable to construct systems that will **pass** (have little affect on) a complex exponential signal  $e^{j2\pi ft}$  for certain frequencies  $f$ , but will **reject** (significantly attenuate) these signals for other frequencies. Such systems are called **filters**. Those frequencies that the filter intends to pass unaffected are said to be in the **pass band** and those frequencies that the filter intends to reject are said to be in the **stop band**.

For example, an **ideal lowpass filter** with **cutoff frequency**  $c$  is the system  $L_c$  with spectrum

$$\Lambda(L_c) = \begin{cases} 1 & -c < f \leq c \\ 0 & \text{otherwise} \end{cases} = \Pi\left(\frac{f}{2c}\right).$$

Applying the inverse Fourier transform to  $\Pi(\frac{f}{2c})$  gives

$$\int_{-\infty}^{\infty} \Pi\left(\frac{f}{2c}\right) e^{j2\pi t f} df = \int_{-c}^c e^{j2\pi t f} df = \frac{\sin(2c\pi t)}{\pi t} = 2c \operatorname{sinc}(2ct).$$

We conclude that the ideal lowpass filter  $L_c$  is a regular linear time invariant system with impulse response  $2c \operatorname{sinc}(2ct)$ .

An **ideal highpass filter** with cutoff frequency  $c$  is given by the linear combination  $T_0 - L_c$  where  $T_0$  is the identity system. The spectrum is

$$\Lambda(T_0 + L_c) = \Lambda(T_0) + \Lambda(L_c) = 1 - \Pi\left(\frac{f}{2c}\right) = \begin{cases} 0 & -c < f \leq c \\ 1 & \text{otherwise.} \end{cases}$$

This ideal highpass filter is not regular because the system  $T_0$  is not regular. The system does not have an impulse response. Nevertheless, it is common to represent one by  $\delta(t) - 2c \operatorname{sinc}(2ct)$  using the delta function as described in Section 3.1.

An **ideal bandpass filter** with upper cutoff frequency  $u$  and lower cutoff frequency  $\ell$  is given by the linear combination  $L_u - L_\ell$ . The spectrum is

$$\Lambda(L_u - L_\ell) = \Pi\left(\frac{f}{2u}\right) - \Pi\left(\frac{f}{2\ell}\right) = \begin{cases} 1 & -u < f \leq -\ell \\ 1 & u < f \leq \ell \\ 0 & \text{otherwise.} \end{cases}$$

It follows that the ideal bandpass filter has impulse response  $2u \operatorname{sinc}(2ut) - 2\ell \operatorname{sinc}(2\ell t)$ . The spectrum and impulse response of the ideal lowpass, highpass, and bandpass filters are plotted in Figure 5.2.

Ideal filters are not realisable in practice. One reason for this is that they are not causal because the sinc function is unbounded in time. We now describe a popular practical low-pass filter discovered by Butterworth

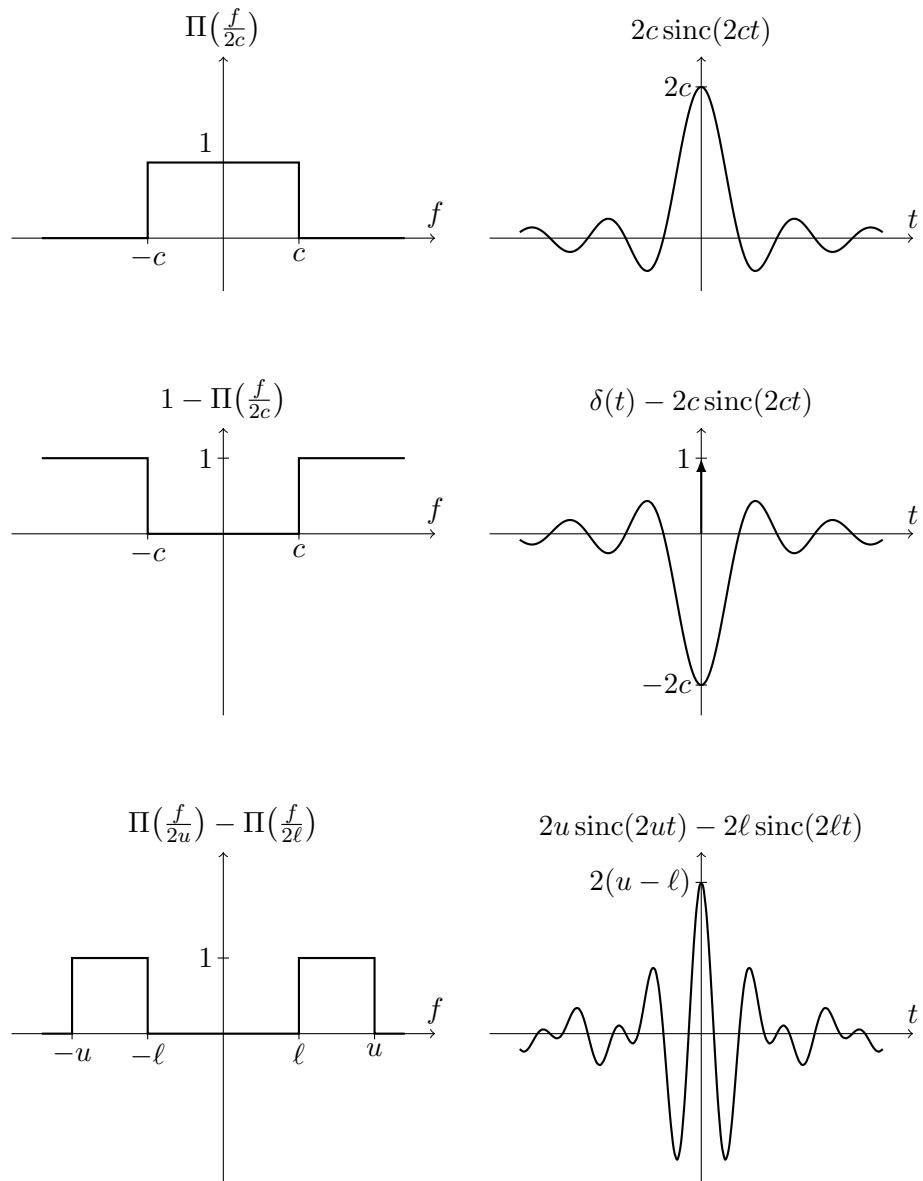


Figure 5.2: Spectrum and impulse response of the ideal lowpass filter  $L_c$  (top), the ideal highpass filter  $T_0 - L_c$  (middle), and the ideal bandpass filter  $L_u - L_\ell$  (bottom). The ideal highpass filter is not regular and does not have an impulse response. We plot the ‘pretend’ impulse response using the delta function described in Section 3.1.

[1930]. A **normalised low pass Butterworth filter** of order  $m$ , denoted by  $B_m$ , has transfer function

$$\lambda(B_m) = \frac{1}{\prod_{i=1}^m (\frac{s}{2\pi} - \beta_i)} = \frac{(2\pi)^m}{\prod_{i=1}^m (s - 2\pi\beta_i)},$$

where  $\beta_1, \dots, \beta_m$  are the roots of the polynomial  $s^{2m} + (-1)^m$  that lie strictly in the left half of the complex plane (have negative real part). It is convenient to precisely define these roots as

$$\beta_k = \begin{cases} \exp(j\frac{\pi}{2}(1 + \frac{2k-1}{m})), & k = 1, \dots, m \\ \exp(j\frac{\pi}{2}(1 - \frac{2k-1}{m})), & k = m+1, \dots, 2m \end{cases}$$

or equivalently

$$\beta_k = \begin{cases} j \cos(\frac{\pi(2k-1)}{2m}) - \sin(\frac{\pi(2k-1)}{2m}), & k = 1, \dots, m \\ j \cos(\frac{\pi(2k-1)}{2m}) + \sin(\frac{\pi(2k-1)}{2m}), & k = m+1, \dots, 2m. \end{cases}$$

The roots are plotted in Figure 5.3. Observe that the roots  $\beta_{m+1}, \dots, \beta_{2m}$  are given by negating the real parts of  $\beta_1, \dots, \beta_m$ , that is,  $\beta_{m+i} = j(\beta_i/j)^*$ .

The spectrum of  $B_m$  is

$$\Lambda(B_m) = \frac{1}{\prod_{i=1}^m (jf - \beta_i)}.$$

The squared magnitude of the polynomial on the denominator is

$$\begin{aligned} \left| \prod_{i=1}^m (jf - \beta_i) \right|^2 &= \left( \prod_{i=1}^m (jf - \beta_i) \right) \left( \prod_{i=1}^m (jf - \beta_i) \right)^* \\ &= \prod_{i=1}^m (jf - \beta_i)(jf - \beta_i)^* \\ &= \prod_{i=1}^m (jf - \beta_i)j^*(f - (\beta_i/j)^*) \end{aligned}$$

and because  $j^*/j = -1$  we have

$$\begin{aligned} \left| \prod_{i=1}^m (jf - \beta_i) \right|^2 &= (-1)^m \prod_{i=1}^m (jf - \beta_i)(jf - j(\beta_i/j)^*) \\ &= (-1)^m \prod_{i=1}^m (jf - \beta_i)(jf - \beta_{m+i}) \\ &= (-1)^m \prod_{i=1}^{2m} (jf - \beta_i). \end{aligned}$$

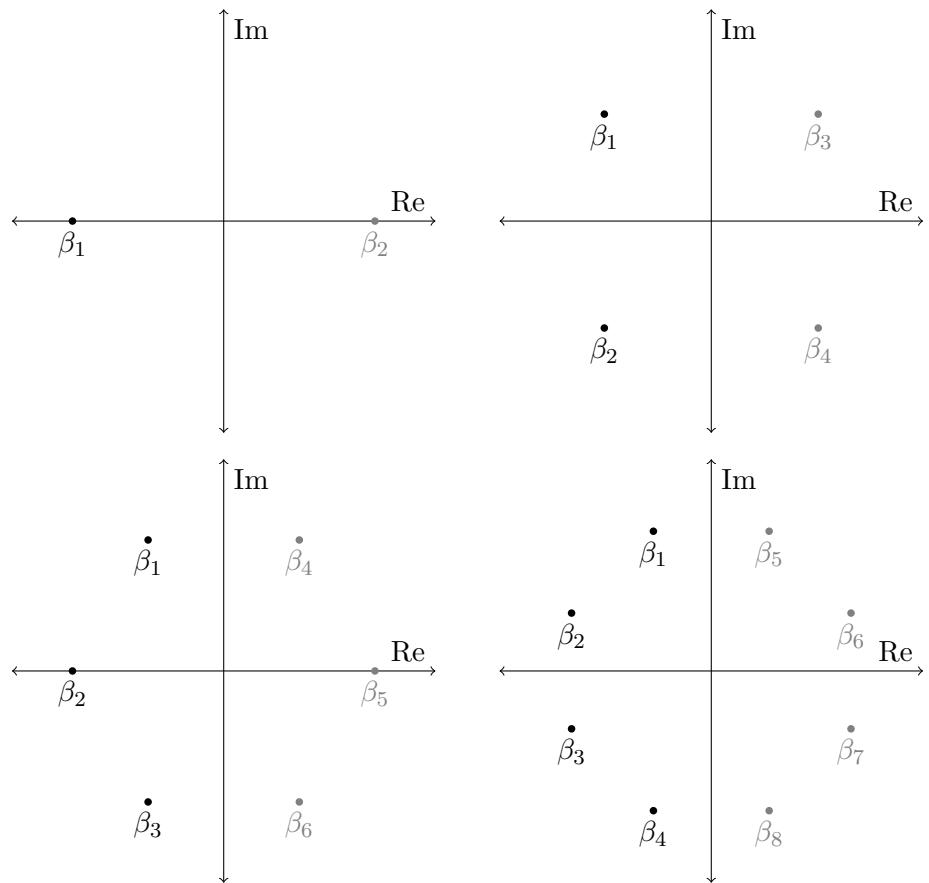


Figure 5.3: Roots of the polynomial  $s^{2m} + (-1)^m$  for  $m = 1$  (top left),  $m = 2$  (top right),  $m = 3$  (bottom left), and  $m = 4$  (bottom right). All the roots lie on the complex unit circle and have magnitude one. The poles of the normalised Butterworth filter  $B_m$  are those roots from the left half of the complex plane (unshaded).

Because  $\beta_1, \dots, \beta_{2m}$  are the roots of the polynomial  $s^{2m} + (-1)^m$  we have

$$\left| \prod_{i=1}^m (jf - \beta_i) \right|^2 = (-1)^m ((jf)^{2m} + (-1)^m) = f^{2m} + 1.$$

It follows that the magnitude spectrum of  $B_m$  is

$$|\Lambda(B_m)| = \sqrt{\frac{1}{f^{2m} + 1}}.$$

The magnitude and phase spectrum of the filters  $B_1$ ,  $B_2$ ,  $B_3$ , and  $B_4$  are plotted in Figure 5.4.

The **cutoff frequency** of the lowpass filter  $B_m$  is defined as the positive real number  $c$  such that  $|\Lambda(B_m, f)|^2 < \frac{1}{2}$  for all  $f > c$ . The normalised Butterworth filters have cutoff frequency  $c = 1\text{Hz}$ . A lowpass Butterworth filter of order  $m$  and cutoff frequency  $c$ , denoted  $B_m^c$ , has transfer function

$$\lambda(B_m^c, s) = \lambda(B_m, \frac{s}{c}) = \frac{1}{\prod_{i=1}^m (\frac{s}{2\pi c} - \beta_i)}.$$

The magnitude spectrum satisfies

$$|\Lambda(B_m^c, f)|^2 = |\Lambda(B_m, \frac{f}{c})|^2 = \frac{1}{(\frac{f}{c})^{2m} + 1} = \frac{c^{2m}}{f^{2m} + c^{2m}}. \quad (5.2.1)$$

A first order Butterworth filter  $B_1^c$  has spectrum

$$\Lambda(B_1^c) = \frac{1}{j\frac{f}{c} + 1} = \frac{c}{jf + c}.$$

Putting  $\frac{1}{c} = 2\pi RC$  we find that this is the same as the spectrum of the RC electrical circuit (Figure 2.1) or the active RC circuit after negation (3.5.2). Thus, the RC electrical circuit is a first order Butterworth filter with cutoff frequency  $c = \frac{1}{2\pi RC}$ . In Test 4 we constructed the active RC circuit with  $R \approx 27\text{k}\Omega$  and  $C \approx 10\text{nF}$  and measured its magnitude spectrum. The cutoff frequency was  $c = \frac{5 \times 10^4}{27\pi} \approx 589\text{Hz}$ .

A second order electrical Butterworth filter can be constructed using the Sallen-Key circuit described in Section 2.2 and Figure 2.10. The input voltage  $x$  and output voltage  $y$  of the Sallen-Key satisfy the differential equation (2.2.9)

$$x = y + C_2(R_1 + R_2)D(y) + R_1R_2C_1C_2D^2(y).$$

The transfer function is

$$\frac{\mathcal{L}(y)}{\mathcal{L}(x)} = \frac{1}{1 + C_2(R_1 + R_2)s + R_1R_2C_1C_2s^2}.$$

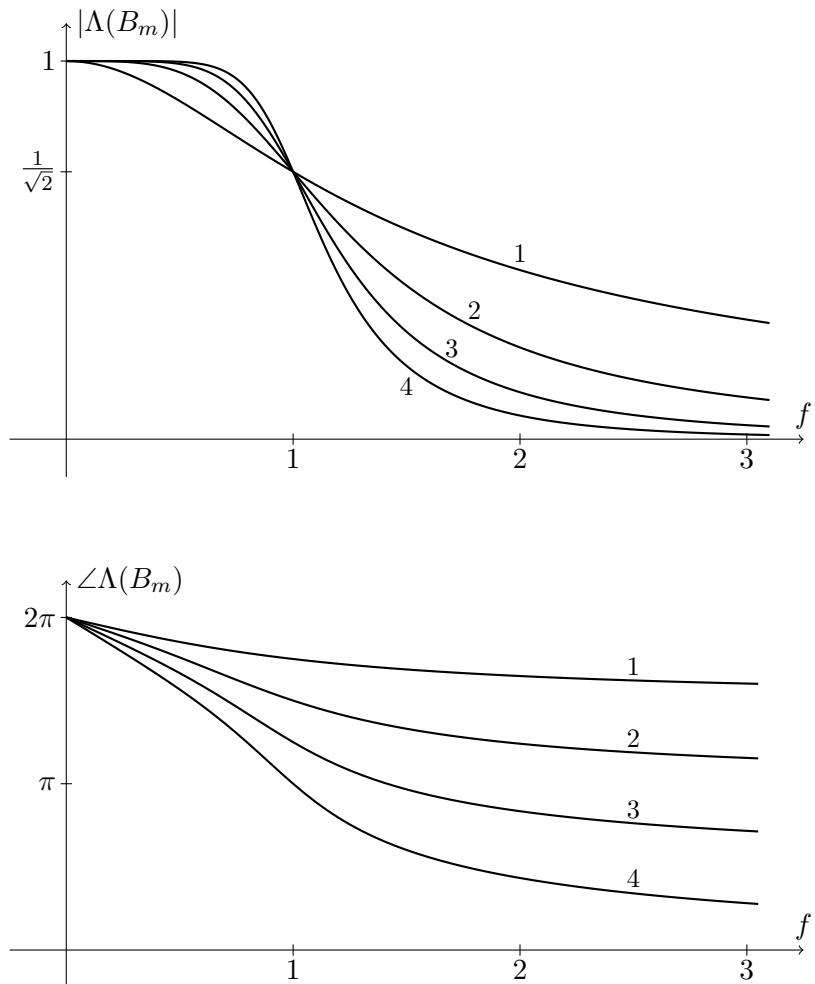


Figure 5.4: Magnitude spectrum (top) and phase spectrum (bottom) of normalised Butterworth filters  $B_1, B_2, B_3$  and  $B_4$ .

The second order Butterworth filter  $B_2^c$  has transfer function

$$\Lambda(B_m^c) = \frac{1}{(\frac{1}{2\pi c}s - \beta_1)(\frac{1}{2\pi c}s - \beta_2)},$$

where  $\beta_1 = \beta_2^* = e^{j3\pi/4}$ . Expanding the quadratic on the denominator gives

$$\Lambda(B_m^c) = \frac{1}{1 + \frac{1}{\sqrt{2}\pi c}s + \frac{1}{4\pi^2 c^2}s^2}.$$

Choosing the resistors and capacitors of the Sallen-Key to satisfy

$$C_2(R_1 + R_2) = \frac{1}{\sqrt{2}\pi c}, \quad R_1 R_2 C_1 C_2 = \frac{1}{4\pi^2 c^2}$$

leads to a second order Butterworth filter. A convenient solution is to put  $C_1 = 2C_2$  and  $R_1 = R_2$ . This gives a second order Butterworth filter with cutoff

$$c = \frac{1}{\sqrt{2}\pi C_2(R_1 + R_2)} = \frac{1}{\sqrt{2}\pi C_1 R_1}.$$

In Test 6 we construct a second order Butterworth filter using a Sallen-Key and measure its spectrum.

Butterworth filters of orders larger than  $m = 2$  can be constructed by concatenating Sallen-Key circuits and RC circuits. If  $m$  is even then  $m/2$  Sallen-Key circuits are required. Each Sallen-Key is used to construct a conjugate pair of poles, that is, the  $k$ th Sallen-Key would have poles  $2\pi c\beta_k$  and  $2\pi c\beta_k^* = 2\pi c\beta_{m-k+1}$ . If  $m$  is odd then  $(m-1)/2$  Sallen-Key circuits and a single RC circuit (or active RC circuit) can be used. The RC circuit is designed to have the real valued pole  $\beta_{(m+1)/2} = 2\pi c$ .

### Test 6 (Butterworth filter)

We construct a second order Butterworth filter using the Sallen-Key circuit from Figure 2.10 with capacitors  $C_2 \approx 100\text{nF}$ ,  $C_1 \approx 2C_2 \approx 200\text{nF}$  and resistors  $R_1 \approx R_2 \approx 1000\Omega$ . The cutoff frequency is

$$c = \frac{1}{\sqrt{2}\pi C_1 R_1} \approx 1125\text{Hz}.$$

Sinusoids of the form

$$\sin(2\pi f_k t), \quad f_k = \lceil 110 \times 2^{k/2} \rceil, \quad k = 0, 1, \dots, 12$$

are input to the filter using a computer soundcard and the magnitude and phase spectrum are measured using the procedure described in Test 4. Figure 5.5 shows the measurements (dots) plotted alongside the hypothesised magnitude spectrum

$$|\Lambda(B_2^c)| = \sqrt{\frac{1}{(f/1125)^4 + 1}}$$

and Figure 5.6 shows measurements (dots) plotted alongside the hypothesised phase spectrum  $\angle\Lambda(B_2^c)$ .

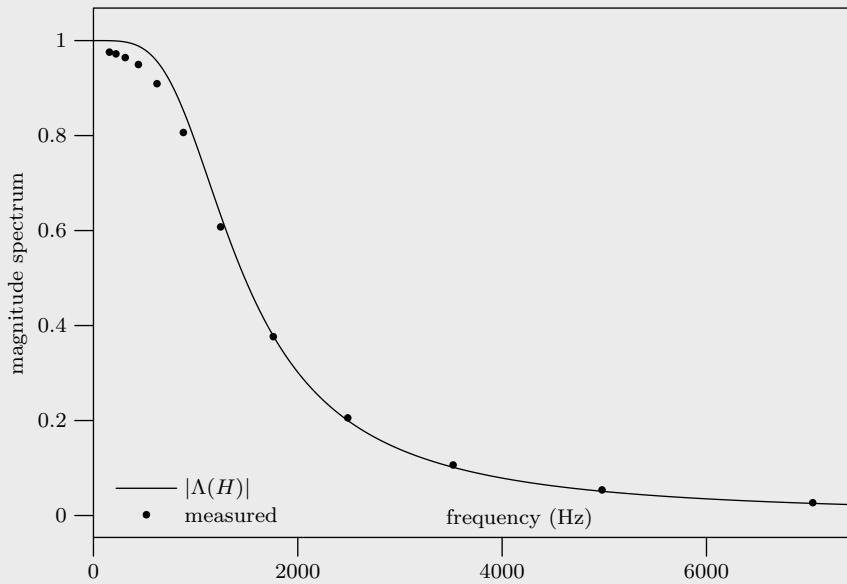


Figure 5.5: Plot of the hypothesised magnitude spectrum of the second order Butterworth filter  $|\Lambda(B_m^c)|$  (solid line) and of the measured magnitude spectrum of the filter implemented with a Sallen-Key active electrical circuit (dots).

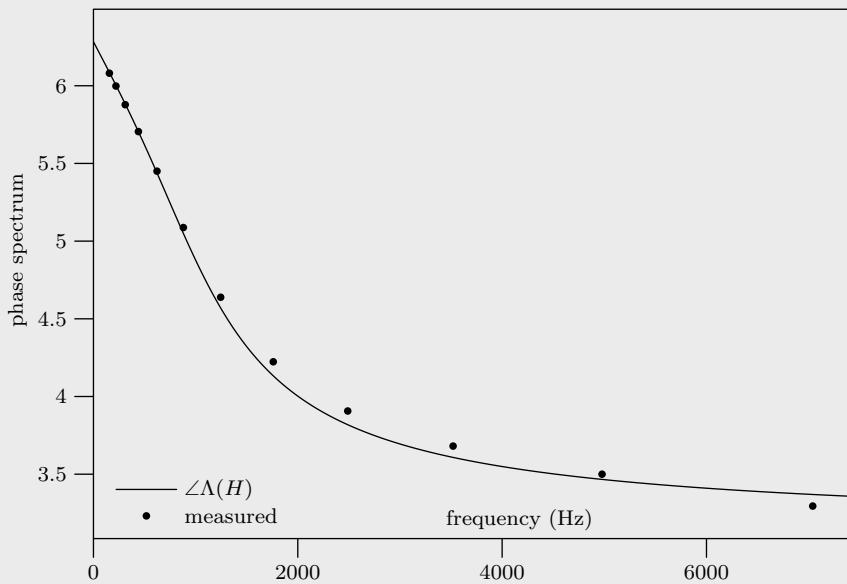


Figure 5.6: Plot of the hypothesised phase spectrum of the second order Butterworth filter  $\angle\Lambda(B_m^c)$  (solid line) and of the measured phase spectrum of the filter implemented with a Sallen-Key active electrical circuit (dots).

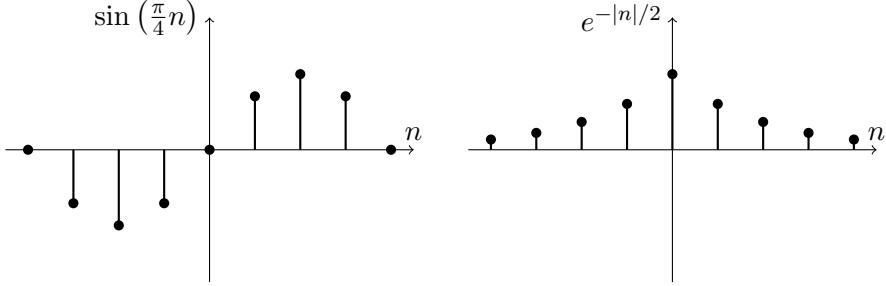


Figure 5.7: 1-dimensional real valued sequences (discrete time signals).

### 5.3 Real and complex valued sequences

Let  $x$  be a signal with Fourier transform  $\hat{x} = \mathcal{F}(x)$ . The signal  $x$  is said to be **bandlimited** if there exists a positive real number  $b$  such that

$$\hat{x}(f) = \mathcal{F}(x, f) = 0 \quad \text{for all } |f| > b.$$

The value  $b$  is called the **bandwidth** of the signal  $x$ . For example, the sinc function is bandlimited with bandwidth  $\frac{1}{2}$  because its Fourier transform  $\mathcal{F}(\text{sinc}, f) = \Pi(f) = 0$  for all  $|f| > \frac{1}{2}$ . Bandlimited signals have a number of properties that make them suitable for representation and manipulation by a computer. They are of particular importance for this reason. Before we can study bandlimited signals we first require some properties of real and complex valued **sequences**.

A **sequence** is a function with domain given by the integers  $\mathbb{Z}$ . The value of the sequence corresponding with the integer  $n$  could be denoted by  $x(n)$  but it is conventional to instead write  $x_n$ . We are primarily interested in sequences that take real or complex values, that is,  $x_n \in \mathbb{R}$  or  $x_n \in \mathbb{C}$ . For example, the expressions

$$\sin(\frac{\pi}{4}n), \quad n^3, \quad e^{-|n|/2}$$

each denote a real valued sequence. In what follows the term **sequence** will always mean a real or complex valued sequence unless otherwise stated. Real and complex valued sequences are commonly called **discrete time signals** and the  $n$ th element in the sequence is denoted by using squared brackets  $x[n]$  [Oppenheim et al., 1996]. Here, we use the subscript notation  $x_n$ . This notation is also common [Vetterli et al., 2014; Rudin, 1986]. Sequences are plotted using vertical lines with dotted ends as in Figure 6.1. Sequences have a number of properties analogous to the properties of signals (Section 1.1).

A sequence  $x$  is bounded if there exists a real number  $M$  such that

$$|x_n| \leq M \quad \text{for all } n \in \mathbb{Z}.$$

Both  $\sin(\frac{\pi}{4}n)$  and  $e^{-|n|/2}$  are examples of bounded sequences, but  $n^3$  is not bounded because its magnitude grows indefinitely as  $n$  moves away from the origin. A sequence  $x$  is periodic if there exists a positive integer  $T$  such that

$$x_n = x_{n+kT} \quad \text{for all integers } k \text{ and } n.$$

The smallest such positive  $T$  is called the period. The sequence  $\sin(\frac{\pi}{4}n)$  is periodic with period  $T = 8$ . Neither  $n^3$  or  $e^{-n^2/4}$  are periodic. A sequence  $x$  is even (or symmetric) if  $x_n = x_{-n}$  for all  $n \in \mathbb{Z}$  and odd (or antisymmetric) if  $x_n = -x_{-n}$  for all  $n \in \mathbb{Z}$ . Both  $\sin(\frac{\pi}{4}n)$  and  $n^3$  are odd and  $e^{-|n|/2}$  is even.

A sequence  $x$  is **absolutely summable** if

$$\|x\|_1 = \sum_{n \in \mathbb{Z}} |x_n| < \infty,$$

that is, if the sum of absolute values of the elements in the sequence converges to a finite number. The real number  $\|x\|_1$  is commonly called the  $\ell^1$ -norm of  $x$ . The sequences  $\sin(\frac{\pi}{4}n)$  and  $n^3$  are not absolutely summable, but  $e^{-|n|/2}$  is because (Exercise 5.8)

$$\sum_{n \in \mathbb{Z}} |e^{-|n|/2}| = \sum_{n \in \mathbb{Z}} e^{-|n|/2} = 1 + \frac{2}{\sqrt{e} - 1}.$$

It is common to denote the set of absolutely summable sequences by  $\ell^1$ . So,  $e^{-|n|/2} \in \ell^1$  and  $\sin(\frac{\pi}{4}n) \notin \ell^1$ .

A sequence  $x$  is **square summable** if

$$\|x\|_2^2 = \sum_{n \in \mathbb{Z}} |x_n|^2 < \infty,$$

that is, if the sum of squared magnitudes of elements converge to a finite number. The real number  $\|x\|_2$  is commonly called the  $\ell^2$ -norm or **energy** of  $x$ . The sequences  $\sin(\frac{\pi}{4}n)$  and  $n^3$  are not square summable, but  $e^{-|n|/2}$  is because (Exercise 5.8)

$$\sum_{n \in \mathbb{Z}} |e^{-|n|/2}|^2 = \sum_{n \in \mathbb{Z}} e^{-|n|} = 1 + \frac{2}{e - 1}.$$

It is common to denote the set of square summable sequences by  $\ell^2$ . So,  $e^{-|n|/2} \in \ell^2$  and  $\sin(\frac{\pi}{4}n) \notin \ell^2$ . If a sequence is absolutely summable then it is also square summable (Exercise 5.9). The corresponding property is not true of signals, that is, absolutely integrable signals are not necessarily square integrable (Exercise 1.4).

## 5.4 Bandlimited signals

Let  $x$  be a signal with Fourier transform  $\hat{x} = \mathcal{F}(x)$ . The signal  $x$  is said to be **bandlimited** if there exists a positive real number  $b$  such that

$$\hat{x}(f) = \mathcal{F}(x, f) = 0 \quad \text{for all } |f| > b.$$

The value  $b$  is called the **bandwidth** of the signal  $x$ . The quintessential bandlimited signal is the sinc function  $\text{sinc}(t)$  that has Fourier transform  $\Pi(f)$  and bandwidth  $\frac{1}{2}$ . It is often simpler to discover examples of bandlimited signals by first specifying their Fourier transform  $\hat{x}$  and applying the inverse Fourier transform to obtain the time domain signal  $x$ . For example, the signal with Fourier transform given by a **raised cosine**

$$\hat{x}(f) = \Pi(f)(1 + \cos(2\pi f)) = \begin{cases} 1 + \cos(2\pi f) & |f| < \frac{1}{2} \\ 0 & \text{otherwise} \end{cases}$$

is bandlimited with bandwidth  $\frac{1}{2}$ . The time domain signal is found by applying the inverse Fourier transform (Exercise 5.6)

$$x(t) = \text{sinc}(t) + \frac{1}{2} \text{sinc}(t+1) + \frac{1}{2} \text{sinc}(t-1).$$

Another example is the signal with Fourier transform given by the **triangle pulse**

$$\hat{x}(f) = \Delta(f) = \begin{cases} f+1 & -1 \leq f < 0 \\ 1-f & 0 \leq f < 1 \\ 0 & \text{otherwise.} \end{cases}$$

The corresponding time domain signal is given by the square of the sinc function  $\text{sinc}^2(t)$  (Exercises 5.2). These bandlimited signals and their Fourier transforms are plotted in Figure 5.8.

It happens that bandlimited signals are always unbounded in time (Section 1.1) and that signals that are finite in time are never bandlimited. We can reasonably suppose that all signals ever encountered in practice are finite in time and so no signals encountered in practice are truly bandlimited. However, many practically occurring signals are approximately bandlimited, that is, their Fourier transform is small for frequencies larger than some positive number  $b$ . For example, in Test 7 the Fourier transform of an audio signal taken from a lecture recording is plotted Figure ???. This signal appears bandlimited below approximately 9 kHz.

A surprising result is that every square integrable bandlimited signal  $x$  with bandwidth  $b$  can be written as a sum of time scaled and time shifted sinc functions, that is, in the form

$$x(t) = \sum_{n \in \mathbb{Z}} c_n \text{sinc}(Rt - n)$$

where  $c$  is a square integrable complex valued sequence and  $R = 2b$ . This is a consequence of the **Riez-Fisher theorem** [Rudin, 1986, page ]. Evaluating the signal  $x$  at integer multiples of  $P = \frac{1}{R}$  we find that

$$x(\ell P) = \sum_{n \in \mathbb{Z}} c_n \operatorname{sinc}(\ell - n) = c_\ell$$

because  $\operatorname{sinc}(\ell - n)$  is equal to one when  $\ell = n$  and zero otherwise. Thus, the elements of the sequence  $c_\ell$  correspond with samples of the signal  $x$  taken at integer multiples of  $P = \frac{1}{R} = \frac{1}{2b}$ . The positive real number  $P$  is often called the **sampling period** and its reciprocal  $R$  the **sampling rate**. It follows that every square integrable bandlimited signal  $x$  with bandwidth  $b$  can be recovered from it samples taken at rate  $R = 2b$ , that is,

$$x(t) = \sum_{n \in \mathbb{Z}} x(nP) \operatorname{sinc}(Rt - n).$$

This is called the **Nyquist sampling theorem**.

## 5.5 The discrete Fourier transform

Let  $c$  be a square integrable sequence

$$x(t) = \sum_{n \in \mathbb{Z}} c_n \operatorname{sinc}(Rt - n)$$

be the corresponding bandlimited signal with sampling rate  $R$  and sampling period  $P = \frac{1}{R}$ . The elements of the sequence  $c$  are samples of  $x$ , that is,  $c_n = x(nP)$ . The Fourier transform of  $x$  is

$$\begin{aligned} \hat{x}(f) &= \sum_{n \in \mathbb{Z}} c_n \mathcal{F}(\operatorname{sinc}(Rt - n)) \\ &= P\Pi(fP) \sum_{n \in \mathbb{Z}} c_n e^{-j2\pi Pnf} \\ &= P\Pi(fP) \mathcal{F}_d(c, Pf) \end{aligned}$$

where

$$\mathcal{F}_d(c, f) = \sum_{n \in \mathbb{Z}} c_n e^{-j2\pi nf}$$

is called the **discrete Fourier transform** of the sequence  $c$ . We will write  $\hat{c}(f) = \mathcal{F}(c, f)$  for the discrete Fourier transform of the sequence  $c$ . The discrete Fourier transform  $\hat{c}$  is a periodic signal with period 1. The Fourier transform of the bandlimited signal  $x$  with sampling period  $P$  is related to the discrete Fourier transform of its sequence of samples  $c_n = x(nP)$  by

$$\hat{x}(f) = \mathcal{F}(x, f) = P\Pi(fP) \mathcal{F}_d(c, Pf) = P\Pi(fP) \hat{c}(Pf).$$

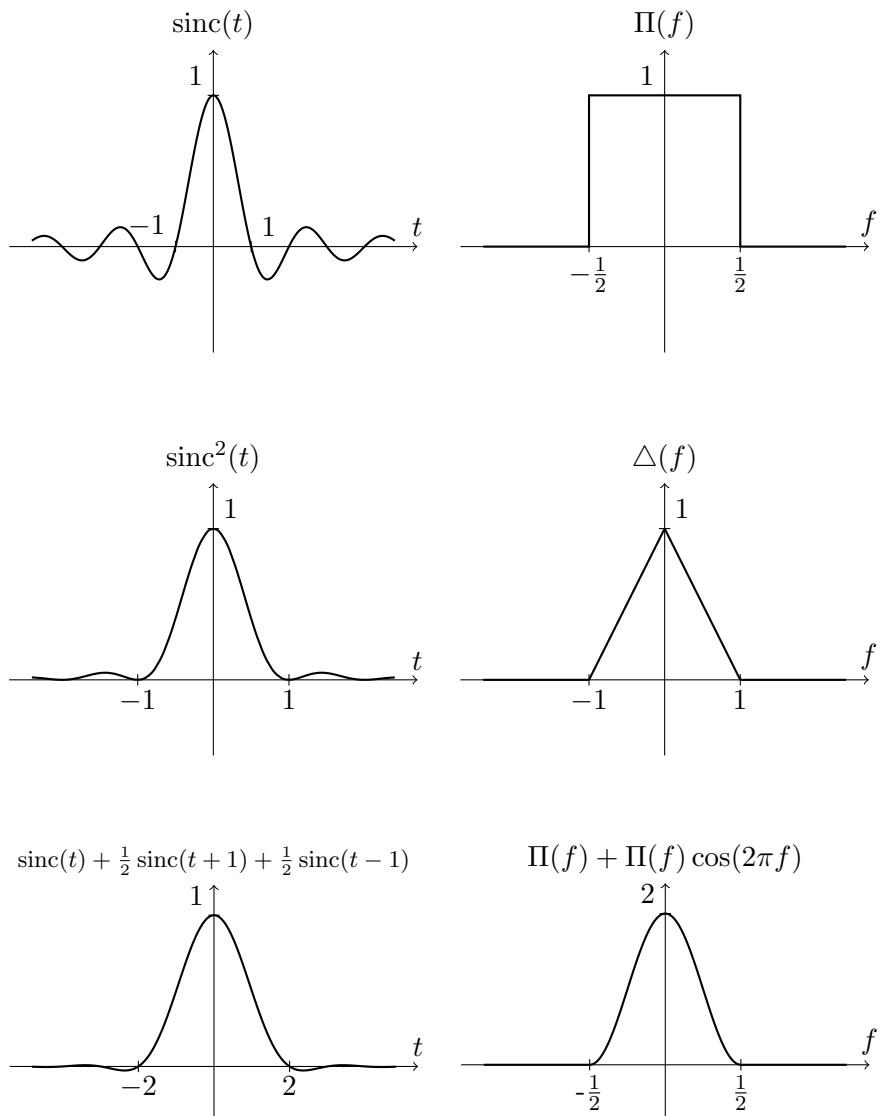


Figure 5.8: Bandlimited signals  $\text{sinc}(t)$ ,  $\text{sinc}^2(t)$ , and  $\text{sinc}(t) + \frac{1}{2} \text{sinc}(t+1) + \frac{1}{2} \text{sinc}(t-1)$  with bandwidth  $\frac{1}{2}$ , 1, and 1 respectively.

In Test 7 we compute the Fourier transform of a 20 s segment of audio from a lecture recording.

### Test 7 (The Fourier transform of a lecture recording)

In this test we consider a 20 s segment of audio taken from the lecture video ch1sec3.mp4. This 34.8 MB file contains both compressed video (H.264 codec) and audio (mp3 codec) of duration 23 min and 36 s. The audio is mono and sampled at rate  $R = 22050$  Hz. The `avconv` program from the libav library is used to extract a 20 s segment of audio starting at time 85 s and ending at time 105 s. The segment is decompress to wav format. The command used is:

```
avconv -i ch1sec3.mp4 -ss 85 -t 20 audio.wav
```

The resulting file `audio.wav` is 882 kB in size and contains  $N = 441216$  samples that we denote by  $c_0, c_1, \dots, c_{N-1}$ . Each sample takes a value in the interval  $[-1, 1]$ . We put  $c_n = 0$  when  $n < 0$  or  $n \geq N$ . The reconstructed audio signal is given by

$$x(t) = \sum_{n \in \mathbb{Z}} c_n \operatorname{sinc}(Rt - n) = \sum_{n=0}^{N-1} c_n \operatorname{sinc}(Rt - n).$$

The Fourier transform of this signal is  $\mathcal{F}(x) = P\Pi(Pf)\mathcal{F}_d(c, Pf)$  where

$$\mathcal{F}_d(c, f) = \sum_{n \in \mathbb{Z}} c_n e^{-j2\pi n f} = \sum_{n=0}^{N-1} c_n e^{-j2\pi n f}$$

is the discrete Fourier transform of the samples. Figure 5.9 shows a plot of the magnitude of the Fourier transform for frequencies in the interval  $-12$  kHz to  $12$  kHz. The plot is constructed by evaluating  $|\mathcal{F}(x, f)|$  at all  $K = 12000$  frequencies from the grid

$$f_k = -12000 + 20k \quad k = 0, \dots, K-1$$

that is, in steps of 20 Hz. It takes approximately 137 s to compute the Fourier transform at all of these frequencies. The computer used is an Intel Core 2 running at 2.4 GHz.

The audio recording contains human voice that primarily resides at the lower frequencies below 5000 Hz. Audible in the recording is a faint high pitched hum. The cause of this could be electronic interference in the (probably low quality) webcam microphone used to record the audio. This hum is represented in Figure 5.9 by the spikes occurring at approximately  $\pm 8$  kHz.

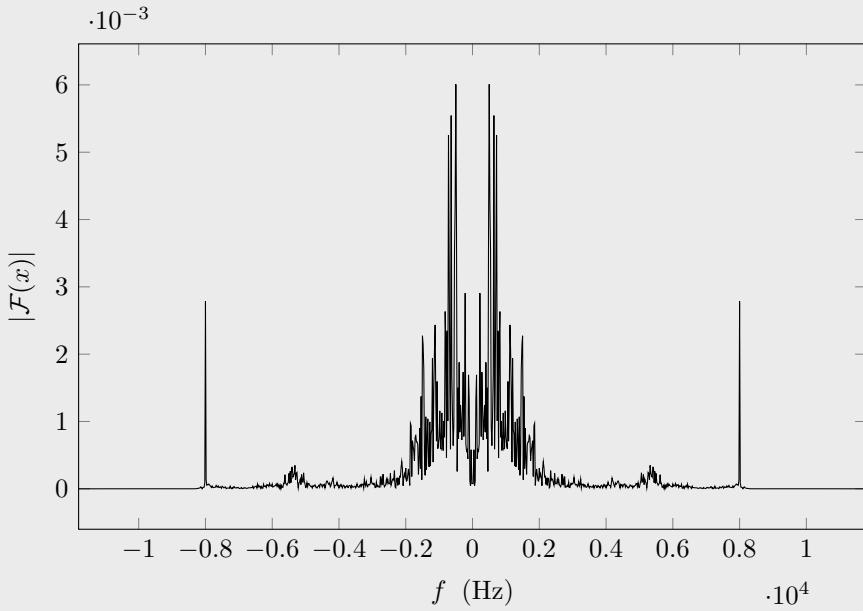


Figure 5.9: Magnitude of the Fourier transform of 20s of audio from a lecture recording. The human voice signals is primarily contained in the low frequency region below 5000 Hz. The spikes occurring at approximately  $\pm 8\text{ kHz}$  might be caused by some type of interference and are audible in the recording as a high pitched hum.

#### Test 8 (The fast Fourier transform of a lecture recording)

## 5.6 Exercises

- 5.1. Plot the signal  $e^{-\alpha|t|}$  where  $\alpha > 0$  and find its Fourier transform.

**Solution:**

$$\begin{aligned}\mathcal{F}(e^{-\alpha|t|}) &= \int_{-\infty}^{\infty} e^{-\alpha|t|} e^{-j2\pi ft} dt \\ &= \int_0^{\infty} e^{-\alpha t} e^{-j2\pi ft} dt + \int_{-\infty}^0 e^{\alpha t} e^{-j2\pi ft} dt \\ &= \int_0^{\infty} e^{-(j2\pi f+\alpha)t} dt + \int_{-\infty}^0 e^{-(j2\pi f-\alpha)t} dt \\ &= \left[ \frac{e^{-(j2\pi f+\alpha)t}}{-(j2\pi f+\alpha)} \right]_0^{\infty} + \left[ \frac{e^{-(j2\pi f-\alpha)t}}{-(j2\pi f-\alpha)} \right]_{-\infty}^0.\end{aligned}$$

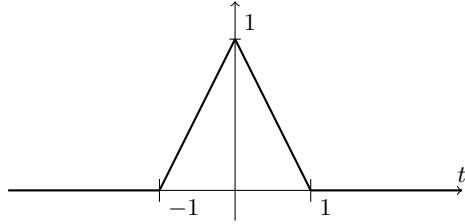
Because  $\alpha > 0$ , the limits as  $t \rightarrow \infty$  and  $t \rightarrow -\infty$  go to zero leaving

$$\frac{1}{j2\pi f + \alpha} - \frac{1}{j2\pi f - \alpha} = \frac{j2\pi f + \alpha - j2\pi f + \alpha}{(j2\pi f + \alpha)(j2\pi f - \alpha)} = \frac{2\alpha}{4\pi^2 f^2 + \alpha^2}.$$

5.2. Plot the signal

$$\Delta(t) = \begin{cases} t+1 & -1 \leq t < 0 \\ 1-t & 0 \leq t < 1 \\ 0 & \text{otherwise} \end{cases}$$

and find its Fourier transform. **Solution:** This signal is often called the **triangle function** or **triangle pulse**.



You can do this directly using the formula for the Fourier transform and integrating by parts. However, it is easier to first realise that the triangle pulse is the convolution of the rectangular function with itself. That is  $\Pi * \Pi = \Delta$ . To see this write

$$(\Pi * \Pi)(t) = \int_{-\infty}^{\infty} \Pi(\tau) \Pi(t - \tau) d\tau = \int_{-1/2}^{1/2} \Pi(t - \tau) d\tau$$

Now  $\Pi(t - \tau) = 1$  for  $\tau$  in the interval  $(-\frac{1}{2} + t, \frac{1}{2} + t)$  and zero otherwise. Thus, the integral evaluates to zero if  $t \geq 1$  or  $t \leq -1$ . When  $t \in [-1, 0)$

$$(\Pi * \Pi)(t) = \int_{-1/2}^{1/2+t} d\tau = t + 1$$

and when  $t \in [0, 1)$

$$(\Pi * \Pi)(t) = \int_{-1/2+t}^{1/2} d\tau = 1 - t$$

as required. Now, by the convolution theorem (5.0.3)

$$\mathcal{F}(\Pi * \Pi) = \mathcal{F}(\Delta) = \mathcal{F}(\Pi)\mathcal{F}(\Pi) = \text{sinc}^2(t).$$

5.3. Show that the sinc function is square integrable, but not absolutely integrable. **Solution:** Our proof is by contradiction. Assume that sinc is absolutely integrable. Then

$$\begin{aligned} \|\text{sinc}\|_1 &= \int_{-\infty}^{\infty} |\text{sinc}(t)| dt \\ &> \int_0^{\infty} |\text{sinc}(t)| dt \\ &= \sum_{n=1}^{\infty} \int_{n-1}^n \left| \frac{\sin(\pi t)}{\pi t} \right| dt \\ &= \sum_{n=1}^{\infty} a_n \end{aligned}$$

where we put

$$a_n = \int_{n-1}^n \left| \frac{\sin(\pi t)}{\pi t} \right| dt.$$

Under our assumption that sinc is absolutely integrable we must have that the infinite sum  $a_1 + a_2 + \dots$  converge to a finite number. Now

$$a_n \geq \int_{n-1}^n \left| \frac{\sin(\pi t)}{\pi n} \right| dt = \frac{1}{\pi n} \int_{n-1}^n |\sin(\pi t)| dt = \frac{2}{\pi^2 n}.$$

However, the sum

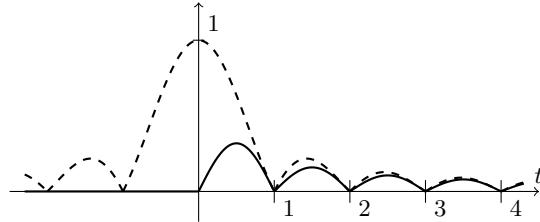
$$\sum_{n=1}^{\infty} a_n = \frac{2}{\pi^2} \sum_{n=1}^{\infty} \frac{1}{n}$$

involves the harmonic series (a  $p$ -series with  $p = 1$ ) and so diverges (to show this use either an integral test or the condensation test). Thus, our initial hypothesis that sinc is absolutely integrable is false.

Graphically, the argument we have used bounds  $|\text{sinc}|$  above the function

$$b(t) = \begin{cases} 0 & t \leq 0 \\ \left| \frac{\sin(\pi t)}{\pi n} \right| & t \in (n-1, n] \end{cases}$$

and then shows that  $b$  is not absolutely integrable. The function  $|\text{sinc}|$  (dashed) and  $b$  (solid) are plotted in the figure below.



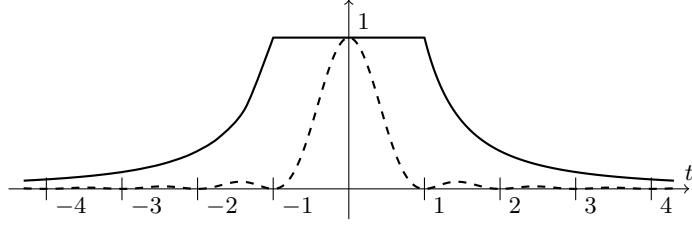
To show that sinc is square integrable observe that  $\text{sinc}^2(t)$  is bounded below the function

$$g(t) = \begin{cases} 1 & |t| \leq 1 \\ \frac{1}{t^2} & \text{otherwise,} \end{cases}$$

that is  $\text{sinc}^2(t) \leq g(t)$  for all  $t \in \mathbb{R}$ . Thus

$$\begin{aligned} \|\text{sinc}\|_2 &= \int_{-\infty}^{\infty} |\text{sinc}(t)|^2 dt \\ &\leq \int_{-\infty}^{\infty} g(t) dt \\ &= \int_{-1}^1 dt + 2 \int_1^{\infty} \frac{1}{t^2} dt \\ &= 2 - \left[ \frac{1}{t} \right]_1^{\infty} = 2 + 1 = 3. \end{aligned}$$

The figure below plots  $\text{sinc}^2$  (dashed) and the bounding function  $g$ .

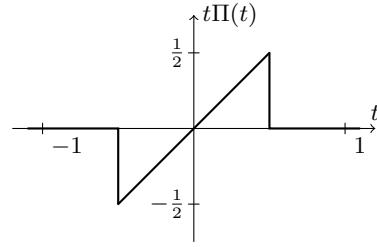


- 5.4. Find and plot the impulse response of the normalised lowpass Butterworth filters  $B_1, B_2$  and  $B_3$ .

- 5.5. Plot the signal

$$t\Pi(t) = \begin{cases} t & -\frac{1}{2} < t \leq \frac{1}{2} \\ 0 & \text{otherwise} \end{cases}$$

and find its Fourier transform. **Solution:**



A direct approach is

$$\mathcal{L}(t\Pi(t)) = \int_{-\infty}^{\infty} t\Pi(t)e^{-st} dt = \int_{-1/2}^{1/2} te^{-st} dt$$

and integrating by parts gives

$$\begin{aligned} \mathcal{L}(t\Pi(t)) &= \left[ t \frac{e^{-st}}{-s} \right]_{-1/2}^{1/2} - \int_{-1/2}^{1/2} \frac{e^{-st}}{-s} dt \\ &= \frac{e^{-s/2} + e^{s/2}}{-2s} - \left[ \frac{e^{-st}}{s^2} \right]_{-1/2}^{1/2} \\ &= \frac{e^{-s/2} + e^{s/2}}{-2s} - \frac{e^{-s/2} - e^{s/2}}{s^2} \\ &= \frac{1}{s} \left( -\frac{e^{s/2} + e^{-s/2}}{2} + \frac{e^{s/2} - e^{-s/2}}{s} \right). \end{aligned}$$

Because  $t\Pi(t)$  is absolutely integrable its region of convergence includes the imaginary axis and we can obtain the Fourier transform by evaluating the Laplace transform at  $s = j2\pi f$ ,

$$\begin{aligned} \mathcal{F}(t\Pi(t), f) &= \mathcal{L}(x, j2\pi f) \\ &= \frac{1}{2\pi j f} \left( -\frac{e^{j\pi f} + e^{-j\pi f}}{2} + \frac{e^{j\pi f} - e^{-j\pi f}}{2j\pi f} \right) \\ &= \frac{1}{2\pi j f} (\operatorname{sinc}(f) - \cos(\pi f)). \end{aligned}$$

An alternative approach is to observe that

$$\mathcal{F}(D(\text{sinc})) = \Lambda(D)\mathcal{F}(\text{sinc}) = j2\pi f\Pi(f),$$

and so, by duality,

$$\mathcal{F}(j2\pi t\Pi(t), f) = \mathcal{F}(\mathcal{F}(D(\text{sinc})), f) = D(\text{sinc}, -f)$$

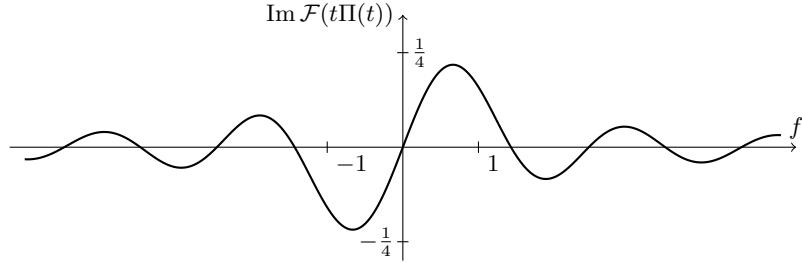
The derivative of the sinc function is given in (2.2.5)

$$D(\text{sinc}, -f) = \frac{1}{\pi f^2} (\sin(\pi f) - \pi f \cos(\pi f)) = \frac{1}{f} (\text{sinc}(f) - \cos(\pi f)).$$

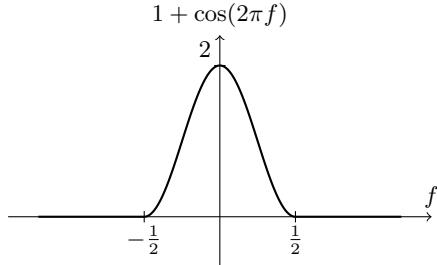
Dividing by  $j2\pi$  we obtain

$$\mathcal{F}(t\Pi(t), f) = \frac{1}{2j\pi^2 f^2} (\pi f \cos(\pi f) - \sin(\pi f)) = \frac{1}{2j\pi f} (\text{sinc}(f) - \cos(\pi f))$$

again. A plot of the Fourier transform is below. Observe that the Fourier transform is purely imaginary so we plot the imaginary part.



- 5.6. Let  $x$  be the signal with Fourier transform  $\hat{x}(f) = (1 + \Pi(f)) \cos(2\pi f)$ . Plot the Fourier transform  $\hat{x}$  and find and plot  $x$ . **Solution:**



We will solve the problem in two ways. Firstly, by direct application of the inverse Fourier transform and secondly, by applications of the duality and modulation properties. Application of the inverse Fourier transform gives

$$x = \mathcal{F}^{-1}(\Pi(f) + \Pi(f) \cos(2\pi f)) = \mathcal{F}^{-1}(\Pi(f)) + \mathcal{F}^{-1}(\Pi(f) \cos(2\pi f)).$$

Now

$$\begin{aligned} \mathcal{F}^{-1}(\Pi(f)) &= \int_{-\infty}^{\infty} \Pi(f) e^{2\pi jft} df \\ &= \int_{-1/2}^{1/2} e^{2\pi jft} df \\ &= \frac{e^{\pi jt} - e^{-\pi jt}}{2\pi jt} = \frac{2j \sin(\pi t)}{2\pi jt} = \frac{\sin(\pi t)}{\pi t} = \text{sinc}(t) \end{aligned}$$

and

$$\begin{aligned}
 \mathcal{F}^{-1}(\Pi(f) \cos(2\pi f)) &= \int_{-\infty}^{\infty} \Pi(f) \cos(2\pi f) e^{2\pi jft} df \\
 &= \int_{-1/2}^{1/2} \cos(2\pi f) e^{2\pi jft} df \\
 &= \int_{-1/2}^{1/2} \frac{1}{2}(e^{2\pi jf} + e^{-2\pi jf}) e^{2\pi jft} df \\
 &= \int_{-1/2}^{1/2} e^{2\pi jf(t+1)} df + \frac{1}{2} \int_{-1/2}^{1/2} e^{2\pi jf(t-1)} df \\
 &= \frac{1}{2} \text{sinc}(t+1) + \frac{1}{2} \text{sinc}(t-1)
 \end{aligned}$$

by working similarly to the previous equation. Putting these together we obtain the time domain signal

$$x(t) = \text{sinc}(t) + \frac{1}{2} \text{sinc}(t+1) + \frac{1}{2} \text{sinc}(t-1).$$

We now derive the same result using duality (5.1.4) and the modulation (5.0.6) properties. By duality

$$x(-f) = \mathcal{F}(\Pi(f)) + \mathcal{F}(\Pi(f) \cos(2\pi f))$$

Because  $\mathcal{F}(\Pi) = \text{sinc}$  the time shift properties yields

$$\mathcal{F}(\Pi(t)) = \text{sinc}(t).$$

Now, by the modulation property

$$\begin{aligned}
 \mathcal{F}(\Pi(t) \cos(2\pi f), f) &= \frac{1}{2} \mathcal{F}(\Pi(t), f-1) + \frac{1}{2} \mathcal{F}(\Pi(t), f+1) \\
 &= \frac{1}{2} \text{sinc}(f-1) + \frac{1}{2} \text{sinc}(f+1).
 \end{aligned}$$

Now

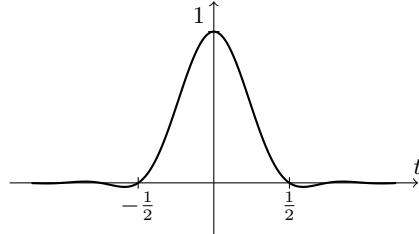
$$x(-f) = \text{sinc}(f) + \frac{1}{2} \text{sinc}(f-1) + \frac{1}{2} \text{sinc}(f+1)$$

Putting  $t = -f$

$$x(t) = \text{sinc}(t) + \frac{1}{2} \text{sinc}(t+1) + \frac{1}{2} \text{sinc}(t-1)$$

because sinc is even. A plot of the Fourier transform is below. The shape of the Fourier transform is somewhat sinc-like, but the oscillations decay faster as  $|t| \rightarrow \infty$ .

$$\text{sinc}(t) + \frac{1}{2} \text{sinc}(t+1) + \frac{1}{2} \text{sinc}(t-1)$$



- 5.7. State whether the following signals are bandlimited and, if so, find the bandwidth.

- (a)  $\text{sinc}(4t)$ ,

- (b)  $\Pi(t/4)$ ,
- (c)  $\cos(2\pi t) \operatorname{sinc}(t)$ ,
- (d)  $e^{-|t|}$ .

**Solution:** Let  $S_\alpha(x, t) = x(\alpha t)$  be the time scalar system. We have

$$\begin{aligned}\mathcal{F}(S_\alpha(x), f) &= \int_{-\infty}^{\infty} x(\alpha t) e^{-2\pi j t} dt \\ &= \frac{1}{\alpha} \int_{-\infty}^{\infty} x(\gamma) e^{-2\pi j \gamma/\alpha} d\gamma \quad (\text{ch. var. } \gamma = \alpha t) \\ &= \frac{1}{\alpha} \mathcal{F}(x, f/\alpha) \\ &= \frac{1}{\alpha} S_{1/\alpha}(\mathcal{F}(x), f).\end{aligned}$$

The Fourier transform of  $S_\alpha(\operatorname{sinc})(t) = \operatorname{sinc}(4t)$  is

$$\mathcal{F}(\operatorname{sinc}(4t)) = \frac{1}{4} \Pi(f/4),$$

and the signal is bandlimited with bandwidth 2 because  $\Pi(f/4) = 0$  whenever  $|f| > 2$ . By duality

$$\frac{1}{4} \mathcal{F}(\Pi(f/4)) = \operatorname{sinc}(4t)$$

and so  $\mathcal{F}(\Pi(f/4)) = 4 \operatorname{sinc}(4t)$ . This signal is not bandlimited because the sinc function is unbounded in time. By the modulation property of Fourier transform (5.0.6),

$$\mathcal{F}(\cos(2\pi t) \operatorname{sinc}(t), f) = \mathcal{F}(\operatorname{sinc}, f - 1) + \mathcal{F}(\operatorname{sinc}, f + 1) = \Pi(f - 1) + \Pi(f + 1).$$

This is bandlimited with bandwidth  $\frac{3}{2}$ . In Exercise 5.1 we showed that

$$\mathcal{F}(e^{-|t|}) = \frac{2}{4\pi^2 f^2 + 1}.$$

This signal is not bandlimited.

- 5.8. Show that  $\sum_{n \in \mathbb{Z}} e^{\alpha|n|} = 1 + 2/(e^{-\alpha} - 1)$  if  $\alpha < 0$  (Hint: solve Exercise 3.5 first). **Solution:** Put  $r = e^\alpha$ . Because  $\alpha < 0$  the series  $r + r^2 + r^3 + \dots$  converges absolutely and so

$$\sum_{n \in \mathbb{Z}} e^{\alpha|n|} = \sum_{n \in \mathbb{Z}} r^{|n|} = 1 + 2 \sum_{n=1}^{\infty} r^n$$

The sum in the equation on the right is a geometric series evaluating to

$$\frac{r}{1-r} = \frac{e^\alpha}{1-e^\alpha} = \frac{1}{e^{-\alpha}-1}.$$

- 5.9. Show that if a sequence absolutely summable then it is also square summable. **Solution:** Suppose that the discrete time signal  $a$  is absolutely summable so that  $\|a\|_1 = \sum_{n \in \mathbb{Z}} |a_n| < \infty$ . Let  $A$  be the subset of  $\mathbb{Z}$  such that  $|a_n| \geq 1$  whenever  $n \in A$ . That is

$$A = \{n ; |a_n| \geq 1\}.$$

Let  $|A|$  denote the number of elements in the set  $A$ . We have

$$\infty > \|a\|_1 = \sum_{n \in \mathbb{Z}} |a_n| \geq \sum_{n \in A} |a_n| \geq |A|$$

and so,  $A$  contains a finite number of elements. Thus

$$B = \sum_{n \in A} |a_n|^2 < \infty$$

Now,  $|a_n|^2 < |a_n| < 1$  for all  $n \notin A$  and so

$$C = \sum_{n \notin A} |a_n|^2 \leq \sum_{n \notin A} |a_n| \leq \|a\|_1 < \infty.$$

Thus

$$\|a\|_2^2 = \sum_{n \in \mathbb{Z}} |a_n|^2 = \sum_{n \in A} |a_n|^2 + \sum_{n \notin A} |a_n|^2 = B + C < \infty.$$

- 5.10. Let  $\{x_n\}$  be a square summable sequence. Show that the interpolated signal  $x(t) = \sum_{n \in \mathbb{Z}} x_n \text{sinc}(Rt - n)$  where  $R$  is a positive real number is square integrable.

# Chapter 6

## Discrete time systems

We have shown that a number of electrical and mechanical devices can be modelled using linear differential equations with constant coefficient. These equations themselves can be used to describe linear time invariant systems. By using the Laplace transform the transfer function of these systems can be found. One particular linear time invariant system has so far been absent. This is the time shifter  $T_\tau$  with non zero time shift  $\tau \neq 0$ . These systems describe pure delays or advances in time of a signal.

We now consider systems constructed from combinations of time shifters of the form  $T_{Pn}$  where  $n \in \mathbb{Z}$  and  $P$  is a positive real number called the **sample period**. Such systems are called **discrete time systems**. We will actually only be interested in a particular type of discrete time system that takes the form

$$H(x) = \sum_{n \in \mathbb{Z}} h_n T_{Pn}(x).$$

where  $h$  is a sequence of real or complex numbers with elements denoted by  $h_n$  for  $n \in \mathbb{Z}$ .

Let  $h$  be a sequence of complex numbers indexed by  $\mathbb{Z}$ . The  $n$ th element in the sequence is denoted by  $h_n$ . Such a sequence is called a **discrete time signal**. Let  $P$  be a positive real number. A system of the form

$$H(x) = \sum_{n \in \mathbb{Z}} h_n T_{Pn}(x).$$

is called a **discrete time system** with period  $P$ . The discrete time signal  $h$  is called the **discrete impulse response** of  $H$ .

### 6.1 Discrete time signals (complex valued sequences)

A **sequence**  $x$  is a function with domain given by the integers  $\mathbb{Z}$ . The value of the sequence corresponding with the integer  $n$  could be denoted by  $x(n)$

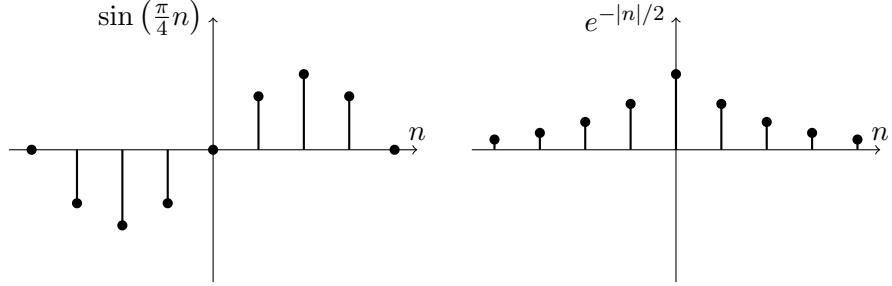


Figure 6.1: 1-dimensional real valued sequences (discrete time signals).

but we will prefer the common notation  $x_n$ . A **discrete time signal** is a sequence that takes real or complex values, that is,  $x_n \in \mathbb{R}$  or  $x_n \in \mathbb{C}$ . For example,

$$\sin(\frac{\pi}{4}n), \quad n^3, \quad e^{-|n|/2}$$

both represent discrete time signals. Discrete time signals are plotted using vertical lines with dotted ends as in Figure 6.1. It is common to use square bracket notation  $x[n]$  to represent the value at or  $x$  at  $n$ , but we will prefer the subscript notation  $x_n$ . This notation is also common [Vetterli et al., 2014; Rudin, 1986]. Discrete time signals have a number of properties analogous to the properties of (continuous time) signals that were described in Section 1.1.

A discrete time signal  $x$  is bounded if there exists a real number  $M$  such that

$$|x_n| \leq M \quad \text{for all } n \in \mathbb{Z}$$

Both  $\sin(\frac{\pi}{4}n)$  and  $e^{-|n|/2}$  are examples of bounded discrete time signals, but,  $n^3$  is not bounded because its magnitude grows indefinitely as  $n$  moves away from the origin.

A discrete time signal  $x$  is periodic if there exists a positive integer  $T$  such that

$$x_n = x_{n+kT} \quad \text{for all integers } k \text{ and } n.$$

The smallest such positive  $T$  is called the period. The discrete time signal  $\sin(\frac{\pi}{4}n)$  is periodic with period  $T = 8$ . Neither  $n^3$  or  $e^{-n^2/4}$  are periodic.

A discrete time signal  $x$  is even (or symmetric) if  $x_n = x_{-n}$  for all  $n \in \mathbb{Z}$  and odd (or antisymmetric) if  $x_n = -x_{-n}$  for all  $n \in \mathbb{Z}$ . Both  $\sin(\frac{\pi}{4}n)$  and  $n^3$  are odd and  $e^{-|n|/2}$  is even.

A discrete time signal  $x$  is **absolutely summable** if

$$\|x\|_1 = \sum_{n \in \mathbb{Z}} |x_n| < \infty,$$

that is, if the sum of absolute values of the elements converges to a finite number. The real number  $\|x\|_1$  is commonly called the  $\ell^1$ -norm of  $x$ . The

discrete time signals  $\sin(\frac{\pi}{4}n)$  and  $n^3$  are not absolutely summable, but  $e^{-|n|/2}$  is because (Exersize 5.8)

$$\sum_{n \in \mathbb{Z}} |e^{-|n|/2}| = \sum_{n \in \mathbb{Z}} e^{-|n|/2} = 1 + \frac{2}{\sqrt{e} - 1}.$$

It is common to denote the set of absolutely summable discrete time signals by  $\ell^1$ . So,  $e^{-|n|/2} \in \ell^1$  and  $\sin(\frac{\pi}{4}n) \notin \ell^1$ .

A discrete time signal  $x$  is **square summable** if

$$\|x\|_2^2 = \sum_{n \in \mathbb{Z}} |x_n|^2 < \infty,$$

that is, if the sum of squared magnitudes of the elements converges to a finite number. The real number  $\|x\|_2$  is commonly called the  $\ell^2$ -norm or **energy** of  $x$ . The discrete time signals  $\sin(\frac{\pi}{4}n)$  and  $n^3$  are not square summable, but  $e^{-|n|/2}$  is because (Excersize 5.8)

$$\sum_{n \in \mathbb{Z}} |e^{-|n|/2}|^2 = \sum_{n \in \mathbb{Z}} e^{-|n|} = 1 + \frac{2}{e - 1}.$$

It is common to denote the set of square summable discrete time signals by  $\ell^2$ . So,  $e^{-|n|/2} \in \ell^2$  and  $\sin(\frac{\pi}{4}n) \notin \ell^2$ . If a discrete time signal is absolutely summable then it is also square summable (Exersize 5.9). The corresponding property is not true of (continuous time) signals, that is, absolutely integrable signals are not necessarily square integrable (Excersize 1.4).

## 6.2 The discrete time impulse response

Discrete time systems are linear combinations of time shifters  $T_{Pn}$ . Because the time shifter is linear and time invariant it follows that discrete time systems are also linear and time invariant. The time shifter is not a regular system and so discrete time systems are not regular, they do not have an impulse response. However, many of the properties

Let  $H_1$  and  $H_2$  be linear time invariant discrete time systems with discrete impulse responses  $h_1$  and  $h_2$ .

composition

discrete convolution

System is stable if and only if discrete impulse response  $h_n$  is absolutely summable.

### 6.3 The z-transform and the discrete time Fourier transform

From (4.1.1) the transfer function of the linear time invariant discrete time system  $H$  is

$$\lambda(H, s) = \sum_{n \in \mathbb{Z}} h_n \lambda(T_{Pn}) = \sum_{n \in \mathbb{Z}} h_n e^{-sPn}.$$

It is convenient define the complex number  $z = e^{Ps}$  and define

$$\mathcal{Z}(h, z) = \lambda(H)(\frac{1}{P} \log z) = \sum_{n \in \mathbb{Z}} h_n z^{-n}.$$

This is called the **z-transform** of the discrete time signal  $h$ . Thus, the transfer function of a discrete time signal is related with z-transform of its discrete impulse response. This relationship is analogous to the relationship between the Laplace transform and the impulse response of a regular linear time invariant system.

The spectrum of the linear time invariant discrete time system  $H$  is

$$\Lambda(H, f) = \lambda(H, j2\pi f) = \sum_{n \in \mathbb{Z}} h_n e^{-2\pi j f P n}.$$

The spectrum is periodic with period  $\frac{1}{P}$ . Define

$$\mathcal{F}_d(h, f) = \Lambda(H, \frac{f}{P}) = \sum_{n \in \mathbb{Z}} h_n e^{-2\pi j f n}$$

This is called the discrete time Fourier transform of the discrete time signal  $h$ .

### 6.4 Finite impulse response filters

### 6.5 Difference equations

### 6.6 Exercises

6.1.

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