# Fakultät für Mathematik und Naturwissenschaften Fachgruppe Mathematik und Informatik 

# System Theoretical Properties of linear port-Hamiltonian Systems on Infinite-dimensional Spaces 

Dissertation zu Erlangung des akademischen Grades Doktor der Naturwissenschaften (Dr. rer. nat.)

vorgelegt von
Julia Theresa Kaiser
aus Haan

Februar, 2021
betreut durch
Prof. Dr. Birgit Jacob

The PhD thesis can be quoted as follows:
urn:nbn:de:hbz:468-20210325-145302-3
[http://nbn-resolving.de/urn/resolver.pl?urn=urn\%3Anbn\%3Ade\%3Ahbz\%3A468-20210325-145302-3]
DOI: 10.25926/wp3p-y863
[https://doi.org/10.25926/wp3p-y863]

## Contents

1 Introduction ..... 5
2 Preliminaries ..... 9
2.1 Preliminaries on evolution equations ..... 10
2.2 Preliminaries on systems theory ..... 14
3 Introduction to port-Hamiltonian systems ..... 23
3.1 Class of port-Hamiltonian systems ..... 23
3.2 Generation theorems ..... 27
3.3 Boundary control and observation port-Hamiltonian systems ..... 31
3.4 Spectrum of port-Hamiltonian systems with ..... $\square$
$P_{1} \mathcal{H}(\zeta)$ diagonal
$P_{1} \mathcal{H}(\zeta)$ diagonal ..... 36 ..... 36
4 Exact controllability of port-Hamiltonian systems ..... 39
4.1 Sufficient condition for exact controllability ..... 39
4.2 Closing remarks and open problems ..... 44
5 Riesz bases of port-Hamiltonian systems ..... 45
5.1 Preliminaries of bases ..... 46
5.1.1 Toy examples ..... 48
5.2 Discrete Riesz spectral operators ..... 50
5.3 Discrete Riesz spectral port-Hamiltonian operators ..... 56
5.3.1 Proof of the Main Result ..... 57
5.3.2 $\quad$ Proof of the equivalence 2$) \Leftrightarrow 3$ ) of Theorem 5.3 .3 ..... 57
5.3.3 Proof of the implication 2 ) $\Rightarrow 1$ ) of Theorem 5.3 .3 ..... 57
5.3.4 Proof of the implication 1$) \Rightarrow 2$ ) of Theorem 5.3 .3 ..... 61
5.4 Examples ..... 65
5.4.1 Wave equation with boundary feedback ..... 65
5.4.2 Timoshenko beam with boundary damping ..... 66
5.5 Closing remarks and open problems ..... 68
6 Generalization of port-Hamiltonian systems ..... 75
6.1 Port-Hamiltonian systems in the infinite-dimensional setting ..... 76
6.1.1 Examples for port-Hamiltonian systems in the infinite-dimensional setting86
6.2 Port-Hamiltonian systems on the semi-axis ..... 87
6.2.1 Examples for port-Hamiltonian systems on the semi-axis. ..... 91
6.3 Closing remarks and open problems . . . . . . . . . . . . . . . . 92

Bibliography 101
Acknowledgement 103

## Chapter 1

## Introduction

The interest in describing dynamics, e.g. the vibration of strings and beams, started in the 17th century by the publication of Newton's Philosophia Naturalis Principia Mathematica, New87. Since then the questions of how to model a system arises and in [CCD81 the beginning of the vibration theory is described from the first mathematical formulations by Isaac Newton and Leonhard Euler. These developments are summarized by Joseph-Louis de Lagrange in Mécanique Analytique, see Lag11. The port-Hamiltonian formulation is an extension of the Hamilton formalism, which was introduced by Hamilton. The Hamilton formalism is a further development of the Lagrange formalism. In both formalism the idea is to start from the kinetic and the potential energy to get the partial differential equation model of the system, see Lan12. Up to now a model is always just an approximation of the reality and one way, on which this thesis is based, is the port-Hamiltonian way of modelling, see DMSB09. Port-based network modeling of complex physical systems leads to port-Hamiltonian systems. Therefore, we introduce modeling in the port-Hamiltonian framework. Here, the idea is to use an energy-based perspective by modeling physical systems. The idea is that a physical system can be viewed as the interconnection of simpler systems, which exchange energy. This structure implies that portHamiltonian systems are closed under power conserving interconnections. The important role of the energy is taken into account with the introduction of the energy norm and the state space as energy space. Therefore, one introduce power conjugated variables, which are connected via the Bond graphs. These are introduced in Bre82 and lead to the introduction of Dirac structures, see for example [DvdS99]. The power conjugated variables are denoted by flow and effort and their product equals power. In this thesis we restrict ourselves to the introduction of port variables, see the introduction of the boundary flow and boundary effort in Chapter 3 .
The advantages of the port-Hamiltonian approach is on the one hand that the model comes from differential geometry and so it is useful for model reduction and on the other hand it fits for a functional analysis approach and therefore also systems theory.
For finite-dimensional systems there is by now a well-established theory vdS06, EMvdS07, DMSB09. The port-Hamiltonian approach has been extended to the
infinite-dimensional situation by a geometric differential approach vdSM02, MM05, JvdS09, ZLMV10] and also by a functional analytic approach Vil07, ZLMV10, JZ12, JMZ15, Aug16, JZ18]. In this thesis we take the functional analytic point of view. This approach has been successfully used to derive simple verifiable conditions for well-posedness LGZM05, Vil07, ZLMV10, JZ12, JMZ15, JK19b, stability [JZ12, AJ14 and stabilization RZLG17, RLGMZ14, AJ14, SZ18 and robust regulation [HP18].
The port-Hamiltonian systems considered in this thesis can be formulated as a partial differential equation

$$
\frac{\partial}{\partial t} x(\zeta, t)=P_{1} \frac{\partial}{\partial \zeta}(\mathcal{H} x)(\zeta, t)+P_{0}(\mathcal{H} x)(\zeta, t), t \geqslant 0, \zeta \in(0,1) .
$$

Also for the more general class of port-Hamiltonian systems, which we consider in Chapter 6, a similar partial differential equation describes the system.
This class of partial differential equations covers (coupled) wave and beam equations and in particular infinite networks of these equations, that means a network with an infinite number of edges.
A functional analytic approach to the partial differential equation is the formulation as an abstract Cauchy problem.

$$
(\mathrm{ACP})\left\{\begin{array}{l}
\dot{x}(t)=A x(t), \quad t \geqslant 0, \\
x(0)=x_{0} .
\end{array}\right.
$$

There has been an enormous development in the study of the Cauchy problem (ACP) and its well-posedness, see for example BC16, Eng13, JZ12, LGZM05, vdSM02, Vil07, ZLMV10 and the references therein. These systems are also known as port-Hamiltonian systems, Hamiltonian partial differential equations or systems of linear conservation laws.
For more information we refer to [JZ12, $\mathrm{JZ18}]$. In the following we denote by port-Hamiltonian systems infinite-dimensional linear port-Hamiltonian systems. Thus, having the results of well-posedness, i.e. existence of mild solutions in mind, this thesis answers the further questions:

1. Which port-Hamiltonian systems are exactly controllable?
2. Which port-Hamiltonian operators are discrete Riesz spectral operators?
3. How can well-posedness of infinite-dimensional systems of port-Hamiltonian system characterized? And what is about well-posedness of portHamiltonian systems on the semi-axis?

In the following, we give a brief overview on this thesis. To introduce portHamiltonian systems and port-Hamiltonian operators in Chapter 3, we recall some basics of functional analysis, strongly continuous semigroups and evolution equations, and systems theory in Chapter 2. This Chapter is mainly based on Wer00, EN00, TW09, JZ12, RW94, and Wei94. In this thesis we focus on port-Hamiltonian systems on a one-dimensional spatial domain.
With this basics in mind, we start in Chapter 3 with the definition of portHamiltonian systems and port-Hamiltonian operators following the functional
analytic approach. In Chapter 3, 4 and 5 we focus on port-Hamiltonian systems without internal damping on a finite interval. This is a special class of portHamiltonian systems, which however is rich enough to cover in particular the wave equation, the transport equation and the Timoshenko beam equation, and also coupled beam and wave equations each with possibly damping on the boundary. For more information on this class of port-Hamiltonian systems we refer to the monograph $[\mathrm{JZ12}$ and the survey [JZ18].
After introducing the class of port-Hamiltonian systems in 3.1 , in 3.2 we formulate generation theorems for this class of systems and in 3.3 we formulate portHamiltonian systems as boundary control and observation systems. All these definitions and results are motivated and illustrated by examples as the wave equation. The results listed there can mostly be found in LGZM05, Vil07, [ZLMV10], JZ12], [JMZ15], Aug16] or [JZ18]. In Section 3.4 we give a new result about the location of the spectrum of a special class of port-Hamiltonian system. In Chapter 4 we consider port-Hamiltonian systems with full boundary control and without internal damping. The main result shows that well-posed port-Hamiltonian systems, with state space $L^{2}\left(0,1 ; \mathbb{C}^{d}\right)$ and input space $\mathbb{C}^{d}$, are exactly controllable and is published in JK19a.
In Chapter 5 the Riesz basis property of port-Hamiltonian systems is studied. Here, we do not follow the ideas of Tre00a, Tre00b] as Villegas in Vil07 but we combine results of systems theory and complex analysis. In Section 5.1 we give a general introduction in bases of Hilbert spaces and in Section 5.2 we give a general characterization of discrete Riesz spectral operators and their properties. In the following Section 5.3 we specify our ideas in the port-Hamiltonian setting and without any technical condition, we give a characterization for the Riesz basis property and show that this is equivalent to the fact that system operator generates a strongly continuous group. Moreover, we get in this situation some more information about the location of the spectrum: Then, the spectrum consists of eigenvalues only, located in a strip parallel to the imaginary axis and they can decomposed into finitely many sets having each a uniform gap. The results of this chapter are published in JKZ20.
In Chapter 6 we consider generalizations of the port-Hamiltonian systems studied so far. We allow port-Hamiltonian systems with internal damping and consider two kinds of generalizations. In Section 6.1 we consider infinitedimensional networks of infinite-dimensional port-Hamiltonian systems on a finite interval and in Section 6.2 we consider infinite-dimensional port-Hamiltonian systems on the semi-axis. This class includes in particular infinite networks of transport, wave and beam equations, or even combinations of these. We formulate equivalent conditions for contraction $C_{0}$-semigroup generation and these results can be found in JK19a.

## Chapter 2

## Preliminaries

In this chapter we introduce some basic notations and ideas of functional analysis, evolution equations and systems theory.
In the following $X$ and $Y$ will always be complex and separable Hilbert spaces. We denote the space of all bounded linear operators from $X$ to $Y$ by $\mathcal{L}(X, Y)$. To shorten notation we write $\mathcal{L}(X):=\mathcal{L}(X, X)$.
We use the notation $s-A:=s I-A$, where $I$ denotes the identity operator, and define the resolvent set of a linear operator $A: \mathcal{D}(A) \subset X \rightarrow X$ as

$$
\rho(A):=\{s \in \mathbb{C} \mid s-A: \mathcal{D}(A) \subset X \rightarrow X \text { is bijective }\} .
$$

For each $s \in \rho(A)$ we denote the resolvent operator of $A$ by $(s-A)^{-1}: X \rightarrow$ $\mathcal{D}(A)$. The spectrum of $A$ is defined as the set $\sigma(A):=\mathbb{C} \backslash \rho(A)$. The point spectrum $\sigma_{p}(A)$ is defined by

$$
\sigma_{p}(A)=\{s \in \mathbb{C} \mid \exists x \in \mathcal{D}(A), x \neq 0, A x=s x\},
$$

and consists of eigenvalues of $A$. We note that in general $\sigma_{p}(A) \subsetneq \sigma(A)$.
Definition 2.0.1. (Wer00, Definition V.5.1]) Let $A: \mathcal{D}(A) \subset X \rightarrow X$ be a densely defined linear operator. Then the (Hilbert space) adjoint operator $A^{*}$ of $A$ is defined as

$$
\begin{aligned}
D\left(A^{*}\right) & :=\{y \in X \mid \exists w \in X, \forall z \in \mathcal{D}(A) \text { such that }\langle A z, y\rangle=\langle z, w\rangle\} ; \\
A^{*} y & :=w .
\end{aligned}
$$

The operator $A$ is called self-adjoint if $A^{*}=A$, and skew-adjoint if $A^{*}=-A$. Note that $A^{*}=A$ implies $\mathcal{D}(A)=\mathcal{D}\left(A^{*}\right)$ in particular.

A further important property of a linear operator is dissipativity.
Definition 2.0.2. The operator $A: \mathcal{D}(A) \subset X \rightarrow X$ is called dissipative if $\operatorname{Re}\langle A x, x\rangle \leqslant 0$ for every $x \in \mathcal{D}(A)$.

Moreover, we introduce the Sobolev spaces $\mathcal{W}^{m, 2}(I)$ for $I \subset \mathbb{R}$ open interval and $m \in \mathbb{N}$. For this purpose, we define the weak derivative of a function $f \in L^{2}(I)$, which is a generalization of the derivative.

Definition 2.0.3. ([Wer00, Definition V.1.11]) Let $I \subset \mathbb{R}$ be an open interval and $m \in \mathbb{N}$. A function $f \in L^{2}(I)$ is $m$-times weakly differentiable if there exists a function $g \in L^{2}(I)$, also denoted by $\frac{d^{m}}{d x^{m}} f$, such that

$$
\begin{equation*}
\langle g, \varphi\rangle=(-1)^{m}\left\langle f, \frac{d^{m}}{d x^{m}} \varphi\right\rangle \quad \forall \varphi \in \mathcal{D}(I), \tag{2.1}
\end{equation*}
$$

where

$$
\mathcal{D}(I):=\left\{\varphi \in C^{\infty}(I) \mid \operatorname{supp}(\varphi):=\overline{\{x: \varphi(x) \neq 0\}} \subset I \text { is compact }\right\}=\mathcal{C}_{c}^{\infty}(I)
$$

denotes the set of $C^{\infty}$-functions with compact support. $\mathcal{D}(I)$ is also called the set of test functions.

Now we are in the situation to give the definition of Sobolev spaces.
Definition 2.0.4. Let $I \subset \mathbb{R}$ be an open interval and $m \in \mathbb{N}$. Then the Sobolev space of $m$-th order over $I$ is given by

$$
\mathcal{W}^{m, 2}(I):=\left\{f \in L^{2}(I) \left\lvert\, \frac{d^{k}}{d x^{k}} f \in L^{2}(I)\right. \text { exists for all } k \leqslant m\right\}
$$

with norm

$$
\|f\|_{m, 2}=\left(\sum_{0 \leqslant k \leqslant m}\left\|\frac{d^{k}}{d x^{k}} f\right\|_{L^{2}(I)}^{2}\right)^{\frac{1}{2}}
$$

Thus, the Sobolev space of first order on the interval $(0,1)$ with values in $\mathbb{C}^{d}$ is given by

$$
\mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{d}\right)=\left\{f \in L^{2}\left(0,1 ; \mathbb{C}^{d}\right) \left\lvert\, \frac{d}{d x} f \in L^{2}\left(0,1 ; \mathbb{C}^{d}\right)\right. \text { exists }\right\}
$$

A more detailed introduction into Sobolev spaces can be found in Ada75.

### 2.1 Preliminaries on evolution equations

Within this section we give a short overview of the theory of evolution equations, i.e., equations which describe the development of a system in time. We consider the following so called abstract Cauchy problem

$$
(\mathrm{ACP})\left\{\begin{array}{l}
\dot{x}(t)=A x(t), \quad t \geqslant 0  \tag{2.2}\\
x(0)=x_{0}
\end{array}\right.
$$

where $A: \mathcal{D}(A) \subset X \rightarrow X$ denotes a closed and densely defined linear operator. This leads to the question whether (2.2) has a unique solution. This property is also known as the well-posedness of (2.2). For a bounded operator $A \in \mathcal{L}(X)$ or a matrix $A \in \mathbb{C}^{n \times n}$ the solution of (ACP) is given by $x(t)=\mathrm{e}^{t A} x_{0}$ where

$$
\begin{equation*}
\mathrm{e}^{t A}:=\sum_{n=1}^{\infty} \frac{(t A)^{n}}{n!}, \quad t \geqslant 0 . \tag{2.3}
\end{equation*}
$$

However, (2.3) does not make sense for general unbounded operators $A$.
In the following we introduce the concepts for solutions of (ACP), c.f. EN00, Section II.6].

Definition 2.1.1. Let $x:[0, \infty) \rightarrow X$ be a continuous function. Then:

1. The function $x$ is called classical solution of (ACP) if $x$ is differentiable, $x(t) \in \mathcal{D}(A)$ for all $t \geqslant 0$ and $x$ satisfies equation (ACP).
2. The function $x$ is called mild solution of (ACP) if $\int_{0}^{t} x(s) d s \in \mathcal{D}(A)$ and

$$
\begin{equation*}
x\left(t, x_{0}\right):=x(t)=x_{0}+A \int_{0}^{t} x(s) d s, t \geqslant 0 . \tag{2.4}
\end{equation*}
$$

Using the concepts of solution we can now introduce well-posedness for the abstract Cauchy problem.
Definition 2.1.2. The abstract Cauchy problem (ACP) is well-posed if

1. $\mathcal{D}(A)$ is dense in $X$;
2. For every $x_{0} \in \mathcal{D}(A)$ there exists a unique classical solution;
3. For every sequence $\left(x_{n}\right)_{n \in \mathbb{N}} \subset \mathcal{D}(A)$ with $\lim _{n \rightarrow \infty} x_{n}=0$ it holds $\lim _{n \rightarrow \infty} x\left(t ; x_{n}\right)=0$ uniformly on compact intervals $\left[0 ; t_{0}\right]$.

Well-posedness of $(A C P)$ is closely related to the notion of $C_{0}$-semigroups, see also Theorem 2.1.8. They can be seen as a generalization of the exponential function. $C_{0}$-semigroups goes back to the work of Hille and Yoshida Hil48 and Yos48] and are studied in detail in the monographs by Engel and Nagel EN00], Pazy Paz83 and Goldstein Gol85.
Definition 2.1.3. A family $(T(t))_{t \geqslant 0} \in \mathcal{L}(X)$ of bounded operators is called a strongly continuous (operator) semigroup, or $C_{0}$-semigroup for short, if it has the following properties:

1. $T(t+s)=T(s) T(t)$ for all $t, s \geqslant 0$,
2. $T(0)=I$,
3. $\lim _{t \searrow 0}\|T(t) x-x\|=0$ for all $x \in X$.

If property 1 . holds for all $t, s \in \mathbb{R}$, the family $(T(t))_{t \geqslant 0} \in \mathcal{L}(X)$ is called a strongly continuous group, or $C_{0}$-group for short.
The strong continuity implies that $T(\cdot) x \in C\left(\mathbb{R}_{+}, X\right)$ for every $x \in X$.
The following example shows that the notation of $C_{0}$-semigroups is also consistent for bounded operators $A$.
Example 2.1.4. Let $A \in \mathbb{C}^{n \times n}$ or $A \in \mathcal{L}(X)$. Then $T(t)=\mathrm{e}^{t A}, t \geqslant 0$, is a $C_{0}$-semigroup and even a $C_{0}$-group.
For a $C_{0}$-semigroup $(T(t))_{t \geqslant 0}$ we define its generator.
Definition 2.1.5. The operator $A: \mathcal{D}(A) \subset X \rightarrow X$ with

$$
\begin{aligned}
A x & :=\lim _{t \searrow 0} \frac{T(t) x-x}{t} \\
\mathcal{D}(A) & =\left\{x \in X \left\lvert\, \lim _{t \searrow 0} \frac{T(t) x-x}{t}\right. \text { exists in } X\right\}
\end{aligned}
$$

is called the (infinitesimal) generator of the $C_{0}$-semigroup $(T(t))_{t \geqslant 0}$.

In the following we mention some of the important properties of generators of $C_{0}$-semigroups.

Lemma 2.1.6. ([EN00, Lemma II.1.3]) Let $A$ be the generator of a $C_{0}$-semigroup $(T(t))_{t \geqslant 0}$. Then the following holds:

1. $x \in \mathcal{D}(A)$ implies $T(t) x \in \mathcal{D}(A)$ and

$$
\frac{d}{d t} T(t) x=T(t) A x=A T(t) x \quad \forall t \geqslant 0 .
$$

2. For every $x \in X$ and every $t \geqslant 0$ it holds $\int_{0}^{t} T(s) x d s \in \mathcal{D}(A)$ and

$$
T(t) x-x=A \int_{0}^{t} T(s) x d s
$$

3. For every $x \in \mathcal{D}(A)$ and every $t \geqslant 0$ it holds

$$
T(t) x-x=\int_{0}^{t} T(s) A x d s
$$

Proposition 2.1.7. ([EN00, Theorem II.1.4]) Let $(T(t))_{t \geqslant 0}$ be a $C_{0}$-semigroup on $X$ with generator $A$. Then $A$ is linear, closed, densely defined and determines the $C_{0}$-semigroup $(T(t))_{t \geqslant 0}$ uniquely.

The next theorem describes the relationship between the well-posedness of the abstract Cauchy problem (ACP) and the generator of a $C_{0}$-semigroup and is a combination of Corollary II.6.9 and Proposition II.6.2 in EN00].

Theorem 2.1.8. Let $A: \mathcal{D}(A) \subset X \rightarrow X$ be a closed linear operator. Then the following statements are equivalent:

1. The abstract Cauchy problem (2.2) is well-posed.
2. A generates a $C_{0}$-semigroup on $X$.

In particular, for every $x_{0} \in \mathcal{D}(A)$ the unique classical solution of (2.2) is given by $x(t):=T(t) x_{0}$.

In the following, we mention some properties of $C_{0}$-semigroups, which can be found in EN00.

Proposition 2.1.9. (EN00, Proposition I.5.5]) For a $C_{0}$-semigroup $(T(t))_{t \geqslant 0}$ on $X$ there exist constants $M \geqslant 1$ and $\omega \in \mathbb{R}$ such that

$$
\begin{equation*}
\|T(t)\| \leqslant M \mathrm{e}^{\omega t}, \quad t \geqslant 0 . \tag{2.5}
\end{equation*}
$$

Definition 2.1.10. Let $(T(t))_{t \geqslant 0}$ be a $C_{0}$-semigroup on $X$. Then $(T(t))_{t \geqslant 0}$ is

- a bounded $C_{0}$-semigroup if there exists $M>0$ such that $\|T(t)\| \leqslant M$,for all $t \geqslant 0$;
- a contractive $C_{0}$-semigroup, if it is bounded with $M=1$, i.e., $\|T(t)\| \leqslant 1$.

Moreover, a $C_{0}$-group $(T(t))_{t \in \mathbb{R}}$ is called a unitary group, if $\|T(t) x\|=\|x\| \forall x \in$ $X$ and $t \in \mathbb{R}$.

Definition 2.1.11. Let $(T(t))_{t \geqslant 0}$ be a $C_{0}$-semigroup on $X$ with generator $A$. Then its growth bound is defined by

$$
\begin{equation*}
\omega_{0}(A):=\inf \left\{\omega \in \mathbb{R} \mid \exists M_{\omega} \geqslant 1 \text { such that }\|T(t)\| \leqslant M_{\omega} \mathrm{e}^{\omega t}, t \geqslant 0\right\} . \tag{2.6}
\end{equation*}
$$

Furthermore,

$$
\begin{equation*}
s(A):=\sup \{\operatorname{Re} s \mid s \in \sigma(A)\} \tag{2.7}
\end{equation*}
$$

denotes the spectral bound of $A$.
If the growth bound $\omega_{0}$ is negative, then the corresponding $C_{0}$-semigroup is called exponentially stable. We note that for an exponentially stable $C_{0}$-semigroup the right half plane $\{s \in \mathbb{C} \mid \operatorname{Re} s>0\}$ of $\mathbb{C}$ lies in the resolvent set $\rho(A)$ of its generator $A$, c.f. EN00, Theorem V.1.11].
Now, we describe some properties of the resolvent operator.
Proposition 2.1.12. ([EN00, Theorem II.1.10]) Let $A$ be the generator of a $C_{0}$-semigroup $(T(t))_{t \geqslant 0}$ with growth bound $\omega_{0} \in \mathbb{R}$ and $\omega \in \mathbb{R}, M \geqslant 1$ are the constants described in Proposition 2.1.9. Then for all $s \in \mathbb{C}$ with $\operatorname{Re} s>\omega_{0}$, it holds that $s$ lies in the resolvent set of $A$, i.e., $s \in \rho(A)$ and the integral

$$
\begin{equation*}
R(s):=\int_{0}^{\infty} \mathrm{e}^{-s t} T(t) x d s \tag{2.8}
\end{equation*}
$$

exists for all $x \in X$ and $R(s)=(s-A)^{-1}$. Moreover, the following estimate for the resolvent holds:

$$
\begin{equation*}
\left\|(s-A)^{-1}\right\| \leqslant \frac{M}{\operatorname{Re} s-\omega} \text { for all } \operatorname{Re} s \geqslant \omega \tag{2.9}
\end{equation*}
$$

Definition 2.1.13. The linear operator $A: \mathcal{D}(A) \subset X \rightarrow X$ has compact resolvent if there exists an $s \in \rho(A)$ such that the operator $(s-A)^{-1}$ is a compact operator.

Note that by EN00, Proposition II.4.25] $A$ has compact resolvent if the embedding of $D(A)$ equipped with the graph norm in $X$ is compact.
In Theorem 2.1.8 we have seen the relation between the abstract Cauchy problem (ACP) and the corresponding $C_{0}$-semigroup. Thus, the question occurs under which conditions $A$ generates a contraction $C_{0}$-semigroup. It is answered by Hille and Yoshida in 1948, c.f. Hil48 and Yos48, and reformulated in a more applicable way by Lumer and Phillips in 1961, LP61, which we state here.

Theorem 2.1.14. (EN00, Theorem II.3.15]) Let $A: \mathcal{D}(A) \subset X \rightarrow X$ be a linear, densely defined, and closed operator on a Hilbert space $X$. Then $A$ generates a contraction $C_{0}$-semigroup on $X$ if and only if

1. $A$ is dissipative and
2. $s-A: \mathcal{D}(A) \subset X \rightarrow X$ is surjective for one (and then for all) $s>0$.

A simpler characterization of generators of contraction $C_{0}$-semigroups is given in the following corollary.

Corollary 2.1.15. ([|EN00, Corollary II.3.17]) Let $A: \mathcal{D}(A) \subset X \rightarrow X$ be a linear, densely defined, and closed operator on a Hilbert space $X$. Then $A$ generates a contraction $C_{0}$-semigroup on $X$ if and only if $A$ and $A^{*}$ are dissipative.

We now formulate the Theorem of Stone, which characterizes generators of unitary groups.

Theorem 2.1.16. ([EN00, II.3.24]). Let $A: \mathcal{D}(A) \subset X \rightarrow X$ be a linear, densely defined, and closed operator on a Hilbert space $X$. Then the following statements are equivalent:

1. A generates a unitary group $(T(t))_{t \in \mathbb{R}}$ on $X$;
2. $A$ is skew-adjoint,
3. $A$ and $-A$ both generate a contraction $C_{0}$-semigroup.

We close this section with the result that the generation of $C_{0}$-semigroups is inherited on closed subspaces. We recall that a closed subspace $V \subset X$ is called $(T(t))_{t \geqslant 0 \text {-invariant }}$ if $T(t) V \subseteq V$ for all $t \geqslant 0$.

Proposition 2.1.17. ([CZ95, Lemma 2.5.3]) Let A generate a $C_{0}$-semigroup $(T(t))_{t \geqslant 0}$ on $X$. In this case the restriction $\left(\left.T(t)\right|_{V}\right)_{t \geqslant 0}$ is again a $C_{0}$-semigroup with generator $\left.A\right|_{V}$ on $V$, where $\left.A\right|_{V}=A v$ for $v \in \mathcal{D}\left(\left.A\right|_{V}\right)=\mathcal{D}(A) \cap V$ and $\left(\left.T(t)\right|_{V}\right)_{t \geqslant 0}$ is generated by the part of $A$ in $V$.

### 2.2 Preliminaries on systems theory

The standard formulation of systems in system theory extends the formulation of a partial differential equation as a (homogenous) abstract Cauchy problem taking into account the interaction of the system with its environment. Thus, in addition to the state space $X$, we need an input space $U$ and an output space $Y$. In general, $X, U$, and $Y$ may be Banach spaces, but within this thesis we focus on the Hilbert space setting. Thus, in the following $X, U$, and $Y$ are supposed to be Hilbert spaces. In this setting the operators in the standard formulation of a system are not necessary bounded in general, but they are bounded in a weaker way. To see this we introduce the extrapolation and the interpolation space.
Therefore, we need to introduce some notation and concepts, which are wellknown and can be found in e.g. EN00, Chapter II] and TW09, Chapter 2].

Definition 2.2.1. Let $A: \mathcal{D}(A) \subset X \rightarrow X$ be the generator of a $C_{0}$-semigroup $(T(t))_{t \geqslant 0}$. Hence, $\rho(A) \neq \emptyset$ and let be $s \in \rho(A)$.

1. Then the extrapolation space $X_{-1}$ is defined as the completion of $X$ with respect to the norm

$$
\|x\|_{-1}=\left\|(s-A)^{-1} x\right\|, \quad x \in X .
$$

2. The interpolation space $X_{1}$ is defined as $\mathcal{D}(A)$ equipped with the norm

$$
\|x\|_{1}=\|(s-A) x\|, \quad x \in X .
$$

Note that the definitions of $X_{1}$ and $X_{-1}$ are independent of the choice of $s \in$ $\rho(A)$, since different $s \in \rho(A)$ lead to equivalent norms. The following inclusions are dense with a continuous embedding:

$$
X_{1} \subset X \subset X_{-1} .
$$

Note that the space $X_{1}$ is a Hilbert space and $A$ can be seen as an operator in $\mathcal{L}\left(X_{1}, X\right)$. Then we consider the restriction and the continuous extension of $(T(t))_{t \geqslant 0}$ to the interpolation and the extrapolation space, respectively.

Proposition 2.2.2. Let $A$ be the generator of a $C_{0}$-semigroup on $X$. Then the following statements hold:

1. $\left(T_{1}(t)\right)_{t \geqslant 0}$, the restriction of $(T(t))_{t \geqslant 0}$ to $X_{1}$, is a $C_{0}$-semigroup on $X_{1}$, with generator

$$
A_{1} x=A x, x \in \mathcal{D}\left(A_{1}\right), \mathcal{D}\left(A_{1}\right)=\mathcal{D}\left(A^{2}\right) .
$$

2. $\left(T_{-1}(t)\right)_{t \geqslant 0}$, the continuous extension of $(T(t))_{t \geqslant 0}$ to $X_{-1}$, is a $C_{0}$-semigroup on $X_{-1}$, whose generator $A_{-1} \in \mathcal{L}\left(X, X_{-1}\right)$, is the unique bounded extension of $A$.

Moreover, we can identify $X_{-1}$ with the dual space of $\mathcal{D}\left(A^{*}\right)$ with respect to the pivot space $X$, that is $X_{-1}=\mathcal{D}\left(A^{*}\right)^{\prime}$. Now, let $A_{-1} \in \mathcal{L}\left(X, X_{-1}\right)$ be the extension of the operator $A$ describing the dynamics of the system, $B \in \mathcal{L}\left(U, X_{-1}\right)$ denotes the control operator, $C \in \mathcal{L}\left(X_{1}, Y\right)$ the observation operator and $D \in \mathcal{L}(U, Y)$ the bounded feedthrough operator mapping from the input to the output. Then the standard formulation in system theory for a control system $\Sigma(A, B, C, D)$ is given by

$$
\begin{align*}
& \dot{x}(t)=A_{-1} x(t)+B u(t), \quad t \geqslant 0, \quad x(0)=x_{0},  \tag{2.10}\\
& y(t)=C x(t)+D u(t), \quad t \geqslant 0 . \tag{2.11}
\end{align*}
$$

Note that we denote by $\Sigma(A, B)$ a control system $\Sigma(A, B, C, D)$ with $C=D=0$ and by $\Sigma(A, C)$ an observation system $\Sigma(A, B, C, D)$ with $B=D=0$. We consider the first equation (2.10) as an abstract inhomogeneous Cauchy problem on the extrapolation space $X_{-1}$ and we give its mild solution for $x_{0} \in X \subset X_{-1}$.

Definition 2.2.3. For $x_{0} \in X$ and $u \in L^{2}(0, t ; U)$ the mild solution of 2.10 ) is given by the variation of parameters formula

$$
\begin{equation*}
x(t)=T(t) x_{0}+\int_{0}^{t} T_{-1}(t-s) B u(s) d s, \quad t \geqslant 0 \tag{2.12}
\end{equation*}
$$

We note that even for initial values $x_{0} \in X$ the values $x(t)$ of the solution may lie in the extrapolation space $X_{-1}$ : The control operator is a map $B \in \mathcal{L}\left(U, X_{-1}\right)$, $T_{-1}(t)$ acts on $X_{-1}$, and thus $\int_{0}^{t} T_{-1}(t-s) B u(s) d s \in X_{-1}$.
To ensure that the solution $x(t)$ lies in $X$ we introduce the idea of admissibility of control operators following Chapters 4, 6, and 11 in [TW09].

Definition 2.2.4. A control operator operator $B \in \mathcal{L}\left(U, X_{-1}\right)$ is an admissible control operator for $(T(t))_{t \geqslant 0}$ if for all $t \geqslant 0$

$$
\int_{0}^{t} T_{-1}(t-s) B u(s) d s \in X
$$

for every $u \in L^{2}(0, t ; U)$.
Admissibility implies that the mild solution $x$ of 2.10 satisfies $x \in C(0, t ; X)$ for every initial condition $x_{0} \in X$ and every $u \in L^{2}(0, t ; U)$.

Proposition 2.2.5. (TW14, Proposition 4.4.6]) Let $B \in \mathcal{L}\left(U, X_{-1}\right)$ be an admissible control operator for $(T(t))_{t \geqslant 0}$. Then for $\omega>\omega_{0}(A)$ exists a constant $M_{\omega}>0$ such that

$$
\begin{equation*}
\left\|\left(s-A_{-1}\right)^{-1} B\right\|_{\mathcal{L}(U, X)} \leqslant \frac{M_{\omega}}{\sqrt{\operatorname{Re} s-\omega}} \text { for } \operatorname{Re} s \geqslant \omega \tag{2.13}
\end{equation*}
$$

In the same manner as in the motivation of admissibility for control operators, it might happen that for $x \in \mathcal{D}(A)$ the solution $x$ does not lie in the domain of $A$ implying that $C x(t)$ is not well-defined. To avoid this we introduce admissibility for observation operators.

Definition 2.2.6. An observation operator $C \in \mathcal{L}\left(X_{1}, Y\right)$ is an admissible observation operator for $(T(t))_{t \geqslant 0}$ if there exists a positive constant $K>0$ such that

$$
\int_{0}^{\infty}\left\|C T(t) x_{0}\right\|^{2} d t \leqslant K\left\|x_{0}\right\|^{2}, \quad x_{0} \in \mathcal{D}(A)
$$

Definition 2.2.7. A system $\Sigma(A, B)$ with an admissible control operator $B \in$ $\mathcal{L}\left(U, X_{-1}\right)$ is exactly controllable, if there exists a time $\tau>0$ such that for all $x_{1} \in X$ there exists a control function $u \in L^{2}(0, \tau ; U)$ such that the corresponding mild solution satisfies $x(0)=0$ and $x(\tau)=x_{1}$.

Note that this definition of exactly controllable is often also denoted as exactly controllable in finite time.

Definition 2.2.8. A system $\Sigma(A, C)$ with an admissible observation operator $C \in \mathcal{L}\left(X_{1}, Y\right)$ is exactly observable, if there exists a positive constant $k$ such that

$$
\int_{0}^{\tau}\left\|C T(t) x_{0}\right\|^{2} d t \geqslant k\left\|x_{0}\right\|^{2}, \quad x_{0} \in \mathcal{D}(A)
$$

This is equivalent to the fact, that every initial state $x_{0} \in X$ can be uniquely and continuously reconstructed from the output $y \in L^{2}(0, \tau ; Y)$.
Furthermore, we note that the concepts of controllability and observability are dual in the following sense.

Proposition 2.2.9. ([TW09, Theorem 11.2.1]) Let $B \in \mathcal{L}\left(U, X_{-1}\right)$ an admissible control operator for the $C_{0}$-semigroup $(T(t)) \geqslant 0$ generated by $A$. Then the system $\Sigma(A, B)$ is exactly controllable if and only if $\Sigma\left(A^{*}, B^{*}\right)$ is exactly observable.

In the following we formulate the so-called Hautus test giving a necessary condition for exact observability. In RW94 the Hautus test is formulated for systems which are exactly observable in infinite time, i.e., there exists a positive constant $\tilde{k}$ such that $\int_{0}^{\infty}\left\|C T(t) x_{0}\right\|^{2} d t \geqslant \tilde{k}\left\|x_{0}\right\|^{2}, \quad x_{0} \in \mathcal{D}(A)$. Of course, this concept of exact observability follows from our notion of exact observability.

Theorem 2.2.10. ( $\overline{\mathrm{RW} 94}$, Theorem 1]) Let $A$ be the generator of an exponentially stable $C_{0}$-semigroup and let $C \in \mathcal{L}\left(X_{1}, Y\right)$ be an admissible observation operator. If the system $\Sigma(A, C)$ is exactly observable, then there exists a positive constant $m$ such that

$$
\begin{equation*}
\frac{1}{|\operatorname{Re} s|^{2}}\|(s-A) x\|^{2}+\frac{1}{|\operatorname{Re} s|}\|C x\|^{2} \geqslant m\|x\|^{2}, \quad \operatorname{Re} s<0, x \in \mathcal{D}(A) . \tag{2.14}
\end{equation*}
$$

Partial differential equations which can be handled via control and observation of the boundary occur frequently in applications. Therefore, we introduce the so-called boundary control and observation systems. In Theorem 2.2 .22 we will see that these kind of systems even fit in the standard formulation $(2.10)-(2.11)$. The following is extracted from the Chapters 11 and 13 in [JZ12].

Definition 2.2.11. Let $X, U$, and $Y$ denote Hilbert spaces and let $\mathfrak{A}: \mathcal{D}(\mathfrak{A}) \subset$ $X \rightarrow X$ and $\mathfrak{B}: \mathcal{D}(\mathfrak{A}) \rightarrow U$ be linear operators. Then $\mathfrak{S}(\mathfrak{A}, \mathfrak{B})$ is a boundary control system if the following hold:

1. The operator $A: \mathcal{D}(A) \subset X \rightarrow X$ with $\mathcal{D}(A)=\mathcal{D}(\mathfrak{A}) \cap \operatorname{ker}(\mathfrak{B})$ and $A x=$ $\mathfrak{A} x$ for $x \in \mathcal{D}(A)$ is the infinitesimal generator of a strongly continuous semigroup on $X$.
2. There exists a right inverse $\widetilde{B} \in \mathcal{L}(U, X)$ of $\mathfrak{B}$ in the sense that for all $u \in U$ we have $\widetilde{B} u \in \mathcal{D}(\mathfrak{A}), \mathfrak{B} \widetilde{B} u=u$ and $\mathfrak{A} \widetilde{B}: U \rightarrow X$ is bounded.

If it holds additionally for a linear operator $\mathfrak{C}: \mathcal{D}(\mathfrak{A}) \rightarrow Y$ the statement
3. the operator $\mathfrak{C}$ is bounded from $\mathcal{D}(A)$ to $Y$, where $\mathcal{D}(A)$ is equipped with the graph norm of $A$,
then the triple $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ is a boundary control and observation system.
Note that a boundary control system is a boundary control and observation system with $\mathfrak{C}=0$.

Remark 2.2.12. In the literature, see e.g. Staffans [Sta05] or Tucsnak and Weiss [TW09], there exists a slightly more general formulation of boundary control systems taking into account that $\mathcal{D}(\mathfrak{B}) \neq \mathcal{D}(\mathfrak{A})$ and $\mathcal{D}(\mathfrak{C}) \neq \mathcal{D}(\mathfrak{A})$, but then they have to satisfy $\mathcal{D}(\mathfrak{A}) \subset \mathcal{D}(\mathfrak{A})$ and $\mathcal{D}(\mathfrak{A}) \subset \mathcal{D}(\mathfrak{C})$.
We write operators of Definition 2.2 .11 as a system of the following form:

$$
\begin{align*}
\dot{x}(t) & =\mathfrak{A} x(t), \quad x(0)=x_{0}, \\
u(t) & =\mathfrak{B} x(t),  \tag{2.15}\\
y(t) & =\mathfrak{C} x(t), \quad t \geqslant 0 .
\end{align*}
$$

Now we are interested in classical and mild solutions of (2.15).
Definition 2.2.13. Let $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ be a boundary control and observation system, with initial value $x_{0} \in \mathcal{D}(\mathfrak{A})$ and $u \in C^{2}(0, \infty ; U)$. A function $x$ : $[0, \infty) \rightarrow X$ is a classical solution of the boundary control and observation system if $x \in C^{1}(0, \infty ; \mathcal{D}(\mathfrak{A}))$ and $x(t)$ satisfies the first two equations of (2.15) for every $t \geqslant 0$.

Lemma 2.2.14. ([JZ12, Lemma 13.1.5]) Let $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ be a boundary control and observation system and $x_{0} \in \mathcal{D}(\mathfrak{A})$ and $u \in C^{2}(0, t ; U)$ satisfying $\mathfrak{B} x_{0}=$ $u(0)$. Then the unique classical solution on $[0, t]$ of (2.15) is given by

$$
\begin{equation*}
x(t)=T(t) x_{0}+\int_{0}^{t} T(t-s) \mathfrak{A} \tilde{B} u(s) d s-A \int_{0}^{t} T(t-s) \tilde{B} u(s) d s \tag{2.16}
\end{equation*}
$$

where $\tilde{B}$ is described in Definition 2.2.11. This implies

$$
y(t)=\mathfrak{C} T(t) x_{0}+\mathfrak{C} \int_{0}^{t} T(t-s) \mathfrak{A} \tilde{B} u(s) d s-\mathfrak{C} A \int_{0}^{t} T(t-s) \tilde{B} u(s) d s
$$

In general, we are also interested in mild solutions, since the initial value $x_{0}$ might be an arbitrary element of $X$, not necessary in the domain of $\mathfrak{A}$, and we also want to allow arbitrary input functions $u \in L^{2}(0, t ; U)$. Nevertheless, the solution $x$ should be a continuous function with values in $X$ and the output $y$ should be a $L^{2}$-function. This leads to the definition of well-posedness.

Definition 2.2 .15 . The boundary control and observation system $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ is called well-posed if there exist a $t>0$ and $m_{t} \geqslant 0$ such that for every initial value $x_{0} \in \mathcal{D}(\mathfrak{A})$ and every input function $u \in C^{2}(0, t ; U)$ with $u(0)=\mathfrak{B} x_{0}$ the classical solution $x$ and $y$ satisfy

$$
\|x(t)\|_{X}^{2}+\int_{0}^{t}\|y(s)\|^{2} d s \leqslant m_{t}\left(\left\|x_{0}\right\|_{X}^{2}+\int_{0}^{t}\|u(s)\|^{2} d s\right) .
$$

There exists a rich literature on well-posed systems, see e.g. Staffans Sta05] or Tuscnak and Weiss TW14]. In general, it is not easy to verify well-posedness for a boundary control and observation system. Nevertheless, the following proposition allows us to do so for a special class of systems.

Proposition 2.2.16. ([JZ12, Proposition 13.1.4]) Let $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ be a boundary control and observation system (2.15). If every classical solution of (2.15) satisfies

$$
\frac{d}{d t}\|x(t)\|^{2}=\|u(t)\|^{2}-\|y(t)\|^{2}
$$

then the system is well-posed.
Now we formulate the mild solution of a well-posed boundary control system and observation system.

Definition 2.2.17. Let $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ be a well-posed boundary control and observation system with initial value $x_{0} \in X$ and $u \in L^{2}(0, t ; U)$. Then the function $x(t)$ given in (2.16) is called mild solution and $x \in C(0, \infty, X)$.

Lemma 2.2.18. ([JZ12, Lemma 13.1.5]) Let $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ be a well-posed boundary control and observation system with initial value $x_{0} \in X$ and $u \in L^{2}(0, t ; U)$. Then the unique mild solution $x$ is given by

$$
\begin{equation*}
x(t)=T(t) x_{0}+\int_{0}^{t} T_{-1}(t-s)\left(\mathfrak{A} \widetilde{B}-A_{-1} \widetilde{B}\right) u(s) d s, \tag{2.17}
\end{equation*}
$$

where $\tilde{B}$ is described in Definition 2.2.11.
For the class of boundary control and observation systems we introduce the concept of transfer functions following Chapter 12 in [JZ12].

Definition 2.2.19. Let $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ be a boundary control and observation system. For $s \in \rho(A)$ and $u \in U, G(s) u$ is the unique solution of

$$
\begin{aligned}
s x & =\mathfrak{A} x, x_{0} \in \mathcal{D}(A), \\
u & =\mathfrak{B} x, \\
y & =\mathfrak{C} x .
\end{aligned}
$$

Then $y=G(s) u, G(s) \in \mathcal{L}(U, Y)$, and $G: \rho(A) \rightarrow \mathcal{L}(U, Y)$ is called the transfer function of the system $\mathfrak{S}$.
For the following proposition we define the weighted $L^{2}$-spaces. For any Hilbert space $H$ a function $v$ is an element of the weighted $L^{2}$-space $L_{\mu}^{2}(0, \infty ; H)$ if and only if $\mathrm{e}^{-\mu t} v \in L^{2}(0, \infty, H) . \quad L_{\mu}^{2}(0, \infty ; H)$ equipped with the norm $\|v\|_{L_{\mu}^{2}(0, \infty ; H)}:=\left\|\mathrm{e}^{-\mu t} v\right\|_{L^{2}(0, \infty ; H)}$ is a Hilbert space.
Proposition 2.2.20. ( $\|$ JZ12, Theorem 12.1.3] and Wei94, Proposition 4.1 and Proposition 3.2]) The transfer function of a well-posed system $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ is given by

$$
G(s)=\mathfrak{C}(s-A)^{-1}(\mathfrak{A} \widetilde{B}-s \widetilde{B})+\mathfrak{C} \widetilde{B}, \quad s \in \rho(A) .
$$

For a well-posed system there exists a $\mu \geqslant w_{0}$ such that the transfer function equals the Laplace transform of the linear and bounded mapping

$$
\mathbb{L}: L_{\mu}^{2}(0, \infty ; U) \rightarrow L_{\mu}^{2}(0, \infty ; Y)
$$

Furthermore, the transfer function is bounded on the right half plane $\mathbb{C}_{\mu}:=$ $\{s \in \mathbb{C} \mid \operatorname{Re} s>\mu\}$.

However, the boundedness of the transfer function on a right half plane does not imply the convergence along the real axis. But convergence along the real axis implies a suitable representation of the feedthrough operator.

Definition 2.2.21. [JZ12, Definition 13.1.11] A well-posed boundary control and observation system $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ with transfer function $G$ is called regular if $\lim _{s \in \mathbb{R}, s \rightarrow \infty} G(s)$ exists. In this case the feedthrough operator $D$ is defined as

$$
D:=\lim _{s \in \mathbb{R}, s \rightarrow \infty} G(s) .
$$

The next assertion can be found in Chapter 13, Section 1 in [JZ12] and makes the connection between boundary control and observation systems and the standard formulation in system theory.

Theorem 2.2.22. Every regular well-posed boundary control and observation system $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})(2.15)$ can equivalently formulated in the standard formulation (2.10)-(2.11) in system theory with a control operator $B=\left(\mathfrak{A} \widetilde{B}-A_{-1} \widetilde{B}\right)$.

So far, we have only considered open-loop systems, that is, the input $u(t)$ is independent of the output $y(t)$, see Figure 2.1. Systems, where input and output are connected via a feedback law

$$
\begin{equation*}
u(t)=F y(t)+v(t), \tag{2.18}
\end{equation*}
$$

are called closed-loop systems, see Figure 2.2. Here $F$ denotes the so called feedback operator and $v(t)$ the new input.


Figure 2.1: open-loop system $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$


Figure 2.2: closed-loop system $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ with feedback $F$
The proofs of the following results about feedback systems can be found in Wei94 and Sta05. We give a brief overview on closed-loop systems and start with the considerations which operators are admissible as a feedback operator. Then we will see that not only well-posedness is preserved under an admissible feedback operator, but also exact controllability.

Definition 2.2.23. ( $(\overline{W e i} 94, ~ P r o p o s i t i o n ~ 4.9]) ~ L e t ~ t h e ~ s y s t e m ~(~ S ~(~ A ~, ~ \mathfrak{B}, \mathfrak{C})$ be a regular boundary control and observation system (2.15) and we denote by $D \in \mathcal{L}(U, Y)$ the corresponding feedthrough operator. Assume that for the transfer function $G$ holds

$$
\begin{equation*}
\lim _{r \rightarrow \infty} \sup _{\omega \in \mathbb{R}}\|G(i \omega+r)-D\|=0 \tag{2.19}
\end{equation*}
$$

Then, an operator $F \in \mathcal{L}(Y, U)$ is called an admissible feedback operator for a regular boundary control and observation system (2.15), if $I-D F$ is invertible.

Proposition 2.2.24. ([JZ12, Theorem 13.1.12]) Let $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ be a well-posed boundary control and observation system 2.15). Assume that $F$ is an admissible feedback operator. Then the closed-loop system $\mathfrak{S}(\mathfrak{A},(\mathfrak{B}-F \mathfrak{C}), \mathfrak{C})$, i.e.,

$$
\begin{align*}
\dot{x}(t) & =\mathfrak{A} x(t), \quad x(0)=x_{0}, \\
v(t) & =(\mathfrak{B}-F \mathfrak{C}) x(t),  \tag{2.20}\\
y(t) & =\mathfrak{C} x(t), \quad t \geqslant 0 .
\end{align*}
$$

with input $v$ and output $y$ is a well-posed boundary control and boundary observation system.

Proposition 2.2.25. (|Wei94, cf. Remark 6.5]) Let $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ be a well-posed boundary control and observation system 2.15. Assume that $F$ is an admissible feedback operator. Then the closed-loop system $\mathfrak{S}(\mathfrak{A},(\mathfrak{B}-F \mathfrak{C}), \mathfrak{C})$ is exactly controllable if and only if the open-loop system $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ is exactly controllable.

## Chapter 3

## Introduction to port-Hamiltonian systems

Here, we consider a special class of partial differential equation on a one-dimensional space, which has an additional structure motivated by the structure of the port-Hamiltonian systems. Thus, we consider linear port-Hamiltonian systems on infinite-dimensional spaces, which we model as boundary control and observation systems as introduced in Chapter 2. Well-known examples in physics and other applications are the transport equation, the wave equation modelling a vibrating string or a transmission line, and beam equations modelling the Timoshenko beam. This class of systems makes use of the physical structure of the equations. A nice feature of the port-Hamiltonian setting is that it allows us to consider in particular boundary control and observation, which is important in applications, e.g. for the wave equation. Having this at hand we can also consider systems having an input and an output.

### 3.1 Class of port-Hamiltonian systems

In this first section we introduce port-Hamiltonian systems which have neither input nor output. In the following section we provide the class of portHamiltonian systems with an input to control and an output to observe these systems.

Assumption 3.1.1. Let $P_{1} \in \mathbb{C}^{d \times d}$ be an invertible Hermitian matrix, $P_{0} \in$ $\mathbb{C}^{d \times d}$ a skew-symmetric matrix, $\left[\begin{array}{ll}\widetilde{W}_{1} & \widetilde{W}_{0}\end{array}\right]$ a full row rank $d \times 2 d$ matrix, and $\mathcal{H}(\zeta)$ a positive $d \times d$ Hermitian matrix for a.e. $\zeta \in(0,1)$ satisfying $\mathcal{H}, \mathcal{H}^{-1} \in$ $L^{\infty}\left(0,1 ; \mathbb{C}^{d \times d}\right)$. Since $\mathcal{H}(\zeta)$ is positive definite and $P_{1}$ a Hermitian matrix, $P_{1} \mathcal{H}(\zeta)$ is similar to a Hermitian matrix, and thus, the matrix $P_{1} \mathcal{H}(\zeta)$ can be diagonalized as $P_{1} \mathcal{H}(\zeta)=S^{-1}(\zeta) \Delta(\zeta) S(\zeta)$, where $\Delta(\zeta)$ is a diagonal matrix and $S(\zeta)$ is an invertible matrix for almost every $\zeta \in(0,1)$. We suppose the technical assumption that $S^{-1}, S, \Delta:[0,1] \rightarrow \mathbb{C}^{d \times d}$ are continuously differentiable.

Note that, for instance, in Kat95, Chapter II] conditions for $P_{1} \mathcal{H}$ such that $S^{-1}, S, \Delta:[0,1] \rightarrow \mathbb{C}^{d \times d}$ are continuously differentiable, are described.

With this assumption in mind, we introduce that type of partial differential equation which is the main subject of this thesis.

Definition 3.1.2. Let Assumption 3.1.1 be fulfilled. A system which is given by the following partial differential equation

$$
\begin{align*}
\frac{\partial x}{\partial t}(\zeta, t) & =\left(P_{1} \frac{\partial}{\partial \zeta}+P_{0}\right)(\mathcal{H}(\zeta) x(\zeta, t)), \zeta \in(0,1), t \geqslant 0 \\
x(\zeta, 0) & =x_{0}(\zeta),  \tag{3.1}\\
0 & =\left[\begin{array}{ll}
\widetilde{W}_{1} & \widetilde{W}_{0}
\end{array}\right]\left[\begin{array}{l}
(\mathcal{H} x)(1, t) \\
(\mathcal{H} x)(0, t)
\end{array}\right], t>0,
\end{align*}
$$

is called a port-Hamiltonian system.
Remark 3.1.3. To shorten the notation, we use the term port-Hamiltonian system instead of linear, first order port-Hamiltonian system.
The energy, also denoted as Hamiltonian, of a port-Hamiltonian system can be described by

$$
\begin{equation*}
E(t)=\frac{1}{2} \int_{0}^{1} x(\zeta, t)^{*} \mathcal{H}(\zeta) x(\zeta, t) d \zeta . \tag{3.2}
\end{equation*}
$$

We choose all states with finite energy as the state space $X$, i.e., all functions $x$ such that $\frac{1}{2} \int_{0}^{1} x(\zeta, t)^{*} \mathcal{H}(\zeta) x(\zeta, t) d \zeta$ is finite. Due to the requirements on $\mathcal{H}$ in Assumption 3.1.1 these are all functions, which are square integrable over the unit interval. Thus, we set the state space $X=L^{2}\left(0,1 ; \mathbb{C}^{d}\right)$ and we equip it not with the standard $L^{2}$-norm, but with the inner product

$$
\langle f, g\rangle=\frac{1}{2} \int_{0}^{1} f(\zeta)^{*} \mathcal{H}(\zeta) g(\zeta) d \zeta, \quad f, g \in L^{2}\left(0,1 ; \mathbb{C}^{d}\right)
$$

Then the squared norm of an element $x \in X$ equals the energy of the state $x$ of the port-Hamiltonian system and therefore, the norm is called energy norm on the energy space $X$. We point out, that the energy norm and the standard $L^{2}-$ norm on $X$ are equivalent. For the energy of port-Hamiltonian systems holds the following power balance equation, which can be proved by integration by parts, cf. JZ12, Lemma 7.1.5].

Proposition 3.1.4. Let $x$ denote the classical solution of the port-Hamiltonian system (3.1). Then the balance equation

$$
\begin{equation*}
\frac{d E}{d t}(t)=\frac{1}{2}\left[((\mathcal{H} x)(\zeta, t))^{*} P_{1}(\mathcal{H} x)(\zeta, t)\right]_{0}^{1} \tag{3.3}
\end{equation*}
$$

holds.
The power balance equation (3.3) explains the name of this class of systems by taking into account that the energy can also change via the boundary of the system.

Remark 3.1.5. Without loss of generality it is possible to consider only portHamiltonian systems on the unit interval instead of port-Hamiltonian systems on an arbitrary interval $[a, b]$. In fact, there is an isometric isomorphism $\alpha$ between the corresponding state spaces:

$$
\begin{aligned}
\alpha: L^{2}\left(a, b ; \mathbb{C}^{d}\right) & \rightarrow L^{2}\left(0,1 ; \mathbb{C}^{d}\right) \\
x(\cdot) & \mapsto x\left(\frac{\cdot-a}{b-a}\right)
\end{aligned}
$$

In what follows, we give some examples to illustrate that there are many physical systems which fit in the class of port-Hamiltonian systems. These and further examples can also be found in (Vil07, [JZ12], [JMZ15], Aug16] or [JZ18], just to mention a few. Examples for a more general class of port-Hamiltonian systems can be found in Chapter 6 of this thesis as well.
Example 3.1.6. The following partial differential equation is called transport equation.

$$
\begin{align*}
\frac{\partial x}{\partial t}(\zeta, t) & =\frac{\partial}{\partial \zeta}(c(\zeta) x(\zeta, t)), x(\zeta, 0)=x_{0}(\zeta), \zeta \in(0,1), t>0  \tag{3.4}\\
0 & =(c x)(1, t)-\mu(c x)(0, t), \mu \in \mathbb{C}, t>0 \tag{3.5}
\end{align*}
$$

where $c(\zeta):[0,1] \rightarrow \mathbb{R}$ is a bounded, continuously differentiable function such that $c(\zeta)>0$ for $\zeta \in[0,1]$. This is the simplest port-Hamiltonian system with $P_{1}=1, P_{0}=0$ and a complex valued function $\mathcal{H}(\zeta)=c(\zeta)$. The boundary conditions 3.5 can reformulated in the port-Hamiltonian setting as

$$
\left[\begin{array}{ll}
1 & -\mu
\end{array}\right]\left[\begin{array}{l}
(c x)(1, t) \\
(c x)(0, t)
\end{array}\right]=0, \mu \in \mathbb{C}, t>0
$$




Figure 3.1: Sketch of the translation shift for $c \equiv 1$ and $\mu=2$
The energy of the system is described by

$$
E(t)=\frac{1}{2} \int_{0}^{1} c(\zeta)|x(\zeta, t)|^{2} d \zeta
$$

Example 3.1.7. The vibrating string can be described by the wave equation. We consider a string which is clamped at the left side and freely vibrating at the right side.

Figure 3.2: vibrating wave clamped at $\zeta=0$

$$
\begin{align*}
\frac{\partial^{2} \omega}{\partial t^{2}}(\zeta, t) & =\frac{1}{\rho(\zeta)} \frac{\partial}{\partial \zeta}\left(T(\zeta) \frac{\partial \omega}{\partial \zeta}(\zeta, t)\right), x(\zeta, 0)=x_{0}(\zeta), \zeta \in(0,1), t>0  \tag{3.6}\\
0 & =\left[\begin{array}{c}
\frac{\partial \omega}{\partial t}(0, t) \\
T(1) \frac{\partial \omega}{\partial \zeta}(1, t)
\end{array}\right], \quad t \geqslant 0 \tag{3.7}
\end{align*}
$$

where $w(\zeta, t)$ is the vertical position of the string at position $\zeta$ and time $t$, $T(\zeta)>0$ is the Young's modulus of the string, and $\rho(\zeta)>0$ is the mass density. We introduce as the new state variables

$$
\begin{array}{ll}
x_{1}(\zeta, t):=\rho(\zeta) \frac{\partial \omega}{\partial t}(\zeta, t) & \text { the momentum, and } \\
x_{2}(\zeta, t):=\frac{\partial \omega}{\partial \zeta}(\zeta, t) & \text { the strain. }
\end{array}
$$

Hence, we can model the wave equation (3.6) as

$$
\begin{align*}
\frac{\partial}{\partial t}\left[\begin{array}{l}
x_{1} \\
x_{2}
\end{array}\right](\zeta, t) & =\left[\begin{array}{ll}
0 & 1 \\
1 & 0
\end{array}\right] \frac{\partial}{\partial \zeta}\left(\left[\begin{array}{cc}
\frac{1}{\rho(\zeta)} & 0 \\
0 & T(\zeta)
\end{array}\right] x(\zeta, t)\right)  \tag{3.8}\\
0 & =\left[\begin{array}{llll}
0 & 0 & 1 & 0 \\
0 & 1 & 0 & 0
\end{array}\right]\left[\begin{array}{l}
(\mathcal{H} x)(1, t) \\
(\mathcal{H} x)(0, t)
\end{array}\right]
\end{align*}
$$

where $x=\left[\begin{array}{ll}x_{1} & x_{2}\end{array}\right]^{\top}$.
Moreover, the energy of the system can be written in the chosen state variables as

$$
E(t)=\frac{1}{2} \int_{0}^{1} \frac{\left|x_{1}(\zeta, t)\right|^{2}}{\rho(\zeta)}+T(\zeta)\left|x_{2}(\zeta, t)\right|^{2} d \zeta .
$$

Example 3.1.8. The Timoshenko beam equations model the effects in a vibrating beam and take into account shear and rotational effects. A beam, which is clamped at both sides, i.e., at $\zeta=0$ and at $\zeta=1$, can be modelled by

$$
\begin{aligned}
\rho(\zeta) \frac{\partial^{2} \omega}{\partial t^{2}}(\zeta, t) & =\frac{\partial}{\partial \zeta}\left(K(\zeta)\left(\frac{\partial \omega}{\partial \zeta}(\zeta, t)-\Phi(\zeta, t)\right)\right), \zeta \in(0,1), t \geqslant 0 \\
I_{\rho(\zeta)} \frac{\partial^{2} \Phi}{\partial t^{2}} & =\frac{\partial}{\partial \zeta}\left(E I(\zeta) \frac{\partial \Phi}{\partial \zeta}(\zeta, t)\right)+K(\zeta)\left(\frac{\partial \omega}{\partial \zeta}(\zeta, t)-\Phi(\zeta, t)\right) \\
\frac{\partial \omega}{\partial t}(0, t) & =\frac{\partial \Phi}{\partial t}(0, t)=\frac{\partial \omega}{\partial t}(1, t)=\frac{\partial \Phi}{\partial t}(1, t)=0, x(\zeta, 0)=x_{0}(\zeta)
\end{aligned}
$$

where $\omega(\zeta, t)$ denotes the transverse displacement of the beam and $\Phi(\zeta, t)$ the rotation angle of a filament of the beam. All physical parameters are positive
and continuously differentiable functions of $\zeta . K(\zeta)$ denotes the shear modulus, $E I(\zeta)$ is the product of Young's modulus of elasticity and the moment of inertia of a cross section, $\rho(\zeta)$ is the mass per unit length and $I_{\rho}(\zeta)$ denotes the rotary moment of inertia of a cross section.
In order to model the Timoshenko beam as a port-Hamiltonian system, we introduce new state variables

$$
\begin{array}{rlrl}
x_{1}(\zeta, t) & =\frac{\partial \omega}{\partial \zeta}(\zeta, t)-\Phi(\zeta, t) & & \text { the shear displacement } \\
x_{2}(\zeta, t) & =\rho(\zeta) \frac{\partial \omega}{\partial t}(\zeta, t) & & \text { the momentum } \\
x_{3}(\zeta, t)=\frac{\partial \Phi}{\partial \zeta}(\zeta, t) & & \text { the angular displacement } \\
x_{4}(\zeta, t)=I_{\rho}(\zeta) \frac{\partial \Phi}{\partial t}(\zeta, t) & & \text { the angular momentum. }
\end{array}
$$

The Timoshenko beam can be modelled as a port-Hamiltonian system using these new state variables. It can be written in the form of (3.1) with

$$
\begin{gathered}
P_{1}=\left[\begin{array}{llll}
0 & 1 & 0 & 0 \\
1 & 0 & 0 & 0 \\
0 & 0 & 0 & 1 \\
0 & 0 & 1 & 0
\end{array}\right], \mathcal{H}(\zeta)=\left[\begin{array}{cccc}
K(\zeta) & 0 & 0 & 0 \\
0 & \frac{1}{\rho(\zeta)} & 0 & 0 \\
0 & 0 & E I(\zeta) & 1 \\
0 & 0 & 0 & \frac{1}{I_{\rho}(\zeta)}
\end{array}\right], \\
P_{0}=\left[\begin{array}{cccc}
0 & 0 & 0 & -1 \\
0 & 0 & 0 & 0 \\
0 & 0 & 0 & 0 \\
1 & 0 & 0 & 0
\end{array}\right],
\end{gathered}
$$

and

$$
\left[\begin{array}{ll}
\widetilde{W}_{1} & \widetilde{W}_{0}
\end{array}\right]=\left[\begin{array}{llllllll}
0 & 1 & 0 & 0 & 0 & 0 & 0 & 0 \\
0 & 0 & 0 & 1 & 0 & 0 & 0 & 0 \\
0 & 0 & 0 & 0 & 0 & 1 & 0 & 0 \\
0 & 0 & 0 & 0 & 0 & 0 & 0 & 1
\end{array}\right] .
$$

The energy of the system is given by

$$
\begin{aligned}
E(t)=\frac{1}{2} \int_{0}^{1} K(\zeta)\left|x_{1}(\zeta, t)\right|^{2} & +\frac{1}{\rho(\zeta)}\left|x_{2}(\zeta, t)\right|^{2} \\
& +E I(\zeta)\left|x_{3}(\zeta, t)\right|^{2}+\frac{1}{I_{\rho}(\zeta)}\left|x_{4}(\zeta, t)\right|^{2} d \zeta
\end{aligned}
$$

### 3.2 Generation theorems

To study the whole class of systems instead of considering each example separately, we aim to formulate port-Hamiltonian systems as abstract Cauchy problems 2.2. Hence, we introduce the port-Hamiltonian operator associated to (3.1) and study the question which port-Hamiltonian operators generate (contractive) $C_{0}$-semigroups.

Definition 3.2.1. Let $P_{0}, P_{1}, \mathcal{H}$ satisfy Assumption 3.1.1 and define $X:=$ $L^{2}\left(0,1 ; \mathbb{C}^{d}\right)$. Then the operator $A: \mathcal{D}(A) \subset X \rightarrow X$ defined by

$$
\begin{gather*}
A x:=\left(P_{1} \frac{d}{d \zeta}+P_{0}\right)(\mathcal{H} x), \quad x \in \mathcal{D}(A),  \tag{3.9}\\
\mathcal{D}(A):=\left\{x \in X \mid \mathcal{H} x \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{d}\right) \text { and }\left[\begin{array}{ll}
\widetilde{W}_{1} & \widetilde{W}_{0}
\end{array}\right]\left[\begin{array}{l}
(\mathcal{H} x)(1) \\
(\mathcal{H} x)(0)
\end{array}\right]=0\right\} \tag{3.10}
\end{gather*}
$$

is called port-Hamiltonian operator.
For port-Hamiltonian systems the boundary conditions are often equivalently reformulated via the boundary flow and the boundary effort. We introduce them in the following.

Definition 3.2.2. For a port-Hamiltonian system we define the boundary flow $f_{\delta, \mathcal{H} x}$ and the boundary effort $e_{\delta, \mathcal{H} x}$ as

$$
\left[\begin{array}{c}
f_{\delta, \mathcal{H} x}  \tag{3.11}\\
e_{\delta, \mathcal{H} x}
\end{array}\right]:=\frac{1}{\sqrt{2}}\left[\begin{array}{cc}
P_{1} & -P_{1} \\
I & I
\end{array}\right]\left[\begin{array}{l}
(\mathcal{H} x)(1) \\
(\mathcal{H} x)(0)
\end{array}\right] .
$$

Define $\widetilde{W}_{B}=\left[\begin{array}{ll}\widetilde{W}_{1} & \widetilde{W}_{0}\end{array}\right]$. Then

$$
\widetilde{W}_{B}\left[\begin{array}{l}
(\mathcal{H} x)(1, t)  \tag{3.12}\\
(\mathcal{H} x)(0, t)
\end{array}\right]=0 \Leftrightarrow W_{B}\left[\begin{array}{l}
f_{\delta, \mathcal{H} x} \\
e_{\delta, \mathcal{H} x}
\end{array}\right]=0,
$$

where

$$
W_{B}=\widetilde{W}_{B} \sqrt{2}\left[\begin{array}{cc}
P_{1} & -P_{1}  \tag{3.13}\\
I & I
\end{array}\right]^{-1}=\widetilde{W}_{B} \frac{1}{\sqrt{2}}\left[\begin{array}{cc}
P_{1}^{-1} & I \\
-P_{1}^{-1} & I
\end{array}\right] .
$$

Thus, we get an equivalent formulation of the boundary conditions of a portHamiltonian system, which is useful for the next corollary.
The following assertion can be found in LGZM05, [JZ12, Thereom 7.2.4], and Aug16] and characterizes the generation of contraction $C_{0}$-semigroups for portHamiltonian operators.

Theorem 3.2.3. Let $A$ be a port-Hamiltonian operator given by (3.9)-(3.10). Let $\widetilde{W}_{B}$ be a matrix with full row rank. Then the following statements are equivalent.

1. $A$ is the generator of a contraction $C_{0}$-semigroup on $X$,
2. $A$ is dissipative,
3. $W_{B} \Sigma W_{B}^{*} \geqslant 0$, where $\Sigma:=\left[\begin{array}{ll}0 & I \\ I & 0\end{array}\right]$.

If one of the above conditions is fulfilled, then A has a compact resolvent. Furthermore, A generates a unitary $C_{0}$-group if and only if $W_{B} \Sigma W_{B}^{*}=0$ holds true.

### 3.2. GENERATION THEOREMS

Remark 3.2.4. Note that for symmetric matrices $M \in \mathbb{C}^{d \times d}$ we write $M \geqslant m I$ with a real constant $m$ if $\langle v, M v\rangle \geqslant m\|v\|^{2}$ for all $v \in \mathbb{C}^{d}$.
The proof of Theorem 3.2 .3 can be found in [LGZM05, Theorem 4.1] and [JZ12, Theorem 7.2.4] and is even a Corollary of Theorem 6.1.3 in Chapter 6 .
Furthermore, the generation of $C_{0}$-semigroups for port-Hamiltonian operators is characterized in JMZ15.

Theorem 3.2.5. Let $A$ be a port-Hamiltonian operator. Let $Z^{+}(\zeta)$ be the span of the eigenvectors of $P_{1} \mathcal{H}(\zeta)$ corresponding to the positive eigenvalues of $P_{1} \mathcal{H}(\zeta)$ and $Z^{-}(\zeta)$ be the span of the eigenvectors of $P_{1} \mathcal{H}(\zeta)$ corresponding to the negative eigenvalues of $P_{1} \mathcal{H}(\zeta)$. Then the following statements are equivalent:

1. $A$ is the generator of a $C_{0}$-semigroup on $X$,
2. $\widetilde{W}_{1} \mathcal{H}(1) Z^{+}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{-}(0)=\mathbb{C}^{d}$.

Using Theorem 2.1.16 we formulate the following corollary:
Corollary 3.2.6. Let $A$ be a port-Hamiltonian operator and let $Z^{+}(\zeta)$ and $Z^{-}(\zeta)$ be defined as in Theorem 3.2.5. Then the following statements are equivalent:

1. $A$ is the generator of a $C_{0}$-group on $X$,
2. $\widetilde{W}_{1} \mathcal{H}(1) Z^{+}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{-}(0)=\widetilde{W}_{1} \mathcal{H}(1) Z^{-}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{+}(0)=\mathbb{C}^{d}$.

With the knowledge of this generation theorems, we consider the Examples 3.1.6 3.1.8 again and study which of these port-Hamiltonian systems are wellposed.
Example 3.2.7. Continuation of Example 3.1.6. For the transport equation we define the associated port-Hamiltonian operator on $X=L^{2}(0,1 ; \mathbb{C})$

$$
\begin{aligned}
A x & =\frac{\partial}{\partial \zeta}(c x), \quad x \in \mathcal{D}(A), \\
\mathcal{D}(A) & =\left\{x \in X \mid c x \in \mathcal{W}^{1,2}(0,1 ; \mathbb{C}) \text { and }\left[\begin{array}{ll}
1 & -\mu
\end{array}\right]\left[\begin{array}{l}
(c x)(1) \\
(c x)(0)
\end{array}\right]=0, \mu \in \mathbb{R}\right\}
\end{aligned}
$$

Since for general port-Hamiltonian systems $\operatorname{Re}\langle A x, x\rangle$ is not easy to determine to check the dissipativity, Theorem 3.2.3 gives an equivalent easy checkable matrix condition. But for this system we can even determine $\operatorname{Re}\langle A x, x\rangle$ using integration by parts

$$
\begin{aligned}
\operatorname{Re}\langle A x, x\rangle=\operatorname{Re}\left\langle\frac{\partial}{\partial \zeta}(c x), c x\right\rangle & =\frac{1}{2}\left[\left\langle c x, \frac{\partial}{\partial \zeta}(c x)\right\rangle+\left\langle\frac{\partial}{\partial \zeta}(c x), c x\right\rangle\right] \\
& =\frac{1}{2}\left[\left.|c x|^{2}\right|_{0} ^{1}-\left\langle\frac{\partial}{\partial \zeta}(c x), c x\right\rangle+\left\langle\frac{\partial}{\partial \zeta}(c x), c x\right\rangle\right] \\
& =\frac{1}{2}\left[|(c x)(1)|^{2}-|(c x)(0)|^{2}\right] .
\end{aligned}
$$

The boundary conditions implies $|(c x)(1)|^{2}=\mu^{2}|(c x)(0)|^{2}$ and so

$$
\operatorname{Re}\langle A x, x\rangle=\frac{1}{2}\left[|(c x)(1)|^{2}-|(c x)(0)|^{2}\right]=\frac{1}{2}\left(\mu^{2}-1\right)|(c x)(0)|^{2} .
$$

Thus, it holds $\operatorname{Re}\langle A x, x\rangle \leqslant 0$ if and only if $|\mu| \leqslant 1$ and hence $A$ generates a contraction $C_{0}$-semigroup if and only if $|\mu| \leqslant 1$.
Nevertheless, in this example, the matrix condition in Theorem 3.2.3 also answers the question whether $A$ generates a contraction $C_{0}$-semigroup much faster: $\widetilde{W}_{B}=\left[\begin{array}{ll}1 & -\mu\end{array}\right]$ implies $W_{B}=\frac{1}{\sqrt{2}}\left[\begin{array}{ll}1+\mu & 1-\mu\end{array}\right]$ and thus $W_{B} \Sigma W_{B}^{*}=1-|\mu| \geqslant$ 0 if and only if $|\mu| \leqslant 1$. Furthermore, we can use Theorem 3.2 .5 to study whether $A$ generates a $C_{0}$-semigroup for $|\mu|>1$. Since $P_{1} \mathcal{H}(\zeta)=\mathcal{H}(\zeta)=c(\zeta)$, it holds $Z^{+}(1)=\mathbb{C}$ and $Z^{-}(0)=\{0\}$. Thus, $A$ generates a $C_{0}$-semigroup even for $|\mu|>1$.
Example 3.2.8. Continuation of Example 3.1.7. Again, we consider the wave equation which is in Example 3.1.7 written as the port-Hamiltonian system (3.8) and we define the associated port-Hamiltonian operator for $x=\left[\begin{array}{ll}x_{1} & x_{2}\end{array}\right]^{\top}$ as

$$
\begin{align*}
A x & =\left[\begin{array}{ll}
0 & 1 \\
1 & 0
\end{array}\right] \frac{\partial}{\partial \zeta}\left(\left[\begin{array}{cc}
\frac{1}{\rho(\zeta)} & 0 \\
0 & T(\zeta)
\end{array}\right] x\right), \quad x \in \mathcal{D}(A),  \tag{3.14}\\
\mathcal{D}(A) & =\left\{x \in X \mid \mathcal{H} x \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{2}\right) \text { and } \widetilde{W}_{B}\left[\begin{array}{l}
(\mathcal{H} x)(1) \\
(\mathcal{H} x)(0)
\end{array}\right]=0\right\},
\end{align*}
$$

where $\mathcal{H}=\left[\begin{array}{cc}\frac{1}{\rho(\zeta)} & 0 \\ 0 & T(\zeta)\end{array}\right]$ and $\widetilde{W}_{B}=\left[\begin{array}{llll}0 & 0 & 1 & 0 \\ 0 & 1 & 0 & 0\end{array}\right]$ describes the boundary conditions. These boundary conditions of $A$ model the situation where the string is clamped at the left side and free vibrating at the right side. Then it holds due to (3.13)

$$
W_{B}=\frac{1}{2}\left[\begin{array}{cccc}
0 & -1 & 1 & 0  \tag{3.15}\\
1 & 0 & 0 & 1
\end{array}\right]
$$

and thus, $W_{B} \Sigma W_{B}^{*}=0$. Hence, $A$ generates a contraction $C_{0}$-semigroup and moreover a unitary $C_{0}$-group. To illustrate that a port-Hamiltonian operator does not always generate a unitary $C_{0}$-group, we consider in a slightly modified setting the port-Hamiltonian operator (3.14) with boundary conditions described by

$$
\widetilde{W}_{B}=\left[\begin{array}{llll}
0 & 0 & \kappa & 1 \\
0 & 1 & 0 & 0
\end{array}\right], \quad \kappa>0,
$$

which model a vibrating string with an amplifier at the left end and free at the right end. Then it holds

$$
W_{B}=\frac{1}{\sqrt{2}}\left[\begin{array}{cccc}
-1 & -\kappa & \kappa & 1 \\
1 & 0 & 0 & 1
\end{array}\right]
$$

and $W_{B} \Sigma W_{B}^{*}=\left[\begin{array}{cc}-2 \kappa & 0 \\ 0 & 0\end{array}\right]$, which is not a positive semi-definite matrix. Thus, due to Theorem 3.2.3 the wave equation equipped with these boundary conditions does not generate a contraction $C_{0}$-semigroup.

In the following, we study the question, whether the operator generates at least a $C_{0}$-semigroup. Theorem 3.2 .5 gives a helpful characterization for $C_{0}{ }^{-}$ semigroup generation of port-Hamiltonian operators. This part of the example can also be found in JMZ15]. Defining $\gamma=\sqrt{T(\zeta) / \rho(\zeta)}$, the matrix function $P_{1} \mathcal{H}$ can be factorized as

$$
P_{1} \mathcal{H}=\underbrace{\left[\begin{array}{cc}
\gamma & -\gamma  \tag{3.16}\\
\rho^{-1} & \rho^{-1}
\end{array}\right]}_{S^{-1}} \underbrace{\left[\begin{array}{cc}
\gamma & 0 \\
0 & -\gamma
\end{array}\right]}_{\Delta} \underbrace{\left[\begin{array}{l}
(2 \gamma)^{-1} \\
(2 \gamma)^{-1}
\end{array} \rho / 2\right.}_{S} \begin{array}{l}
\rho / 2
\end{array}] .
$$

Then, $P_{1} \mathcal{H}$ has eigenvalues $\gamma$ and $-\gamma$ with corresponding eigenvectors $\left[\begin{array}{l}T(\zeta) \\ \gamma(\zeta)\end{array}\right]$, and $\left[\begin{array}{c}-T(\zeta) \\ \gamma(\zeta)\end{array}\right]$, respectively. Since the eigenspaces are one-dimensional $Z^{+}(\zeta)$ and $Z^{-}(\zeta)$ are each the span of the corresponding single eigenvector, it equals

$$
\begin{aligned}
\widetilde{W}_{1} \mathcal{H}(1) Z^{+}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{-}(0) & =\left[\begin{array}{ll}
0 & 0 \\
0 & 1
\end{array}\right]\left[\begin{array}{l}
\gamma(1) \\
T(1)
\end{array}\right] \oplus\left[\begin{array}{ll}
\kappa & 1 \\
0 & 0
\end{array}\right]\left[\begin{array}{c}
-\gamma(0) \\
T(0)
\end{array}\right] \\
& =\left[\begin{array}{c}
0 \\
T(1)
\end{array}\right] \oplus\left[\begin{array}{c}
-\kappa \gamma(0)+T(0) \\
0
\end{array}\right]=\mathbb{C}^{2} .
\end{aligned}
$$

Thus, the port-Hamiltonian system (3.8) is well-posed.
Example 3.2.9. Continuation of example 3.1.8. The port-Hamiltonian operator associated to the port-Hamiltonian system of the Timoshenko beam is given by

$$
\begin{aligned}
& A x=\left(\left[\begin{array}{llll}
0 & 1 & 0 & 0 \\
1 & 0 & 0 & 0 \\
0 & 0 & 0 & 1 \\
0 & 0 & 1 & 0
\end{array}\right] \frac{d}{d \zeta}+\left[\begin{array}{cccc}
0 & 0 & 0 & -1 \\
0 & 0 & 0 & 0 \\
0 & 0 & 0 & 0 \\
1 & 0 & 0 & 0
\end{array}\right]\right)\left(\left[\begin{array}{cccc}
K(\zeta) & 0 & 0 & 0 \\
0 & \frac{1}{\rho(\zeta)} & 0 & 0 \\
0 & 0 & E I(\zeta) & 1 \\
0 & 0 & 0 & \frac{1}{I \rho(\zeta)}
\end{array}\right] x\right), \quad x \in \mathcal{D}(A), \\
& \mathcal{D}(A)=\left\{x \in X \mid \mathcal{H} x \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{d}\right) \text { and }\left[\begin{array}{ccccccc}
0 & 1 & 0 & 0 & 0 & 0 & 0 \\
0 & 0 & 0 & 1 & 0 & 0 & 0 \\
0 & 0 & 0 & 0 & 1 & 0 \\
0 & 0 & 0 & 0 & 0 & 0 & 0
\end{array}\right]\left[\begin{array}{l}
(\mathcal{H} x)(1, t) \\
(\mathcal{H} x)(0, t)
\end{array}\right]=0\right\},
\end{aligned}
$$

where $x=\left[\begin{array}{llll}x_{1} & x_{2} & x_{3} & x_{4}\end{array}\right]^{\top}$. Then it holds $W_{B}=\left[\begin{array}{ccccccc}0 & 1 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0\end{array}\right]$ and moreover, $W_{B} \Sigma W_{B}^{*}=0$. Thus, by Theorem $3.2 .3 A$ generates a unitary $C_{0}$-group.

### 3.3 Boundary control and observation port-Hamiltonian systems

Since most of the systems in applications are connected with their environment, we introduce port-Hamiltonian systems with inputs and outputs. In a first step, we add only an input to the port-Hamiltonian system. Thus, we consider infinite-dimensional linear port-Hamiltonian systems on a one-dimensional spatial domain with boundary control of the form

$$
\begin{align*}
\frac{\partial x}{\partial t}(\zeta, t) & =\left(P_{1} \frac{\partial}{\partial \zeta}+P_{0}\right)(\mathcal{H}(\zeta) x(\zeta, t)), \\
x(\zeta, 0) & =x_{0}(\zeta), \zeta \in(0,1),  \tag{3.17}\\
u(t) & =\widetilde{W}_{B}\left[\begin{array}{l}
(\mathcal{H} x)(1, t) \\
(\mathcal{H} x)(0, t)
\end{array}\right], t \geqslant 0 .
\end{align*}
$$

Again, Assumption 3.1.1 has to be fulfilled. These systems are called boundary control port-Hamiltonian systems. To formulate port-Hamiltonian systems with input as boundary control systems, we introduce a port-Hamiltonian operator without boundary conditions. The following is extracted from Chapters 11 and 13 in JZ12.

Definition 3.3.1. The operator

$$
\begin{equation*}
\mathfrak{A} x:=\left(P_{1} \frac{d}{d \zeta}+P_{0}\right)(\mathcal{H} x), \quad x \in \mathcal{D}(\mathfrak{A}), \tag{3.18}
\end{equation*}
$$

on $X:=L^{2}\left(0,1 ; \mathbb{C}^{d}\right)$ with the domain

$$
\begin{equation*}
\mathcal{D}(\mathfrak{A}):=\left\{x \in X \mid \mathcal{H} x \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{d}\right)\right\} \tag{3.19}
\end{equation*}
$$

is called the (maximal) port-Hamiltonian operator.
Furthermore, we introduce the input operator $\mathfrak{B}: \mathcal{D}(\mathfrak{A}) \rightarrow \mathbb{C}^{d}$ by

$$
\mathfrak{B} x=\widetilde{W}_{B}\left[\begin{array}{l}
(\mathcal{H} x)(1, t)  \tag{3.20}\\
(\mathcal{H} x)(0, t)
\end{array}\right] .
$$

Then the partial differential equation (3.17) can be written as a boundary control system

$$
\begin{aligned}
\dot{x}(t) & =\mathfrak{A} x(t), \quad x(0)=x_{0}, \\
u(t) & =\mathfrak{B} x(t) .
\end{aligned}
$$

The first important question is whether the port-Hamiltonian system (3.17) is well-posed in the sense that for every initial condition $x_{0} \in X$ and every $u \in L^{2}\left(0, t ; \mathbb{C}^{d}\right)$ equation (3.17) has a unique mild solution, cf. Definition 2.2.15. Moreover, we see that the port-Hamiltonian operator $A$ associated to the maximal port-Hamiltonian operator with

$$
\begin{align*}
A x & =\mathfrak{A} x, x \in \mathcal{D}(A),  \tag{3.21}\\
\mathcal{D}(A) & :=\left\{x \in \mathcal{D}(\mathfrak{A}) \left\lvert\, \widetilde{W}_{B}\left[\begin{array}{l}
(\mathcal{H} x)(1) \\
(\mathcal{H} x)(0)
\end{array}\right]=0\right.\right\} \tag{3.22}
\end{align*}
$$

equals the port-Hamiltonian operator defined in Definition 3.2.1. The following results gives a characterization for well-posedness of port-Hamiltonian systems.

Theorem 3.3.2. (|Vil07, ZLMV10, JZ12|) The port-Hamiltonian system (3.17) is well-posed if and only if the port-Hamiltonian operator A generates a strongly continuous $C_{0}$-semigroup on $X$.

We recall, that $A$ generates a contraction $C_{0}$-semigroup on $X$ if and only if $A$ is dissipative on $X$, cf. Theorem 3.2.3. There can be found matrix conditions to guarantee generation of a contraction $C_{0}$-semigroup, too. Matrix conditions for the generation of strongly continuous semigroups are given in Theorem 3.2.5. Now we understand well-posedness for boundary control systems and add as next step an output to these systems.

Definition 3.3.3. Systems of the form

$$
\begin{align*}
\frac{\partial x}{\partial t}(\zeta, t) & =\left(P_{1} \frac{\partial}{\partial \zeta}+P_{0}\right)(\mathcal{H}(\zeta) x(\zeta, t)), \\
x(\zeta, 0) & =x_{0}(\zeta), \zeta \in(0,1),  \tag{3.23}\\
u(t) & =\widetilde{W}_{B}\left[\begin{array}{l}
(\mathcal{H} x)(1, t) \\
(\mathcal{H} x)(0, t)
\end{array}\right] \\
y(t) & =\widetilde{W}_{C}\left[\begin{array}{l}
(\mathcal{H} x)(1, t) \\
(\mathcal{H} x)(0, t)
\end{array}\right], t \geqslant 0
\end{align*}
$$

satisfying Assumption 3.1 .1 and where $\widetilde{W}_{C}$ is a full row rank $k \times 2 d$-matrix, $k \in\{0, \cdots, d\}$, such that the matrix $\left[\begin{array}{c}\widetilde{W}_{B} \\ \widetilde{W}_{C}\end{array}\right]$ has full row rank are called boundary control and observation port-Hamiltonian system.

The case $k=0$ refers to a system without observation, that is, every definition or statement of the port-Hamiltonian system (3.23) also applies to the boundary control port-Hamiltonian system (3.17).
We define $\mathfrak{C}: \mathcal{D}(\mathfrak{A}) \rightarrow \mathbb{C}^{k}$ by

$$
\mathfrak{C} x=\widetilde{W}_{C}\left[\begin{array}{l}
(\mathcal{H} x)(1, t)  \tag{3.24}\\
(\mathcal{H} x)(0, t)
\end{array}\right]
$$

Then we can write the port-Hamiltonian system (3.23) in the following form

$$
\begin{align*}
\dot{x}(t) & =\mathfrak{A} x(t), \quad x(0)=x_{0}, \\
u(t) & =\mathfrak{B} x(t),  \tag{3.25}\\
y(t) & =\mathfrak{C} x(t) .
\end{align*}
$$

We recall, that if $A$, defined by (3.9)-(3.10), generates a strongly continuous semigroup on the state space $X$, then the port-Hamiltonian system (3.23) is a boundary control and observation system.
We note that for $x_{0} \in \mathcal{D}(\mathfrak{A})$ and $u \in C^{2}\left(0, t ; \mathbb{C}^{d}\right), t>0$, satisfying $\mathfrak{B} x_{0}=u(0)$, the boundary control and observation port-Hamiltonian system $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ possesses a unique classical solution, cf. Lemma 2.2.14.
For technical reasons we formulate the boundary conditions of (3.23) equivalently via the boundary flow and the boundary effort denoted by $f_{\delta, \mathcal{H} x}$ and $e_{\delta, \mathcal{H} x}$. Using Definition 3.2 .2 we can write the port-Hamiltonian system (3.23) equivalently as

$$
\begin{align*}
\frac{\partial x}{\partial t}(\zeta, t) & =\left(P_{1} \frac{\partial}{\partial \zeta}+P_{0}\right)(\mathcal{H}(\zeta) x(\zeta, t)) \\
x(\zeta, 0) & =x_{0}(\zeta), \zeta \in(0,1)  \tag{3.26}\\
u(t) & =W_{B}\left[\begin{array}{c}
f_{\delta, \mathcal{H}} \\
e_{\delta, \mathcal{H} x}
\end{array}\right], \\
y(t) & =W_{C}\left[\begin{array}{c}
f_{\delta, \mathcal{H} x} \\
e_{\delta, \mathcal{H} x}
\end{array}\right], t \geqslant 0
\end{align*}
$$

where analogously to $W_{B}$ in (3.13) the matrix $W_{C}$ is defined as

$$
W_{C}=\widetilde{W}_{C} \frac{1}{\sqrt{2}}\left[\begin{array}{cc}
P_{1}^{-1} & I  \tag{3.27}\\
-P_{1}^{-1} & I
\end{array}\right] .
$$

The port-Hamiltonian system (3.23) is uniquely described by $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ given by (3.21), (3.22), (3.20) and (3.24). In the following we give an example for a boundary control and observation port-Hamiltonian system.
Example 3.3.4. Continuation of Example 3.1.7 and 3.2.8. We consider the wave equation with an input and an output, namely

$$
\begin{align*}
\frac{\partial x}{\partial t}(\zeta, t) & =\left[\begin{array}{ll}
0 & 1 \\
1 & 0
\end{array}\right] \frac{\partial}{\partial \zeta}\left[\begin{array}{cc}
\frac{1}{\rho(\zeta)} & 0 \\
0 & T(\zeta)
\end{array}\right] x(\zeta, t), x(\zeta, 0)=x_{0}(\zeta), \zeta \in(0,1), \\
u(t) & =\left[\begin{array}{llll}
0 & 0 & 2 & 0 \\
0 & 0 & 0 & 2
\end{array}\right]\left[\begin{array}{c}
f_{\delta, \mathcal{H} x} \\
e_{\delta, \mathcal{H} x}
\end{array}\right],  \tag{3.28}\\
y(t) & =\left[\begin{array}{llll}
\frac{1}{2} & 0 & 0 & 0 \\
0 & \frac{1}{2} & 0 & 0
\end{array}\right]\left[\begin{array}{c}
f_{\delta, \mathcal{H} x} \\
e_{\delta, \mathcal{H} x}
\end{array}\right], t \geqslant 0 .
\end{align*}
$$

Well-posedness is a fundamental property of boundary control and observation systems. In general it is not easy to show that a boundary control and observation system is well-posed, for the port-Hamiltonian system (3.23) well-posedness is already satisfied if $A$ generates a $C_{0}$-semigroup, cf. [ZLMV10, Theorem 3.3] and (JZ12, Theorem 13.2.2].

Theorem 3.3.5. The port-Hamiltonian system (3.23) is well-posed if and only if the operator $A$ defined by (3.9)-(3.10) generates a strongly continuous semigroup on $X$.

There is a special class of port-Hamiltonian systems for which well-posedness follows immediately.

Definition 3.3.6. A port-Hamiltonian systems (3.23) is called impedance passive if

$$
\begin{equation*}
\operatorname{Re}\langle\mathfrak{A} x, x\rangle \leqslant \operatorname{Re}\langle\mathfrak{B} x, \mathfrak{C} x\rangle \tag{3.29}
\end{equation*}
$$

for every $x \in \mathcal{D}(\mathfrak{A})$. If we have equality in $(3.29)$, then the port-Hamiltonian system is called impedance energy preserving.

The fact that a port-Hamiltonian system is impedance energy preserving can be characterized by an easy checkable matrix condition.

Theorem 3.3.7. ([LGZM05, Theorem 4.4]) The port-Hamiltonian system described in (3.23) is impedance energy preserving if and only if it holds

$$
\left[\begin{array}{ll}
W_{B} \Sigma W_{B}^{*} & W_{B} \Sigma W_{C}^{*}  \tag{3.30}\\
W_{C} \Sigma W_{B}^{*} & W_{C} \Sigma W_{C}^{*}
\end{array}\right]=\left[\begin{array}{cc}
0 & I \\
I & 0
\end{array}\right],
$$

where $\Sigma=\left[\begin{array}{ll}0 & I \\ I & 0\end{array}\right]$.

Remark 3.3.8. Every impedance energy preserving port-Hamiltonian system (3.23) is well-posed; $W_{B} \Sigma W_{B}^{*}=0$ even implies that $A$ generates a unitary strongly continuous group, cf. [JMZ15, Theorem 1.1].
Example 3.3.9. Continuation of Example 3.1.7, 3.2.8, and 3.3.4. The system (3.28) is impedance energy preserving, since it holds $W_{B} \Sigma W_{B}^{*}=0, W_{C} \Sigma W_{C}^{*}=$ 0 and $W_{B} \Sigma W_{C}^{*}=W_{C} \Sigma W_{B}^{*}=I$.
Using the following balance equation we get another property of impedance passive port-Hamiltonian systems.
Lemma 3.3.10. ( $[$ JZ12, Theorem 11.3.5]) Consider the boundary control and observation port-Hamiltonian system (3.26) such that the associated port-Hamiltonian operator $A$ generates a $C_{0}$-semigroup. If the number of outputs $k=d$, then the following balance equation holds:

$$
\frac{d}{d t}\|x(t)\|^{2}=\left[\begin{array}{ll}
u^{*}(t) & y^{*}(t)
\end{array}\right]\left[\begin{array}{ll}
W_{B} \Sigma W_{B}^{*} & W_{B} \Sigma W_{C}^{*}  \tag{3.31}\\
W_{C} \Sigma W_{B}^{*} & W_{C} \Sigma W_{C}^{*}
\end{array}\right]^{-1}\left[\begin{array}{l}
u(t) \\
y(t)
\end{array}\right]
$$

Remark 3.3.11. For an impedance energy preserving port-Hamiltonian system the balance equation (3.31) becomes

$$
\frac{d}{d t} E(t)=\frac{d}{d t}\|x(t)\|^{2}=\left[\begin{array}{ll}
u(t)^{*} & y(t)^{*}
\end{array}\right]\left[\begin{array}{ll}
0 & I  \tag{3.32}\\
I & 0
\end{array}\right]\left[\begin{array}{l}
u(t) \\
y(t)
\end{array}\right]=2 \operatorname{Re}\langle u(t), y(t)\rangle .
$$

Thus, it is easy to see that for impedance energy preserving systems with input $u(t)=0$ there is no change of energy.
Well-posedness implies the existence of $\widetilde{B} \in \mathcal{L}\left(\mathbb{C}^{d}, X\right)$ with $\operatorname{ran} \widetilde{B} \subset \mathcal{D}(\mathfrak{A})$ and $\mathfrak{A} \widetilde{B} \in \mathcal{L}\left(\mathbb{C}^{d}, X\right)$. Applying Lemma 2.2.18 we get the mild solution.
Lemma 3.3.12. The unique mild solution of (3.23) with an initial value $x_{0} \in$ $L^{2}\left(0,1 ; \mathbb{C}^{d}\right)$ and $u \in L^{2}(0, t ; U)$ is given by

$$
x(t)=T(t) x_{0}+\int_{0}^{t} T_{-1}(t-s)\left(\mathfrak{A} \widetilde{B}-A_{-1} \widetilde{B}\right) u(s) d s
$$

Here the operator $\widetilde{B}: \mathbb{C}^{d} \rightarrow L^{2}\left(0,1 ; \mathbb{C}^{d}\right)$ can be defined as

$$
(\widetilde{B} u)(\zeta):=(\mathcal{H}(\zeta))^{-1}\left(S_{1} \zeta+S_{2}(1-\zeta)\right) u
$$

where $S_{1}$ and $S_{2}$ are $d \times d$-matrices given by

$$
\left[\begin{array}{l}
S_{1} \\
S_{2}
\end{array}\right]:=\left[\begin{array}{cc}
P_{1} & -P_{1} \\
I & I
\end{array}\right]^{-1} \widetilde{W}_{B}^{*}\left(\widetilde{W}_{B} \widetilde{W}_{B}^{*}\right)^{-1} .
$$

Then the port-Hamiltonian control system can be written equivalently in the standard control operator formulation (2.10)

$$
\dot{x}(t)=A_{-1} x(t)+B u(t), x(0)=x_{0}, t \geqslant 0,
$$

where $B \in \mathcal{L}\left(\mathbb{C}^{d}, X_{-1}\right)$ is given by

$$
\begin{equation*}
B:=\mathfrak{A} \widetilde{B}-A_{-1} \widetilde{B} . \tag{3.33}
\end{equation*}
$$

We recall that a boundary control and observation system with transfer function $G$ is regular, if $\lim _{s \in \mathbb{R}, s \rightarrow \infty} G(s)$ exists. Regularity of this system ensures, among other things, that the feedthrough operator $D \in \mathcal{L}(U, Y)$ exists and can be described via the transfer function, cf. Definition 2.2.21.

Lemma 3.3.13. ([JZ12, Lemma 13.2.2]) Under the standing assumptions every well-posed port-Hamiltonian system (3.23) is regular and it holds

$$
\begin{equation*}
\lim _{\operatorname{Re} s \rightarrow \infty} G(s)=\lim _{s \rightarrow \infty, s \in \mathbb{R}} G(s) . \tag{3.34}
\end{equation*}
$$

Therefore, the conditions in Definition 2.2 .23 are fulfilled and we define for regular port-Hamiltonian operators admissibility of feedback operators and recall the properties of the closed-loop system.

Definition 3.3.14. A $d \times d$-matrix $F$ is called an admissible feedback operator for a regular port-Hamiltonian system (3.23) with feedthrough operator $D$, if $I-D F$ is invertible.

Proposition 3.3.15. ([JZ12, Theorem 13.1.12]) Let $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ be a well-posed port-Hamiltonian system (3.23). Assume that $F$ is an admissible feedback operator. Then the closed-loop system $\mathfrak{S}(\mathfrak{A},(\mathfrak{B}-F \mathfrak{C}), \mathfrak{C})$, i.e.,

$$
\begin{align*}
\frac{\partial x}{\partial t}(\zeta, t) & =\left(P_{1} \frac{\partial}{\partial \zeta}+P_{0}\right)(\mathcal{H}(\zeta) x(\zeta, t)), \\
x(\zeta, 0) & =x_{0}(\zeta), \zeta \in(0,1),  \tag{3.35}\\
v(t) & =(\mathfrak{B}-F \mathfrak{C}) x(t), \\
y(t) & =\mathfrak{C} x(t), t \geqslant 0,
\end{align*}
$$

with input $v$ and output $y$ is a well-posed port-Hamiltonian system.
We recall that the open-loop system is exactly controllable, if and only if the closed-loop system is exactly controllable. In Chapter 4 we will see that wellposed port-Hamiltonian control systems are always exactly controllable. We close this chapter with a section about the spectrum of port-Hamiltonian systems.

### 3.4 Spectrum of port-Hamiltonian systems with $P_{1} \mathcal{H}(\zeta)$ diagonal

Since a well-posed port-Hamiltonian operator $A$ has compact resolvent, the spectrum of $A$ consists of isolated eigenvalues only and every point in the spectrum is an eigenvalue which has finite algebraic as well as finite geometric multiplicity, cf. [GGK90, Theorem XV.2.3].
For arbitrary generators of $C_{0}$-semigroups it is well-known that the spectrum lies in a left half-plane, cf. Proposition 2.1.12. Having the generation theorems for $C_{0}$-groups in mind, see EN00, page 79], this implies that the spectrum of operators generating a $C_{0}$-group lies in a strip parallel to the imaginary axis.

Next, for port-Hamiltonian operators with $P_{1} \mathcal{H}(\zeta)$ diagonal and $P_{0}=0$ we prove the same results.
Since the eigenvalues of $P_{1} \mathcal{H}(\zeta)$ are the same as the eigenvalues of $\mathcal{H}(\zeta)^{\frac{1}{2}} P_{1} \mathcal{H}(\zeta)^{\frac{1}{2}}$ it follows by Sylvester's law of inertia that the number of positive and negative eigenvalues of $P_{1} \mathcal{H}(\zeta)$ equal those of $P_{1}$. Let $d_{1}$ denote the number of positive and $d_{2}=d-d_{1}$ the number of negative eigenvalues of $P_{1}$. If $P_{1} \mathcal{H}(\zeta)$ is diagonal, the matrix $P_{1} \mathcal{H}(\zeta)$ can be written as $P_{1} \mathcal{H}(\zeta)=\left[\begin{array}{cc}\Lambda(\zeta) & 0 \\ 0 & \Theta(\zeta)\end{array}\right]$ without loss of generality, where $\Lambda(\zeta)=\operatorname{diag}\left(\lambda_{i}(\zeta)\right) \in \mathbb{C}^{d_{1} \times d_{1}}$ corresponds to the positive eigenvalues and $\Theta(\zeta)=\operatorname{diag}\left(\theta_{i}(\zeta)\right) \in \mathbb{C}^{d_{2} \times d_{2}}$ to the negative ones.
We split the variable $x(\zeta)=\left[\begin{array}{l}x^{+}(\zeta) \\ x^{-}(\zeta)\end{array}\right] \in \mathbb{C}^{d}$ with $x^{+}(\zeta) \in \mathbb{C}^{d_{1}}$ and $x^{-}(\zeta) \in \mathbb{C}^{d_{2}}$. Then we formulate the following proposition, which is part of JZ12, Theorem 13.3.1].

Proposition 3.4.1. Let $A_{K}$ be defined as

$$
\begin{aligned}
A_{K}\left[\begin{array}{l}
x^{+}(\zeta) \\
x^{-}(\zeta)
\end{array}\right] & =\frac{d}{d \zeta}\left(\left[\begin{array}{cc}
\Lambda(\zeta) & 0 \\
0 & \Theta(\zeta)
\end{array}\right]\left[\begin{array}{l}
x^{+}(\zeta) \\
x^{-}(\zeta)
\end{array}\right]\right) \\
\mathcal{D}\left(A_{K}\right) & =\left\{\left[\begin{array}{l}
x^{+}(\zeta) \\
x^{-}(\zeta)
\end{array}\right] \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{d}\right) \left\lvert\, K\left[\begin{array}{l}
\Lambda(1) x^{+}(1) \\
\Theta(0) x^{-}(0)
\end{array}\right]+Q\left[\begin{array}{l}
\Lambda(0) x^{+}(0) \\
\Theta(1) x^{-}(1)
\end{array}\right]\right.\right\}
\end{aligned}
$$

where $K, Q \in \mathbb{C}^{d \times d}$ such that $\left[\begin{array}{ll}K & Q\end{array}\right]$ has full row rank. Then it holds: $A_{K}$ generates a $C_{0}$-semigroup on $X=L^{2}\left(0,1 ; \mathbb{C}^{d}\right)$ if and only if $K$ is invertible.

Proposition 3.4.2. Let $A$ be a port-Hamiltonian operator (3.9)-(3.10) with $P_{1} \mathcal{H}(\zeta)$ diagonal and $P_{0}=0$. If $A$ generates a $C_{0}$-semigroup, then its eigenvalues lie in a strip parallel to the imaginary axis.

Proof: In the first part of the proof we give a condition under which $s \in \mathbb{C}$ is an eigenvalue. Then, we use this condition to show that there are no eigenvalues in a certain left half-plane. We start with the characterization of the eigenvalues of $A$.
Let $s \in \mathbb{C}$ be arbitrarily. Then $z$ is a solution of $s z=A z$ if and only if

$$
s z(\zeta)=\frac{d}{d \zeta}\left[\begin{array}{cc}
\Lambda(\zeta) & 0  \tag{3.36}\\
0 & \Theta(\zeta)
\end{array}\right] z(\zeta), \quad \zeta \in[0,1] .
$$

The solution of (3.36) is given by

$$
z_{i}(\zeta)= \begin{cases}\frac{c_{i}}{\lambda_{i}(\zeta)} \mathrm{e}^{s \int_{0}^{\zeta} \frac{1}{\lambda_{i}(y)} d y} & \text { for } 1 \leqslant i \leqslant d_{1} \\ \frac{c_{i}}{\theta_{i}(\zeta)} \mathrm{e}^{s \int_{0}^{\zeta} \frac{1}{\theta_{i}(y)} d y} & \text { for } d_{1}+1 \leqslant i \leqslant d\end{cases}
$$

The number $s$ is an eigenvalue of $A$ if and only if there exist constants $c_{i}$ such that $z \in \mathcal{D}(A)$. We split the variable $z(\zeta)=\left[\begin{array}{c}z^{+}(\zeta) \\ z^{-}(\zeta)\end{array}\right] \in \mathbb{C}^{d}$ with $z^{+}(\zeta) \in \mathbb{C}^{d_{1}}$ and $z^{-}(\zeta) \in \mathbb{C}^{d_{2}}$, and we define $\widetilde{W}_{1} \mathcal{H}(1)=:\left[\begin{array}{ll}V_{1} & V_{2}\end{array}\right]$ and $\widetilde{W}_{0} \mathcal{H}(0)=:\left[\begin{array}{ll}U_{1} & U_{2}\end{array}\right]$,
where $U_{1}, V_{1} \in \mathbb{C}^{d \times d_{1}}$ and $U_{2}, V_{2} \in \mathbb{C}^{d \times d_{2}}$. Then, $z \in \mathcal{D}(A)$ if and only if

$$
\begin{aligned}
0 & =\left[\begin{array}{ll}
\widetilde{W}_{1} & \widetilde{W}_{0}
\end{array}\right]\left[\begin{array}{l}
(\mathcal{H} z)(1) \\
(\mathcal{H} z)(0)
\end{array}\right] \\
& =\left[\begin{array}{ll}
V_{1} & V_{2}
\end{array}\right]\left[\begin{array}{l}
z^{+}(1) \\
z^{-}(1)
\end{array}\right]+\left[\begin{array}{ll}
U_{1} & U_{2}
\end{array}\right]\left[\begin{array}{l}
z^{+}(0) \\
z^{-}(0)
\end{array}\right] \\
& =\left[\begin{array}{ll}
V_{1} & U_{2}
\end{array}\right]\left[\begin{array}{l}
z^{+}(1) \\
z^{-}(0)
\end{array}\right]+\left[\begin{array}{ll}
U_{1} & V_{2}
\end{array}\right]\left[\begin{array}{l}
z^{+}(0) \\
z^{-}(1)
\end{array}\right] \\
& =K\left[\begin{array}{l}
\Lambda(1) z^{+}(1) \\
\Theta(0) z^{-}(0)
\end{array}\right]+Q\left[\begin{array}{l}
\Lambda(0) z^{+}(0) \\
\Theta(1) z^{-}(1)
\end{array}\right],
\end{aligned}
$$

where $K:=\left[\begin{array}{ll}V_{1} & U_{2}\end{array}\right]\left[\begin{array}{cc}\Lambda(1)^{-1} & 0 \\ 0 & \Theta(0)^{-1}\end{array}\right]$ and $Q:=\left[\begin{array}{ll}U_{1} & V_{2}\end{array}\right]\left[\begin{array}{cc}\Lambda(0)^{-1} & 0 \\ 0 & \Theta(1)^{-1}\end{array}\right]$. Since $A$ is the generator of a $C_{0}$-semigroup, $K$ is invertible, see Proposition 3.4.1. Thus, $s$ is an eigenvalue of $A$ if and only if

$$
\begin{aligned}
0 & =\left[\begin{array}{l}
\Lambda(1) z^{+}(1) \\
\Theta(0) z^{-}(0)
\end{array}\right]+K^{-1} Q\left[\begin{array}{l}
\Lambda(0) z^{+}(0) \\
\Theta(1) z^{-}(1)
\end{array}\right] \\
& =\left(I+K^{-1} Q G(s)\right)\left[\begin{array}{l}
\Lambda(1) z^{+}(1) \\
\Theta(0) z^{-}(0)
\end{array}\right],
\end{aligned}
$$

where it is easy to verify that $G(s)=\operatorname{diag}\left(g_{i}(s)\right)$ with

$$
g_{i}(s)= \begin{cases}\mathrm{e}^{-s \int_{0}^{1} \frac{1}{\lambda_{i}(y)} d y} & \text { for } 1 \leqslant i \leqslant d_{1}  \tag{3.37}\\ \mathrm{e}^{s \int_{0}^{1} \frac{1}{\theta_{i}(y)} d y} & \text { for } d_{1}+1 \leqslant i \leqslant d\end{cases}
$$

Summarising, $s \in \mathbb{C}$ is an eigenvalue of $A$ if and only if

$$
\operatorname{det}\left(K G^{-1}(s)+Q\right)=0
$$

Thus, in order to prove that all eigenvalues of $A$ lie in a strip, it is sufficient to show that there exists a constant $s_{0} \in \mathbb{R}$ such that $\operatorname{det}\left(K G(s)^{-1}+Q\right) \neq 0$ for $\operatorname{Re} s \leqslant s_{0}$. Due to (3.37), we can write $G(s)=\operatorname{diag}\left(\mathrm{e}^{-h_{i} s}\right)$ with $h_{i}$ positive for $i=1, \ldots, d$. Thus, it yields $G(s)^{-1}=\operatorname{diag}\left(\mathrm{e}^{h_{i} s}\right)$ and the determinant of $K G(s)^{-1}+Q$ can be written as a sum of exponentials with $0 \leqslant \tilde{h}_{j}<\tilde{h}_{j+1}$.

$$
\operatorname{det}\left(K G(s)^{-1}+Q\right)=\sum_{j=1}^{N} a_{j} \mathrm{e}^{\tilde{\mathrm{h}}_{j} s},
$$

where $N \in \mathbb{N}$ and $a_{j} \neq 0$. Then

$$
\mathrm{e}^{-\tilde{h}_{1} s} \operatorname{det}\left(K G(s)^{-1}+Q\right)=\sum_{j=1}^{N} a_{j} \mathrm{e}^{\left(\tilde{h}_{j}-\tilde{h}_{1}\right) s} \rightarrow a_{1} \neq 0 \text { for } \operatorname{Re} s \rightarrow-\infty
$$

and thus, $\operatorname{det}\left(K G(s)^{-1}+Q\right) \neq 0$ for $\operatorname{Re} s<s_{0}$ and some $s_{0} \in \mathbb{R}$. Thus, all eigenvalues of $A$ lie in a strip.

## Chapter 4

## Exact controllability of port-Hamiltonian systems

In this chapter, we consider infinite-dimensional linear port-Hamiltonian systems on a one-dimensional spatial domain with boundary control. In Chapter 3, Definition 3.3.3 we have seen that these systems are of the form

$$
\begin{align*}
\frac{\partial x}{\partial t}(\zeta, t) & =\left(P_{1} \frac{\partial}{\partial \zeta}+P_{0}\right)(\mathcal{H}(\zeta) x(\zeta, t)), \\
x(\zeta, 0) & =x_{0}(\zeta), \zeta \in(0,1)  \tag{4.1}\\
u(t) & =\widetilde{W}_{B}\left[\begin{array}{l}
(\mathcal{H} x)(1, t) \\
(\mathcal{H} x)(0, t)
\end{array}\right], t \geqslant 0
\end{align*}
$$

and Assumption 3.1.1 is fulfilled.
Provided the port-Hamiltonian system (4.1) is well-posed, we aim to characterize exact controllability. Exact controllability is a desirable property of a controlled partial differential equation and has been extensively studied, see for example Kom94, CZ95, TW09]. Triggiani Tri91] showed that exact controllability does not hold for many hyperbolic partial differential equations. However, in this chapter we prove, that the port-Hamiltonian system (4.1) is exactly controllable whenever it is well-posed. The main result of this chapter is published in JK19a.

### 4.1 Sufficient condition for exact controllability

This section is devoted to the main result of this chapter, that is, we show that every well-posed port-Hamiltonian system (4.1) is exactly controllable. We remind the definition of exact controllability and start this section with a characterization of exact controllability via optimizability. For the definition of exact controllability see Definition 2.2.7. The definition of optimizability and the following statement is extracted from RW97.

Definition 4.1.1. Let $\mathfrak{S}(\mathfrak{A}, \mathfrak{B})$ denote a boundary control system and let $x$ be its mild solution. Then $\mathfrak{S}(\mathfrak{A}, \mathfrak{B})$ is called optimizable if for every initial value
$x_{0} \in X$ there exists an input function $u \in L^{2}(0, \infty ; U)$ such that

$$
\int_{0}^{\infty}\|x(t)\|^{2} d t<\infty
$$

Note that exact controllability implies optimizability.
Proposition 4.1.2. ([RW97, Corollary 2.2]) The system $\mathfrak{S}(\mathfrak{A}, \mathfrak{B})$ is exactly controllable if $\mathfrak{S}(\mathfrak{A}, \mathfrak{B})$ is optimizable and $-A$ generates a bounded $C_{0}$-semigroup.

Since impedance energy preserving systems are impedance passive as well, the following statement is useful for the consideration of impedance passive systems.

Proposition 4.1.3. ([Vil07, Theorem 5.1] and [HP18, Lemma 7]) A boundary control and observation port-Hamiltonian system with boundary conditions described using $W_{B}$ such that $W_{B} \Sigma W_{B}^{*}>0$ is exponentially stable. An impedance passive port-Hamiltonian system can be exponentially stabilized via a feedback $u(t)=-k y(t), k>0$.

Exact controllability for impedance energy preserving boundary control and observation port-Hamiltonian system has been studied in JZ18.

Proposition 4.1.4. ([JZ18, Corollary 10.7]) An impedance energy preserving port-Hamiltonian system (3.23) is exactly controllable.

For completeness we include the proof of Proposition 4.1.4.
Proof: As the port-Hamiltonian system (3.23) is impedance energy preserving the corresponding operator $A$ generates a unitary strongly continuous group, cf. Remark 3.3.8. Thus, $-A$ generates a bounded strongly continuous semigroup and exact controllability is equivalent to optimizability, cf. Proposition 4.1.2. Thus it is sufficient to show that the port-Hamiltonian system (3.23) is optimizable. Let $x_{0} \in X$ be arbitrarily. Using Proposition 4.1.3 that for every $k>0$ the choice $u(t)=-k y(t)$ leads to a mild solution in $L^{2}(0, \infty ; X)$. This shows optimizability of system (3.23) and concludes the proof.
Now we can formulate the main result of this chapter.
Theorem 4.1.5. Every well-posed port-Hamiltonian system (4.1) is exactly controllable.

For the proof of this result we need the following lemmas.
Lemma 4.1.6. Let $\left[\begin{array}{ll}W_{1} & W_{0}\end{array}\right] \in \mathbb{C}^{d \times 2 d}$ have full row rank with $W_{1}, W_{0} \in \mathbb{C}^{d \times d}$. Then, there exist invertible matrices $\widetilde{R}_{1}, \widetilde{R}_{0} \in \mathbb{C}^{d \times d}$ such that

$$
\left[\begin{array}{ll}
W_{1} & W_{0}
\end{array}\right]\left[\begin{array}{l}
\widetilde{R}_{1} \\
\widetilde{R}_{0}
\end{array}\right]=I .
$$

Proof: Let [ $\left.\begin{array}{ll}W_{1} & W_{0}\end{array}\right]$ have full row rank with $\operatorname{rank} W_{1}=d-k, k \in$ $\{0, \ldots, d\}$, and rank $W_{0}=d-\ell$ with $\ell \in\{0, \ldots, d\}$. Clearly $d-k+d-\ell \geqslant d$, or equivalently, $k+\ell \leqslant d$.

By $W_{1}^{d-k}$ we denote the first $d-k$ rows of $W_{1}$ and $W_{1}^{k}$ denotes the last $k$ rows. Similarly, by $W_{0}^{d-\ell}$ we denote the last $d-\ell$ rows of $W_{0}$ and by $W_{0}^{\ell}$ the first $\ell$ rows. That is

$$
W_{1}=\left[\begin{array}{c}
W_{1}^{d-k} \\
W_{1}^{k}
\end{array}\right] \quad \text { and } \quad W_{0}=\left[\begin{array}{c}
W_{0}^{\ell} \\
W_{0}^{d-\ell}
\end{array}\right] .
$$

Without loss of generality, using row reduction and the fact that it yields $\operatorname{rank}\left[\begin{array}{ll}W_{1} & W_{0}\end{array}\right]=d$, we may assume that $W_{1}^{k}=0$ and that $W_{1}^{d-k}$ and $W_{0}^{d-\ell}$ have full row rank.
We choose right inverses $R_{1}^{d-k} \in \mathbb{C}^{d \times(d-k)}$ for $W_{1}^{d-k}$ and $R_{0}^{d-\ell} \in \mathbb{C}^{d \times(d-\ell)}$ for $W_{0}^{d-\ell}$. Thus,

$$
W_{1}^{d-k} R_{1}^{d-k}=I \quad \text { and } \quad W_{0}^{d-\ell} R_{0}^{d-\ell}=I .
$$

Clearly, the columns of $R_{1}^{d-k}$ and $R_{0}^{d-\ell}$ are linearly independent and are not elements of the kernel of $W_{1}$ and $W_{0}$, respectively.
Let $R_{1}^{k} \in \mathbb{C}^{d \times k}$ consisting of columns spanning the kernel of $W_{1}$, and let $R_{0}^{\ell} \in \mathbb{C}^{d \times \ell}$ consisting of columns spanning the kernel of $W_{0}$. We define $R_{1}=$ $\left[\begin{array}{ll}R_{1}^{d-k} & R_{1}^{k}\end{array}\right] \in \mathbb{C}^{d \times d}$ and $R_{0}=\left[\begin{array}{ll}R_{0}^{\ell} & R_{0}^{d-\ell}\end{array}\right] \in \mathbb{C}^{d \times d}$. Thus, $R_{1}$ and $R_{0}$ are invertible and it yields

$$
\begin{aligned}
& W_{1} R_{1}+W_{0} R_{0} \\
& =\left[\begin{array}{cc}
I_{d-k} & 0_{(d-k) \times k} \\
0_{k \times(d-k)} & 0_{k \times k}
\end{array}\right]+\left[\begin{array}{cc}
0_{\ell \times \ell} & W_{0}^{l} R_{0}^{d-\ell} \\
0_{(d-\ell) \times \ell} & I_{d-\ell}
\end{array}\right] .
\end{aligned}
$$

Thus, $W_{1} R_{1}+W_{0} R_{0}:=M$ is invertible as an upper triangular matrix and we define $\widetilde{R}_{1}:=R_{1} M^{-1}$ and $\widetilde{R}_{0}:=R_{0} M^{-1}$ to obtain the assertion of the lemma.

Lemma 4.1.7. Let $\alpha \neq 0$ and $\mathfrak{S}(\mathfrak{A}, \mathfrak{B})$ be a well-posed port-Hamiltonian system. Then the port-Hamiltonian system $\mathfrak{S}(\mathfrak{A}, \alpha \mathfrak{B})$ is well-posed as well. Moreover, the system $\mathfrak{S}(\mathfrak{A}, \mathfrak{B})$ is exactly controllable if and only if $\mathfrak{S}(\mathfrak{A}, \alpha \mathfrak{B})$ is exactly controllable.

Proof: The well-posedness of the scaled system follows immediately. The controllability of the two systems is equivalent, since we can scale the input function $u$ of one system by $\alpha$ or $\frac{1}{\alpha}$ to get an input for the other system without changing the mild solution.
Using the results above, we can now give the proof of the main result of this chapter.
Proof of Theorem 4.1.5; We start with an arbitrary port-Hamiltonian system (4.1) described by the tuple $\mathfrak{S}(\mathfrak{A}, \mathfrak{B})$.
By Lemma4.1.7, this system is exactly controllable if and only if for some $\alpha>0$ the system $\mathfrak{S}(\mathfrak{A}, \alpha \mathfrak{B})$ is exactly controllable. We aim to prove that there exists an $\alpha>0$ such that the system $\mathfrak{S}(\mathfrak{A}, \alpha \mathfrak{B})$ is exactly controllable.
Thus, we aim to write the system $\mathfrak{S}(\mathfrak{A}, \alpha \mathfrak{B})$ as a closed-loop system of an exactly controllable system $\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{o}, \mathfrak{C}_{o}\right)$. To construct $\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{o}, \mathfrak{C}_{o}\right)$ we find an impedance energy preserving system $\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{o}, \widetilde{\mathfrak{C}}\right)$ which is exactly controllable by Proposition 4.1.4.

By (3.20) and (3.27), the operator $\mathfrak{B}$ is described by a full row rank $d \times 2 d$ matrix

$$
W_{B}=\left[\begin{array}{ll}
W_{1} & W_{0}
\end{array}\right] .
$$

Using Lemma 4.1 .6 there exists a matrix $R=\left[\begin{array}{l}R_{1} \\ R_{0}\end{array}\right] \in \mathbb{C}^{2 d \times d}$ such that

$$
W_{B} R=I
$$

and $R_{1}, R_{0} \in \mathbb{C}^{d \times d}$ are invertible. If $W_{0}=0$, without loss of generality we may assume that $R_{0}=I$ and $R_{1}=W_{1}^{-1}$.
We now consider the port-Hamiltonian system $\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{o}, \widetilde{\mathfrak{C}}\right)$, where

$$
\mathfrak{B}_{o} x=\left[\begin{array}{ll}
R_{1}^{-1} & 0
\end{array}\right]\left[\begin{array}{l}
f_{\delta, \mathcal{H} x} \\
e_{\delta, \mathcal{H} x}
\end{array}\right]
$$

and

$$
\widetilde{\mathfrak{C}} x=\left[\begin{array}{ll}
0 & R_{1}^{*}
\end{array}\right]\left[\begin{array}{l}
f_{\delta, \mathcal{H} x} \\
e_{\delta, \mathcal{H} x}
\end{array}\right] .
$$

Obviously, the port-Hamiltonian system $\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{0}, \widetilde{\mathfrak{C}}\right)$ is impedance energy preserving. Then it follows from Proposition 4.1.4 that $\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{o}, \widetilde{\mathfrak{C}}\right)$ is exactly controllable.
If $W_{0}=0$, then $\mathfrak{S}(\mathfrak{A}, \mathfrak{B})=\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{o}\right)$ and thus the statement is proved with $\alpha=1$.
We now assume that $W_{0} \neq 0$. In this case we consider the port-Hamiltonian system $\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{o}, \mathfrak{C}_{o}\right)$, where

$$
\mathfrak{C}_{o} x=\left[\begin{array}{ll}
\alpha R_{1}^{-1} & \alpha R_{0}^{-1}
\end{array}\right]\left[\begin{array}{l}
f_{\delta, \mathcal{H} x} \\
e_{\delta, \mathcal{H} x}
\end{array}\right] .
$$

The constant $\alpha>0$ will be chosen later. The matrix $\left[\begin{array}{cc}R_{1}^{-1} & 0 \\ \alpha R_{1}^{-1} & \alpha R_{0}^{-1}\end{array}\right]$ is invertible and the port-Hamiltonian system $\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{o}, \mathfrak{C}_{o}\right)$ is still exactly controllable, since changing the output does not influence controllability.
The port-Hamiltonian system $\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{o}, \mathfrak{C}_{o}\right)$ is regular, see Theorem 3.3.5 and Lemma 3.3.13. By $D$ we denote the feedthrough operator of $\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{o}, \mathfrak{C}_{o}\right)$ and we choose

$$
\alpha= \begin{cases}2\|D\|\left\|W_{0} R_{0}\right\|, & D \neq 0 \\ 1, & D=0\end{cases}
$$

Then $\alpha>0$ and the matrix

$$
F=\frac{1}{\alpha} W_{0} R_{0}
$$

is an admissible feedback operator for $\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{o}, \mathfrak{C}_{o}\right)$ as $\|D F\|<1$ (which implies invertibility of $I-D F)$.

We now consider the closed-loop system as shown in Figure 4.1 and obtain

$$
\begin{aligned}
& \dot{x}(t)=\mathfrak{A} x(t), \quad x(0)=x_{0}, \\
& u_{\alpha}(t)=\alpha\left(u_{o}(t)-F y_{o}(t)\right) \\
&=\alpha\left(\mathfrak{B}_{o}-F \mathfrak{C}_{o}\right) x(t) \\
&=\left(\alpha\left[\begin{array}{ll}
R_{1}^{-1} & 0
\end{array}\right]-W_{0} R_{0}\left[\alpha R_{1}^{-1}\right.\right. \\
&\left.\left.\alpha R_{0}^{-1}\right]\right)\left[\begin{array}{c}
f_{\delta, \mathcal{H} x} \\
e_{\delta, \mathcal{H} x}
\end{array}\right] \\
&=\alpha W_{B}\left[\begin{array}{c}
f_{\delta, \mathcal{H} x} \\
e_{\delta, \mathcal{H} x}
\end{array}\right] .
\end{aligned}
$$

Thus, the closed-loop system equals the port-Hamiltonian system $\mathfrak{S}(\mathfrak{A}, \alpha \mathfrak{B})$. As the open-loop system $\mathfrak{S}\left(\mathfrak{A}, \mathfrak{B}_{o}, \mathfrak{C}_{o}\right)$ is exactly controllable, by Theorem 2.2.25 the port-Hamiltonian system $\mathfrak{S}(\mathfrak{A}, \alpha \mathfrak{B})$ is exactly controllable.


Figure 4.1: $\mathfrak{S}(\mathfrak{A}, \alpha \mathfrak{B})$ as a closed-loop system
Thus, every well-posed port-Hamiltonian system is exactly controllable.
We close this section with an example, where we apply Theorem 4.1.5.
Example 4.1.8. Continuation of Example 3.1.7 and 3.2.8. In Example 3.1.7we have seen that an (undamped) vibrating string can be modelled as the portHamiltonian system (3.8). Its boundary control is given by

$$
\left[\begin{array}{ll}
\widetilde{W}_{1} & \widetilde{W}_{0}
\end{array}\right]\left[\begin{array}{l}
(\mathcal{H} x)(1, t)  \tag{4.2}\\
(\mathcal{H} x)(0, t)
\end{array}\right]=u(t),
$$

where $\left[\begin{array}{ll}\widetilde{W}_{1} & \widetilde{W}_{0}\end{array}\right]$ is a $2 \times 4$-matrix with rank 2 . Using the diagonalization of $P_{1} \mathcal{H}$, see equation (3.16) and Theorem 3.2.5, it is easy to see that the portHamiltonian system (3.8), (4.2) is well-posed if and only if

$$
\widetilde{W}_{1}\left[\begin{array}{c}
\gamma(1) \\
T(1)
\end{array}\right] \oplus \widetilde{W}_{0}\left[\begin{array}{c}
-\gamma(0) \\
T(0)
\end{array}\right]=\mathbb{C}^{2}
$$

cf. JMZ15, or equivalently if the vectors $\widetilde{W}_{1}\left[\begin{array}{l}\gamma(1) \\ T(1)\end{array}\right]$ and $\widetilde{W}_{0}\left[\begin{array}{c}-\gamma(0) \\ T(0)\end{array}\right]$ are linearly independent. By Theorem 4.1.5 the port-Hamiltonian system (3.8), (4.2) is exactly controllable if the vectors $\widetilde{W}_{1}\left[\begin{array}{c}\gamma(1) \\ T(1)\end{array}\right]$ and $\widetilde{W}_{0}\left[\begin{array}{c}-\gamma(0) \\ T(0)\end{array}\right]$ are linearly independent. Here we consider $\widetilde{W}_{1}:=I$ and $\widetilde{W}_{0}:=\left[\begin{array}{cc}-1 & 0 \\ 0 & 1\end{array}\right]$, see also Example 3.2.8. Then the port-Hamiltonian system (3.8), (4.2) is exactly controllable if
the vectors $\left[\begin{array}{l}\gamma(1) \\ T(1)\end{array}\right]$ and $\left[\begin{array}{l}\gamma(0) \\ T(0)\end{array}\right]$ are linearly independent, i.e., it depends not only on the boundary conditions but also on the physical coefficients $T(\zeta)$ and $\rho(\zeta)$ whether the associated port-Hamiltonian operator $A$ generates a $C_{0}$-semigroup.

### 4.2 Closing remarks and open problems

In this chapter we have studied the notion of exact controllability for a class of linear port-Hamiltonian systems on a one dimensional spacial domain with full boundary control and no internal damping. We showed that for this class well-posedness implies exact controllability. Further, we applied the obtained results to the wave equation. By duality a well-posed port-Hamiltonian system $\mathfrak{S}(\mathfrak{A}, \mathfrak{B}, \mathfrak{C})$ with state space $L^{2}\left(0, \infty ; \mathbb{C}^{d}\right)$ and output space $\mathbb{C}^{d}$ is exactly observable. An interesting problem for future research is the characterization of exact controllability for port-Hamiltonian systems with internal damping, i.e., portHamiltonian systems where $P_{0}$ is not necessarily skew-adjoint. We note, that the condition that $\widetilde{W}_{B}$ has full rank cannot be neglected, as in general without full boundary control a port-Hamiltonian system is not exact controllable. Further results for the approximate observability of port-Hamiltonian systems with internal damping can be found in JZ21. In particular, there is shown port-Hamiltonian systems with internal damping are not exactly controllable in general.
Another open question is the characterization of exact controllability for portHamiltonian systems of higher order, see [Vil07]. However, for these systems even the characterization of well-posedness is an open problem.

## Chapter 5

## Riesz bases of port-Hamiltonian systems

It is well-known that the eigenvectors of a compact self-adjoint operator form an orthonormal basis of the underlying Hilbert space. In the 1960s Dunford and Schwartz DS71 introduced the more general notion of spectral operators. Further, Curtain [Cur84] analysed discrete spectral operators, i.e., spectral operators with compact resolvent, and the class of Riesz spectral operators was formulated in [CZ95] and extended in [GZ01] to characterize also operators with multiple eigenvalues. For Riesz spectral operators its eigenvectors still form a basis, but this basis is assumed to be Riesz basis. Since a Riesz basis is isomorphic to an orthonormal basis, many of the nice properties of compact self-adjoint operators carry over to Riesz spectral operators. For instance, solutions of the abstract differential equation $\dot{x}(t)=A x(t)+B u(t)$, with $A$ a Riesz spectral operator, can still described by an eigenfunction expansion of non-harmonic Fourier series. This enables that many properties of these infinite-dimensional systems such as stability, stabilizability and controllability can be characterized in an elegant manner, see e.g. [CZ95, CZ20].

In this chapter, we investigate the Riesz basis property of a special class of infinite-dimensional systems, namely port-Hamiltonian systems on a one-dimensional spatial domain. Here by the Riesz basis property we mean that the associated system operator is a discrete Riesz spectral operator, see Definition 5.2.2.

First, we start with a short introduction to the concept of bases in a infinitedimensional vector space and define Riesz bases. Then we give two toy examples in which the Riesz basis consisting of eigenvectors of a port-Hamiltonian operator can be computed exactly and in which we see that these methods are limited to these simplified situations. Finally we give a characterization for discrete spectral operators, where it is not necessary to determine eigenvalues and eigenfunctions. The main result of this chapter is published in JKZ20 at arXiv and also submitted.

### 5.1 Preliminaries of bases

Although there is a wide theory of bases in the Banach space setting, we will only consider Hilbert spaces, due to the fact that we study port-Hamiltonian systems on the Hilbert space $X=L^{2}\left(0,1, \mathbb{C}^{d}\right)$. Nevertheless, we start to recall some notations and definitions, which also hold in the Banach space setting and can for example be found in [GW19, You80, CZ95, AK06].
We recall that a finite sequence of vectors $\left(x_{n}\right)_{n=1}^{N}$ is linearly independent if the only linear combination for the null is trivial, i.e.,

$$
\sum_{n=1}^{N} a_{n} x_{n}=0 \Leftrightarrow a_{n}=0 \text { for } n=1, \ldots, N .
$$

A Hamel basis is a sequence of vectors with which each element of the vector space can be represented as a finite linear combination.
In the setting of finite-dimensional vector spaces the term basis usually means a Hamel basis. Due to the axiom of choice also each infinite-dimensional vector space has a Hamel basis, but it consists of more than countable elements and thus, the possibilities of their applications are limited.
In 1927 Julius Schauder introduced an additional type of basis, cf. Sch27, the so called Schauder basis. This has the advantage that such a basis of an infinite-dimensional separable space is countable.

Definition 5.1.1. Let be $X$ a Banach space. A sequence $\left(x_{n}\right)_{n \in \mathbb{N}}$ in $X$ is a basis of $X$, if every $x \in X$ has a unique representation with a sequence of complex numbers $\left(a_{n}\right)_{n \in \mathbb{N}}$ such that

$$
\begin{equation*}
x=\sum_{n=1}^{\infty} a_{n} x_{n} \tag{5.1}
\end{equation*}
$$

Thus, in the following the term basis of infinite-dimensional spaces means a Schauder basis and thus, every $x \in X$ can be uniquely represented as a convergent series. In general, a Schauder basis is not a Hamel basis, since infinite linear combinations are allowed and so the linear span of a Schauder basis must be dense in $X$, but it may not be the entire space.
In the above definition the convergence in (5.1) holds in the norm topology $\lim _{n \rightarrow \infty}\left\|x-\sum_{i=1}^{n} a_{i} x_{i}\right\|=0$. In the following definition we introduce the more restrictive concept of unconditional basis.

Definition 5.1.2. An unconditional bases is a basis, where (5.1) converges unconditionally, i.e., also all reorderings of the series (5.1) are convergent.

Definition 5.1.3. Let $X$ be a Hilbert space. A sequence $\left(x_{n}\right)_{n \in \mathbb{N}}$ in $X$ is called an orthonormal basis of $X$, if $\left(x_{n}\right)_{n \in \mathbb{N}}$ is a basis and

$$
\left\langle x_{n}, x_{m}\right\rangle=\left\{\begin{array}{l}
1, \text { if } n=m \\
0, \text { if } n \neq m
\end{array}\right.
$$

Example 5.1.4. It is known that the sequence $\left(\frac{1}{\sqrt{2 \pi}} e^{i n \cdot}\right)_{n \in \mathbb{N}}$ is an orthonormal basis of $L^{2}(0,2 \pi ; \mathbb{C})$, cf. Wer00. Using the variable transformation $x \mapsto \frac{x}{2 \pi}$, we see that $\left(e^{2 \pi i n \cdot}\right)_{n \in \mathbb{N}}$ is an orthonormal basis of $L^{2}(0,1 ; \mathbb{C})$.
We define the term equivalence for bases and introduce Riesz bases afterwards.
Definition 5.1.5. Two bases $\left(x_{n}\right)_{n \in \mathbb{N}}$ and $\left(y_{n}\right)_{n \in \mathbb{N}}$ for a Banach space $X$ are equivalent if and only if there exists a boundedly invertible operator $T: X \rightarrow X$ such that $T x_{n}=y_{n}$ for all $n \in \mathbb{N}$.

Definition 5.1.6. A Riesz basis is a basis which is equivalent to an orthonormal basis.

Remark 5.1.7. Due to Definition 5.1.6 and 5.1.5 a Riesz basis is equivalent to every orthonormal basis.

Definition 5.1.8. A basis $\left(x_{n}\right)_{n \in \mathbb{N}}$ of a Hilbert space $X$ is bounded if

$$
0<\inf _{n}\left\|x_{n}\right\|<\sup _{n}\left\|x_{n}\right\|<\infty .
$$

Remark 5.1.9. A Riesz basis is a bounded basis $\left(x_{n}\right)_{n \in \mathbb{N}}$ since every Riesz basis is obtained from an orthonormal basis $\left(e_{n}\right)_{n \in \mathbb{N}}$ by application of a bounded invertible operator. Therefore, we have

$$
\begin{equation*}
\frac{1}{\left\|T^{-1}\right\|} \leqslant\left\|x_{n}\right\| \leqslant\|T\| \quad \forall n \in \mathbb{N} . \tag{5.2}
\end{equation*}
$$

The following lemma provides an important property of Riesz bases and illustrates the relationship between Riesz bases and unconditional bases.

Lemma 5.1.10. The following statements are equivalent:

1. The sequence $\left(x_{n}\right)_{n \in \mathbb{N}}$ is a Riesz basis of $X$.
2. The sequence $\left(x_{n}\right)_{n \in \mathbb{N}}$ is complete in $X$ and there exist positive constants $m_{1}$ and $m_{2}$ such that for an arbitrary number $N \in \mathbb{N}$ and arbitrary scalars $a_{n} \in \mathbb{C}, n=1, \ldots, N$, it holds

$$
\begin{equation*}
m_{1} \sum_{n=1}^{N}\left|a_{n}\right|^{2} \leqslant\left\|\sum_{n=1}^{N} a_{n} x_{n}\right\|^{2} \leqslant m_{2} \sum_{n=1}^{N}\left|a_{n}\right|^{2} . \tag{5.3}
\end{equation*}
$$

3. The sequence $\left(x_{n}\right)_{n \in \mathbb{N}}$ is a bounded unconditional basis.

Proof: The proof of 1.) $\Leftrightarrow$ 2.) can be found for example in GW19, Theorem 2.2] and the proof of the equivalence 1.) $\Leftrightarrow$ 3.) in Hei11, Theorem 7.13].

We close this section with a generalization of the concept of Riesz bases.
Definition 5.1.11. A sequence of closed subspaces $\left\{X_{n}\right\}_{n \in \mathbb{N}}$ in a Hilbert space $X$ is a Riesz basis of subspaces of $X$ if $\operatorname{span}\left\{X_{n}\right\}_{n \in \mathbb{N}}$ is dense and there exists an isomorphism $T \in \mathcal{L}(X)$, such that $\left\{T X_{n}\right\}_{n \in \mathbb{N}}$ is a system of pairwise orthogonal subspaces of $X$.

Remark 5.1.12. If a sequence of vectors $\left(x_{n}\right)_{n \in \mathbb{N}}$ in $X$ is a Riesz basis of $X$, that is, there exists an isomorphism $T \in \mathcal{L}(X)$, such that $\left(T x_{n}\right)_{n \in N}$ is an orthonormal basis of $X$, then clearly $\left\{\operatorname{span} x_{n}\right\}_{n \in \mathbb{N}}$ is a Riesz basis of subspaces of $X$.
The following toy examples show that in special and simplified situations the eigenvalues and eigenfunctions of a port-Hamiltonian operator can be determined exactly. Since this is difficult in general, we give a characterization of the Riesz basis property which is easy to verify in Theorem 5.3.3.

### 5.1.1 Toy examples

The first example is a port-Hamiltonian equation, i.e., a port-Hamiltonian systems with $d=1$ and the second one is a wave equation with constant coefficients.

Lemma 5.1.13. We consider a port-Hamiltonian operator on the interval $[0,1]$, i.e.

$$
A x=P_{1} \frac{\partial}{\partial \zeta}(\mathcal{H} x)+P_{0}(\mathcal{H} x)
$$

on

$$
\mathcal{D}(A)=\left\{x \in L^{2}(0,1 ; \mathbb{C}) \mid \mathcal{H} x \in \mathcal{W}^{1,2}(0,1 ; \mathbb{C}) \& \widetilde{w}_{1}(\mathcal{H} x)(1)+\widetilde{w}_{0}(\mathcal{H} x)(0)=0\right\}
$$

with $\mathcal{H}, \mathcal{H}^{-1} \in L^{\infty}(0,1 ; \mathbb{C})$ and $\mathcal{H}(\zeta) \in(0, \infty), \widetilde{w}_{1}, \widetilde{w}_{0} \in \mathbb{C}$, such that $A$ generates a $C_{0}$-semigroup. Then it holds: The eigenvectors of the operator $A$ form a Riesz basis if and only if $A$ generates a $C_{0}$-group.

Proof: Without loss of generality we may assume $P_{1}>0$ and even $P_{1}=1$. It holds that $A$ generates a $C_{0}$-group if and only if $\widetilde{w}_{1} \neq 0$ and $\widetilde{w}_{0} \neq 0 \mathrm{cf}$. Theorem 3.2.5. Since $A$ generates a $C_{0}$-semigroup by assumption, we have $\widetilde{w}_{1} \neq 0$, cf. Theorem 3.2.5.
The eigenvalue problem $(\mathcal{H} x)^{\prime}(\zeta)+P_{0}(\mathcal{H} x)(\zeta)=\mu x(\zeta)$ is equivalent to

$$
\frac{d}{d \zeta}(\mathcal{H} x)(\zeta)=\left(\mu \mathcal{H}^{-1}(\zeta)-P_{0}\right)(\mathcal{H} x)(\zeta)
$$

and we get the solution

$$
(\mathcal{H} x)(\zeta)=c e^{\int_{0}^{\zeta} \mu \mathcal{H}^{-1}(s)-P_{0} d s} \quad \text { with } c \in \mathbb{R}, c \neq 0
$$

Since $\mathcal{H}, \mathcal{H}^{-1} \in L^{\infty}(0,1 ; \mathbb{C})$, the integral is well-defined. Thus,

$$
\begin{equation*}
x(\zeta)=\mathcal{H}^{-1}(\zeta)(\mathcal{H} x)(\zeta) \in L^{2}(0,1 ; \mathbb{C}) \tag{5.4}
\end{equation*}
$$

Furthermore, 5.4 yields $\mathcal{H} x \in \mathcal{W}^{1,2}(0,1 ; \mathbb{C})$. Therefore, to get that $x \in \mathcal{D}(A)$, only the boundary condition $\widetilde{w}_{1}(\mathcal{H} x)(1)+\widetilde{w}_{0}(\mathcal{H} x)(0)=0$ has to be fulfilled. We define $G(\zeta):=\int_{0}^{\zeta} \mathcal{H}^{-1}(s) d s$. This yields $G(0)=0$ and $G(1)>0$, since $G(\zeta)$ is monotonic increasing. Thus, with $(\mathcal{H} x)(1)=c e^{\mu G(1)} e^{-P_{0}}$ and $(\mathcal{H} x)(0)=c$ the boundary condition becomes

$$
\widetilde{w}_{1} c e^{\mu G(1)} e^{-P_{0}}+\widetilde{w}_{0} c=0 .
$$

If $\widetilde{w}_{0}=0$, this equation has no solution and thus, we have no eigenvalue and therefore $A$ is not a Riesz operator. We define $k:=e^{-P_{0}}$. For $\widetilde{w}_{0} \neq 0$, we obtain eigenvalues $\left(\mu_{n}\right)_{n \in \mathbb{N}}$ with multiplicity one, namely

$$
\begin{aligned}
& \widetilde{w}_{1} c e^{\mu G(1)} e^{-P_{0}}+\widetilde{w}_{0} c=0 \\
& \Leftrightarrow e^{\mu G(1)}=\frac{-\widetilde{w}_{0}}{\widetilde{w}_{1} k} \\
& \Leftrightarrow \ln \left(e^{\mu G(1)}\right)=\ln \left(\frac{-\widetilde{w}_{0}}{\widetilde{w}_{1} k}\right) \\
& \Leftrightarrow \mu_{n} G(1)=\ln \left(\left|\frac{-\widetilde{w}_{0}}{\widetilde{w}_{1} k}\right|\right)+i\left[\cdot \arg \left(\frac{-\widetilde{w}_{0}}{\widetilde{w}_{1} k}\right)+2 \pi n\right] \\
& \Leftrightarrow \mu_{n}=\frac{\ln \left(\left|\frac{-\widetilde{w}_{0}}{\widetilde{w}_{1} k}\right|\right)+i\left[\cdot \arg \left(\frac{-\widetilde{w}_{0}}{\widetilde{w}_{1} k}\right)+2 \pi n\right]}{G(1)}
\end{aligned}
$$

with eigenvectors

$$
\begin{align*}
x_{n}(\zeta) & =\mathcal{H}^{-1}(\zeta) e^{\mu_{n} G(\zeta)} e^{-P_{0} \zeta}  \tag{5.5}\\
& =\mathcal{H}^{-1}(\zeta) e^{(a+i b) G(\zeta)} e^{-P_{0} \zeta} e^{2 \pi i n \frac{G(\zeta)}{G(1)}} \tag{5.6}
\end{align*}
$$

see (5.4), where $a:=\frac{\ln \left(\left|\frac{-\tilde{w}_{0}}{\tilde{w}_{1} k}\right|\right)}{G(1)}$ and $b:=\frac{\arg \frac{-\tilde{w}_{0}}{\tilde{w}_{1} k}}{G(1)}$. Finally, we have to show that these eigenvectors $\left(x_{n}\right)_{n \in \mathbb{N}}$ form a Riesz basis. Since the point spectrum has a uniform gap, we can apply Theorem 1.1 in (Zwa10 and it suffices to prove that the span of eigenvectors $\left(x_{n}\right)_{n \in \mathbb{N}}$ is dense in $L^{2}(0,1 ; \mathbb{C})$. Suppose that $x \in \operatorname{span}\left\{\left(x_{n}\right)_{n \in \mathbb{N}}\right\}^{\perp}$, i.e., for every $n \in \mathbb{N}$

$$
\begin{aligned}
& \int_{0}^{1} x^{*}(\zeta) x_{n}(\zeta) d \zeta=0 \\
\Leftrightarrow & \int_{0}^{1} x^{*}(\zeta) \mathcal{H}^{-1}(\zeta) e^{(a+i b) G(\zeta)} e^{-P_{0} \zeta} e^{2 \pi i n \frac{G(\zeta)}{G(1)}} d \zeta=0 \\
\Leftrightarrow & \int_{0}^{1} x^{*}\left(G^{-1}(G(1) z)\right) e^{(a+i b) G(1) z} e^{-P_{0} G^{-1}(G(1) z)} e^{2 \pi i n z} G(1) d z=0 \\
\Leftrightarrow & \int_{0}^{1} \tilde{x}^{*}(z) e^{2 \pi i n z} d z=0
\end{aligned}
$$

with

$$
\tilde{x}^{*}(z)=x\left(G^{-1}(G(1) z)\right) e^{(a+i b) G(1) z} e^{-P_{0} G^{-1}(G(1) z)} G(1) \text { for } z \in[0,1] .
$$

Here we first used the substitution $z=\frac{G(\zeta)}{G(1)}$ and $d \zeta=\frac{G(1)}{\mathcal{H}^{-1}(\zeta)} d z$ since $G^{\prime}(\zeta)=$ $\mathcal{H}^{-1}(\zeta)$. Since $\left(e^{2 i \pi n z}\right)_{n \in \mathbb{N}}$ is an orthonormal basis of $L^{2}(0,1 ; \mathbb{C})$, see Example 5.1.4 $\tilde{x}^{*}(z)=0$ for $z \in[0,1]$. Therefore, $x\left(G^{-1}(G(1) z)\right)=0$ for $z \in[0,1]$, which implies $x(\zeta)=0$ and thus, the eigenvectors of $A$ form a Riesz basis. The second example can be found in DH20.
Example 5.1.14. We consider the wave equation with constant coefficients $T$ and $\rho$ on the one-dimensional spatial domain with viscous damping on the right side and boundary control and boundary observation at the other side.

$$
\begin{aligned}
\frac{\partial^{2}}{\partial t^{2}} \omega(\zeta, t) & =\frac{1}{\rho} \frac{\partial}{\partial \zeta}(T \omega(\zeta, t)) \\
0 & =T \frac{\partial}{\partial \zeta} \omega(1, t)+\frac{\kappa}{\rho} \frac{\partial}{\partial t} \omega(1, t) \\
u(t) & =\frac{\partial}{\partial \zeta} \omega(0, t) \\
y(t) & =\frac{\partial}{\partial t} \omega(0, t), \zeta \in(0,1), t \geqslant 0
\end{aligned}
$$

where $\kappa>0$ describes the damping constant. Then the associated portHamiltonian operator is described by

$$
\begin{aligned}
A x & =\left[\begin{array}{ll}
0 & 1 \\
1 & 0
\end{array}\right] \frac{\partial}{\partial \zeta}\left(\left[\begin{array}{cc}
\frac{1}{\rho} & 0 \\
0 & T
\end{array}\right] x\right), \quad x \in \mathcal{D}(A), \\
\mathcal{D}(A) & =\left\{x \in X \mid \mathcal{H} x \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{2}\right) \text { and }\left[\begin{array}{cccc}
-\kappa & 1 & 0 & 0 \\
0 & 0 & 0 & \frac{1}{T}
\end{array}\right]\left[\begin{array}{l}
(\mathcal{H} x)(1) \\
(\mathcal{H} x)(0)
\end{array}\right]=0\right\} .
\end{aligned}
$$

Solving the equation $A \varphi_{n}=s_{n} \varphi_{n}$ yields the eigenvectors

$$
\varphi_{n}(\zeta)=\left[\begin{array}{c}
\cosh \left(\frac{\rho}{T} s_{n} \zeta\right) \\
\frac{1}{\rho T} \sinh \left(\frac{\rho}{T} s_{n} \zeta\right)
\end{array}\right]
$$

and the eigenvalues $s_{n}=s_{0}+\sqrt{\frac{T}{\rho}} i \pi n$, where $s_{0}=\frac{1}{2} \sqrt{\frac{T}{\rho}} \ln \left(\frac{\sqrt{\rho T}-\kappa}{\sqrt{\rho T+\kappa}}\right)$.
Using the mapping

$$
M:=\left[\begin{array}{cc}
\cosh \left(\frac{\rho}{T} s_{0} \zeta\right) & -\sqrt{\rho T} \sinh \left(\frac{\rho}{T} s_{0} \zeta\right) \\
i \sinh \left(\frac{\rho}{T} s_{0} \zeta\right) & -i \sqrt{\rho T} \cosh \left(\frac{\rho}{T} s_{0} \zeta\right)
\end{array}\right]
$$

we see that $\left(\varphi_{n}\right)_{n \in \mathbb{N}}$ is a Riesz basis, since

$$
\left(M \varphi_{n}\right)_{n \in \mathbb{N}}=\left(\left[\begin{array}{c}
\cos (n \pi \zeta) \\
\sin (n \pi \zeta)
\end{array}\right]\right)_{n \in \mathbb{N}}
$$

is an orthonormal basis of $X=L^{2}\left(0,1 ; \mathbb{C}^{2}\right)$.

### 5.2 Discrete Riesz spectral operators

The study of the Riesz basis property for infinite-dimensional port-Hamiltonian systems has started with the thesis by Villegas Vil07, Chapter 4]. Using results on first order eigenvalue problems by Tretter Tre00a, Tre00b], he obtained a sufficient condition. However, it is not easy to see when this technical sufficient condition is satisfied.
Many systems have a Riesz basis of eigenfunctions, see e.g. GX04, XG03. In the monograph GW19, Section 4.3] Guo and Wang study the Riesz basis property for a closely related class of systems, that is, hyperbolic systems of the form $\frac{\partial x}{\partial t}=K(\zeta) \frac{\partial x}{\partial \zeta}+C(\zeta) x$ with $K$ and $C$ diagonal. Note, that (almost)
every port-Hamiltonian system on a one dimensional spatial domain can be transformed into a hyperbolic system of this form. However, in general not with a diagonal $C$. Furthermore, the boundary conditions will be more general, see [ZLMV10, JZ12] or the proof of Lemma 5.3.5. Therefore, the main result of this chapter generalizes their theorem GW19, Theorem 4.11]. We remark, that in this situation the notions of Riesz basis of subspaces and Riesz basis with parentheses are equivalent. Moreover, in [XW11] the Riesz basis property is investigated for operators perturbed by output feedback.
The main result shows that a linear infinite-dimensional port-Hamiltonian system on a one-dimensional spatial domain has the Riesz basis property if and only if the system operator generates a strongly continuous group. Here it is important to note that we do not need constant coefficients, nor extra assumption. Since the group property is equivalent to a simple matrix condition, our results enable us to check very quickly whether the Riesz basis property holds, see also the examples in Section 5.4. Our proof combines methods from complex analysis, differential equations and mathematical systems theory. In particular, we use the fact that every well-posed port-Hamiltonian control system (5.15) is exactly controllable in finite time, cf. Chapter 4.
We start with the definition of discrete Riesz spectral operators.
Definition 5.2.1. For an operator $A$ on $X$ we call $\gamma \subset \sigma(A)$ a compact spectral set if $\gamma$ is a compact subset of $\mathbb{C}$ which is open and closed in $\sigma(A)$. The spectral projection on the spectral subset $\gamma$ is defined as

$$
E(\gamma)=\frac{1}{2 \pi i} \int_{\Gamma}(s-A)^{-1} d s
$$

where $\Gamma$ is a closed Jordan curve containing every point of $\gamma$ and no point of $\sigma(A) \backslash \gamma$.

In this chapter operators with compact resolvent are of particular interest. The spectrum of these operators is a denumerable set of points with no finite accumulation point, cf. DS71, Lemma XIX.2]. Furthermore, every point in the spectrum is an eigenvalue which has finite algebraic as well as finite geometric multiplicity, cf. GGK90, Theorem XV.2.3]. If $\left(s_{n}\right)_{n \in \mathbb{N}}$ is the spectrum of an operator with compact resolvent we write $E_{n}:=E\left(\left(s_{n}\right)\right), n \in \mathbb{N}$, for the spectral projection regarding the $n$-th eigenvalue.

Definition 5.2.2. Let $A$ be an operator with compact resolvent and countable spectrum $\sigma(A)=\left(s_{n}\right)_{n \in \mathbb{N}}$. Then $A$ is a discrete Riesz spectral operator, if

1. for every $n \in \mathbb{N}$ there exists $N_{n} \in \mathcal{L}(X)$ such that $A E_{n}=\left(s_{n}+N_{n}\right) E_{n}$,
2. the sequence of closed subspaces $\left(E_{n}(X)\right)_{n \in \mathbb{N}}$ is a Riesz basis of subspaces of $X$.
3. $N:=\sum_{n \in \mathbb{N}} N_{n}$ is bounded and nilpotent.

Remark 5.2.3. If $A$ is a discrete Riesz spectral operator, then clearly $E_{n}$ commute with $A$ and $A$ is equivalent to the infinite matrix

$$
A=\operatorname{diag}\left(A_{1}, A_{2}, \ldots, A_{n}, \ldots\right)
$$

where $A_{n}$ is a square matrix which corresponds to the restriction $\left.A\right|_{E_{n} X}$ of $A$. Then $A=\sum_{n \in \mathbb{N}} i_{n} A_{n} E_{n}$, where $i_{n}$ is the (natural) inclusion operator and $A_{n}$ is identified with $\left.\left(s_{n}+N_{n}\right)\right|_{E_{n} X}$.
Remark 5.2.4. Discrete Riesz spectral operators are spectral operators in the sense of Dunford and Schwartz DS71. However, we additionally assume that the operator has a compact resolvent, which are discrete operators in the sense of Dunford and Schwartz DS71, Definition XIX.1]. Every discrete Riesz spectral operator is a Riesz spectral operator in the sense of Guo and Zwart [GZ01] and these are again spectral operator in the sense of Dunford and Schwartz. In Curtain and Zwart [CZ95] a slightly stronger notion is considered, where all eigenvalues have to be simple. However, they do not require that the operator has a compact resolvent.
Furthermore, we emphasize that a port-Hamiltonian operator which generates a $C_{0}$-semigroup is closed and that its resolvent is compact, see Aug16].
Moreover, we introduce a term for a set of complex numbers each of which are not to close together.

Definition 5.2.5. A set $\left(s_{n}\right)_{n \in \mathbb{N}} \subset \mathbb{C}$ has a uniform gap, if

$$
\inf _{n \neq m}\left|s_{n}-s_{m}\right|>0 \text { for } n, m \in \mathbb{N}
$$

Riesz bases of subspace have the following useful characterizations.
Proposition 5.2.6. ([Zwa10, Definition 1.4]) Let $A$ be an operator with compact resolvent and $\sigma(A)=\left(s_{n}\right)_{n \in \mathbb{N}}$. Then the sequence of subspaces $\left(E_{n}(X)\right)_{n \in \mathbb{N}}$ is a Riesz basis of subspaces of $X$ if and only if there exist positive constants $m_{1}$ and $m_{2}$ such that it holds

$$
m_{1}\|x\|^{2} \leqslant \sum_{n \in \mathbb{N}}\left\|E_{n} x\right\|^{2} \leqslant m_{2}\|x\|^{2}, \quad x \in X .
$$

Lemma 5.2.7, Lemma 5.2.8, Proposition 5.2.9 and Proposition 5.2.10 will be useful for the proof of the main result of this chapter.

Lemma 5.2.7. Let $A$ be a discrete Riesz spectral operator and $M:=\|N\|$, where $N$ is given by Definition 5.2.2. Then there exists a constant $C>0$ such that for $s \in \rho(A)$ with $d(s, \sigma(A))>M$ we have

$$
\begin{equation*}
\left\|(s-A)^{-1}\right\| \leqslant \frac{C}{d(s, \sigma(A))}, \tag{5.7}
\end{equation*}
$$

where $d(s, \sigma(A))$ denotes the distance from $s$ to the spectrum of $A$.
Proof: Let $\sigma(A)=\left(s_{n}\right)_{n \in \mathbb{N}}, E_{n}, N_{n}, N$ as in Definition 5.2.2 and $s \in \rho(A)$ with $d(s, \sigma(A))>M$ be arbitrary. Since $E_{n}$ is a spectral projection, it commutes with $A$ and the resolvent of $A$. By the definition of a discrete spectral operator we have

$$
s-A=\sum_{n=1}^{\infty}\left(\left(s-s_{n}\right)-N_{n}\right) E_{n}
$$

and identifying $\left(s-s_{n}\right)-N_{n}$ with the matrix corresponding to $\left.\left(\left(s-s_{n}\right)-N_{n}\right)\right|_{E_{n} X}$ we obtain

$$
(s-A)^{-1}=\sum_{n=1}^{\infty}\left(\left(s-s_{n}\right)-N_{n}\right)^{-1} E_{n}
$$

Then it holds for $x \in X$

$$
\begin{aligned}
\left\|(s-A)^{-1} x\right\|^{2} & =\left\|\sum_{n \in \mathbb{N}}\left(\left(s-s_{n}\right)-N_{n}\right)^{-1} E_{n} x\right\|^{2} \\
& \leqslant \sum_{n \in \mathbb{N}}\left\|\left(\left(s-s_{n}\right)-N_{n}\right)^{-1} E_{n} x\right\|^{2} \\
& \leqslant \sup _{n \in \mathbb{N}}\left\|\left(\left(s-s_{n}\right)-N_{n}\right)^{-1}\right\|^{2} \sum_{n \in \mathbb{N}}\left\|E_{n} x\right\|^{2} \\
& \leqslant m_{2} \sup _{n \in \mathbb{N}}\left\|\left(\left(s-s_{n}\right)-N_{n}\right)^{-1}\right\|^{2}\|x\|^{2}
\end{aligned}
$$

where $m_{2}$ is the positive constant of the Riesz basis of subspaces $\left(E_{n}\right)_{n \in \mathbb{N}}$, cf. Proposition 5.2.6. Using

$$
\left(\left(s-s_{n}\right)-N_{n}\right)=\left(s-s_{n}\right)\left(I-\frac{1}{\left(s-s_{n}\right)} N_{n}\right)
$$

we get

$$
\left(\left(s-s_{n}\right)-N_{n}\right)^{-1}=\frac{1}{\left(s-s_{n}\right)} \sum_{j=0}^{k_{n}} \frac{1}{\left(s-s_{n}\right)^{j}} N_{n}^{j}
$$

where $k_{n}$ denotes the degree of nilpotency of $N_{n}$. Thus, for $s \in \rho(A)$ such that $d(s, \sigma(A))>M$, it holds

$$
\begin{aligned}
\left\|\left(\left(s-s_{n}\right)-N_{n}\right)^{-1}\right\| & \leqslant \sum_{j=1}^{k_{n}+1} \frac{1}{\left|\left(s-s_{n}\right)\right|^{j}} M^{j-1} \leqslant \sum_{j=1}^{k_{n}+1} \frac{1}{d(s, \sigma(A))^{j}} M^{j-1} \\
& \leqslant \frac{1}{d(s, \sigma(A))} \sum_{j=0}^{\infty}\left(\frac{M}{d(s, \sigma(A))}\right)^{j}
\end{aligned}
$$

which concludes the proof.
Lemma 5.2.8. Let $A$ be a discrete Riesz spectral operator and generator of $a$ $C_{0}$-semigroup, and $P \in \mathcal{L}(X)$. Then there exist constants $K, M>0$ such that for $s \in \rho(A)$ with $d(s, \sigma(A))>M$ we have $s \in \rho(A+P)$ and

$$
\left\|(s-(A+P))^{-1}\right\| \leqslant \frac{K}{d(s, \sigma(A))}
$$

Proof: By Lemma 5.2 .7 there exists $M_{1}, C>0$ such that

$$
\left\|(s-A)^{-1}\right\| \leqslant \frac{C}{d(s, \sigma(A))}
$$

for $s \in \rho(A)$ with $d(s, \sigma(A))>M_{1}$. Set $M:=\max \left\{M_{1}, 2\|P\| C\right\}$. Let $s \in \rho(A)$ with $d(s, \sigma(A))>M$. Then $I-P(s-A)^{-1}$ is invertible and we obtain

$$
\begin{aligned}
\left\|(s-(A+P))^{-1}\right\| & =\left\|(s-A)^{-1}\left[I-P(s-A)^{-1}\right]^{-1}\right\| \\
& \leqslant \frac{C}{d(s, \sigma(A))} \frac{1}{1-\|P\|\left\|(s-A)^{-1}\right\|} \\
& \leqslant \frac{2 C}{d(s, \sigma(A))}
\end{aligned}
$$

which concludes the proof for $K=2 C$.
Proposition 5.2.9. ( $(\widehat{\mathrm{DS} 71} \mid)$ Let $A$ be an operator with $\sigma(A)=\left(s_{n}\right)_{n \in \mathbb{N}}$ such that the family of spectral subspaces is a Riesz basis of subspaces of $X$. Then $A$ has the representation $A=S+N$, where the scalar part $S$ is defined as

$$
\begin{aligned}
S x & :=\sum_{n \in \mathbb{N}} s_{n} E_{n} x, \\
\mathcal{D}(S) & =\left\{x \in X \mid \sum_{n=1}^{\infty}\left\|s_{n} E_{n} x\right\|^{2}<\infty\right\},
\end{aligned}
$$

and $N_{n}:=N E_{n}=\left(A-s_{n}\right) E_{n}$. Furthermore, $N_{n}$ is quasi-nilpotent, i.e., $\sigma\left(N_{n}\right)=\{0\}$ for all $n \in \mathbb{N}$.

Proposition 5.2.10. Let $A$ be a generator of $C_{0}$-group on $X$ with compact resolvent. The eigenvalues are counted with algebraic multiplicity. If the following conditions
I. The span of the (generalized) eigenvectors form a dense set in $X$,
II. The eigenvalues can be decomposed into finitely many sets each having a uniform gap,
are both fulfilled, then $A$ is a discrete Riesz spectral operator.
Proof: By [Zwa10, Theorem 1.1 and Theorem 1.6] it follows that the family of spectral subspaces is a Riesz basis of subspaces of $X$.
Thus, $A$ has the representation $A=S+N$, cf. Proposition 5.2.9 where $S$ denotes the scalar part of the spectral operator $A$ and $N:=\sum_{n \in \mathbb{N}} N_{n}$, where $N_{n}=\left(A-s_{n}\right) E_{n}$ and $N_{n}$ is quasi-nilpotent. To prove that $A$ is a discrete Riesz spectral operator it remains to show that $N$ is bounded and nilpotent.
We can identify $N_{n}$ with a square matrix corresponding to $\left.N_{n}\right|_{E_{n} X}$ and thus $N_{n}$ is bounded and nilpotent.
Since the eigenvalues of $A$ can be decomposed into finitely many sets each having a uniform gap and their algebraic multiplicity is finite, the degree of the nilpotent matrices $N_{n}$ is bounded. Thus, $N$ is nilpotent.
Finally, we verify that $N$ is bounded. Without loss of generality we assume that $A$ generates an exponentially stable $C_{0}$-group. Then by [LW83], there exists an invertible and positive operator $L \in \mathcal{L}(X)$ such that

$$
\begin{equation*}
\langle A x, L x\rangle+\langle x, L A x\rangle=-\langle x, x\rangle \quad \forall x \in \mathcal{D}(A) . \tag{5.8}
\end{equation*}
$$

We define $A_{n}:=A E_{n}=\left(s_{n}+N_{n}\right) E_{n}$, where $E_{n}$ denotes the $n$-th spectral projection, and we identify $A_{n}$ and $N_{n}$ with the corresponding matrices on $E_{n} X$. From now on we fix $n$. Then we get for $x \in X$

$$
\left\langle A E_{n} x, L E_{n} x\right\rangle+\left\langle E_{n} x, L A E_{n} x\right\rangle=-\left\langle E_{n} x, E_{n} x\right\rangle
$$

or equivalently

$$
\begin{equation*}
\left\langle A_{n} E_{n} x, L_{n} E_{n} x\right\rangle+\left\langle E_{n} x, L_{n} A_{n} E_{n} x\right\rangle=-\left\langle E_{n} x, E_{n} x\right\rangle, \tag{5.9}
\end{equation*}
$$

where $L_{n}:=E_{n}^{*} L E_{n}=L_{n}^{*}$. As $L$ is self-adjoint, $L_{n}$ is self-adjoint as well. Again we identify $L_{n}$ with the corresponding matrix on $E_{n} X$ and obtain on $E_{n} X$

$$
\left(s_{n}+N_{n}\right)^{*} L_{n}+L_{n}\left(s_{n}+N_{n}\right)=-I .
$$

Thus we have

$$
\begin{equation*}
N_{n}^{*} L_{n}+L_{n} N_{n}=-I+r_{n} L_{n} \tag{5.10}
\end{equation*}
$$

with $r_{n}:=-2 \operatorname{Re} s_{n}$. Multiplying (5.10) from the right by $N_{n}^{k_{n}-j+1}$ and from the left by $\left(N_{n}^{*}\right)^{k_{n}-j}$ with $j=2,3, \ldots, k_{n}$ results in

$$
\begin{aligned}
\left(N_{n}^{*}\right)^{k_{n}-j+1} L_{n} N_{n}^{k_{n}-j+1} & +\left(N_{n}^{*}\right)^{k_{n}-j} L_{n} N_{n}^{k_{n}-j+2} \\
& =-\left(N_{n}^{*}\right)^{k_{n}-j} N_{n}^{k_{n}-j+1}+r_{n}\left(N_{n}^{*}\right)^{k_{n}-j} L_{n} N_{n}^{k_{n}-j+1}
\end{aligned}
$$

and thus it holds

$$
\begin{align*}
\left\|L_{n}^{1 / 2} N_{n}^{k_{n}-j+1}\right\|^{2} \leqslant & \left\|N_{n}^{k_{n}-j}\right\|\left\|N_{n}^{k_{n}-j+1}\right\|+\left|r_{n}\right|\left\|L_{n}\right\|\left\|N_{n}^{k_{n}-j}\right\|\left\|N_{n}^{k_{n}-j+1}\right\| \\
& +\left\|N_{n}^{k_{n}-j}\right\|\left\|L_{n}\right\|\left\|N_{n}^{k_{n}-j+2}\right\| . \tag{5.11}
\end{align*}
$$

Since $L_{n}$ is boundedly invertible on $E_{n} X$, we get

$$
\begin{equation*}
m\left\|N_{n}^{k_{n}-j+1}\right\|^{2} \leqslant\left\|L_{n}^{1 / 2} N_{n}^{k_{n}-j+1}\right\|^{2} \text { for some } m \text { independent of } n . \tag{5.12}
\end{equation*}
$$

For $j=2$ we use $N_{n}^{k_{n}}=0$ and obtain

$$
m\left\|N_{n}^{k_{n}-1}\right\| \leqslant\left\|N_{n}^{k_{n}-2}\right\|+\left|r_{n}\right|\left\|L_{n}\right\|\left\|N_{n}^{k_{n}-2}\right\| .
$$

Since $A$ is the generator of a $C_{0}$-group, we have $R:=\sup _{n \in \mathbb{N}}\left|r_{n}\right|<\infty$. Moreover, with $M:=\sup _{n \in \mathbb{N}}\left\|L_{n}\right\|<\infty$ and $C:=1+R M$, this implies

$$
\begin{equation*}
\left\|N_{n}^{k_{n}-1}\right\| \leqslant \frac{C}{m}\left\|N_{n}^{k_{n}-2}\right\| . \tag{5.13}
\end{equation*}
$$

For $j=3, \ldots, k_{n}$ we get using (5.12), (5.11), and (5.13) and by induction over j

$$
\begin{equation*}
\left\|N_{n}^{k_{n}-j+1}\right\| \leqslant\left\|N_{n}^{k_{n}-j}\right\| \sum_{l=1}^{j-1} \frac{C}{m^{l}} M^{l-1} \tag{5.14}
\end{equation*}
$$

In particular, for $j=k_{n}$ and using $k_{n} \leqslant K$, this implies

$$
\left\|N_{n}\right\| \leqslant \sum_{l=1}^{k_{n}-1} \frac{C}{m^{l}} M^{l-1} \leqslant \sum_{l=1}^{K-1} \frac{C}{m^{l}} M^{l-1} .
$$

Together with Proposition 5.2.6 this implies that $N$ is bounded.

### 5.3 Discrete Riesz spectral port-Hamiltonian operators

We consider first order linear port-Hamiltonian systems on a one-dimensional spatial domain of the form

$$
\begin{align*}
\frac{\partial x}{\partial t}(\zeta, t) & =\left(P_{1} \frac{\partial}{\partial \zeta}+P_{0}\right)(\mathcal{H}(\zeta) x(\zeta, t)), \\
x(\zeta, 0) & =x_{0}(\zeta),  \tag{5.15}\\
0 & =\left[\begin{array}{ll}
\widetilde{W}_{1} & \widetilde{W}_{0}
\end{array}\right]\left[\begin{array}{l}
(\mathcal{H} x)(1, t) \\
(\mathcal{H} x)(0, t)
\end{array}\right],
\end{align*}
$$

where $\zeta \in[0,1]$ and $t \geqslant 0$ and the Assumption 3.1.1 is fulfilled. In Chapter 3 and 4 a detailed introduction in these kind of systems is given.
Furthermore, we emphasize that a port-Hamiltonian operator which generates a $C_{0}$-semigroup is closed and that its resolvent is compact, see Theorem 3.2.3. Having the Definition 2.2.15 of well-posedness and Theorem 3.3.5 in mind which shows that for a port-Hamiltonian boundary control system (5.15) wellposedness is already satisfied if $A$ generates a $C_{0}$-semigroup, we recall the following definition.
Definition 5.3.1. We call the port-Hamiltonian system (5.15) a well-posed control system if $A$ generates a $C_{0}$-semigroup on $X$ and there exist $\tau>0$ and $m_{\tau} \geqslant 0$ such that for all $x_{0} \in \mathcal{D}(\mathfrak{A})$ and $u \in C^{2}\left([0, \tau] ; \mathbb{C}^{d}\right)$ with $u(0)=$ $\left[\begin{array}{l}\left(\mathcal{H} x_{0}\right)(1,0) \\ \left(\mathcal{H} x_{0}\right)(0,0)\end{array}\right]$ the classical solution $x$ of 5.15$)$ satisfies

$$
\|x(\tau)\|_{X}^{2} \leqslant m_{\tau}\left(\left\|x_{0}\right\|_{X}^{2}+\int_{0}^{\tau}\|u(t)\|^{2} d t\right) .
$$

In the following, we assume that the port-Hamiltonian system (5.15) is a wellposed control system.
Well-posedness implies that for every initial condition $x_{0} \in X$ and every $L^{2}$ control function $u$ the port-Hamiltonian control system has a unique mild solution, cf. Theorem 3.3.12. Due to Theorem 4.1.5 well-posed port-Hamiltonian control systems are always exactly controllable in finite time. As a consequence of exact controllability we obtain that the eigenspaces span the state space.

Proposition 5.3.2. Consider a well-posed port-Hamiltonian control system (5.15) and assume that $A$ generates a $C_{0}$-group on $X$. Let $\sigma(A)=\left(s_{n}\right)_{n \in \mathbb{N}}$. Then

$$
X=\overline{\operatorname{span}_{n \in \mathbb{N}} E\left(\left(s_{n}\right)\right) X},
$$

where $E\left(\left(s_{n}\right)\right)$ is introduced in Definition 5.2.1.
Proof: Follows from [JZ99, Lemma 7.3] together with Proposition 4.1.5. By $Z^{+}(\zeta)$, we denote the span of the eigenvectors of $P_{1} \mathcal{H}(\zeta)$ corresponding to the positive eigenvalues of $P_{1} \mathcal{H}(\zeta)$ and by $Z^{-}(\zeta)$ the span of the eigenvectors of $P_{1} \mathcal{H}(\zeta)$ corresponding to the negative eigenvalues of $P_{1} \mathcal{H}(\zeta)$.

We are now in the position to formulate our main result.

Theorem 5.3.3. Let $A$ be a port-Hamiltonian operator and the generator of a $C_{0}$-semigroup. Then the following is equivalent:

1. $A$ is a discrete Riesz spectral operator.
2. $A$ is the generator of a $C_{0}$-group.
3. $\widetilde{W}_{1} \mathcal{H}(1) Z^{+}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{-}(0)=\widetilde{W}_{1} \mathcal{H}(1) Z^{-}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{+}(0)=\mathbb{C}^{d}$.

If one of the equivalent conditions are satisfied, then $\sigma(A)=\sigma_{p}(A)$ lie in a strip parallel to the imaginary axis, the eigenvalues (counted according to the algebraic multiplicity) can be decomposed into finitely many sets each having a uniform gap and $A$ satisfies the spectrum determined growth assumption, that $i s, \omega_{0}(A)=s(A)$.

The proof of Theorem 5.3.3 will be given in the next Section 5.3.1.
Remark 5.3.4. It is particularly easy to check if a port-Hamiltonian operator $A$ is the generator of a unitary $C_{0}$-group, cf. Theorem 3.2.3. This is actually the case if and only if $W_{B} \Sigma W_{B}^{*}=0$, where $W_{B}:=\left[\widetilde{W}_{1} \widetilde{W}_{0}\right]\left[\begin{array}{cc}P_{1} & -P_{1} \\ I & I\end{array}\right]^{-1}$ and $\Sigma:=\left[\begin{array}{ll}0 & I \\ I & 0\end{array}\right]$. In this case $A$ is even a skew-adjoint operator by Stone's Theorem, cf. Theorem 2.1.16, which implies that the normalized eigenvectors form an orthonormal basis of $X$.

### 5.3.1 Proof of the Main Result

In the following section we give the proof of Theorem 5.3.3. Let us first assume that one and therefore all of the conditions of the theorem are satisfied. As the resolvent of $A$ is compact the spectrum consists of isolated eigenvalues only. That the eigenvalues lie in a strip parallel to the imaginary axis and that they can be decomposed into finitely many sets each having a uniform gap will be shown in the proof of the implication 2$) \Rightarrow 1$ ). Finally, $\omega_{0}(A)=s(A)$ is implied by GZ01, Theorem 2.12].

### 5.3.2 Proof of the equivalence 2$) \Leftrightarrow 3$ ) of Theorem 5.3.3

The operator $A$ generates a $C_{0}$-group, if and only if $A$ and $-A$ generates a $C_{0}{ }^{-}$ semigroup, EN00, Section II.3.11]. In Theorem 3.2.5 it is shown that $A$ is the generator of a $C_{0}$-semigroup if and only if $\widetilde{W}_{1} \mathcal{H}(1) Z^{+}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{-}(0)=\mathbb{C}^{d}$. Since $\mathcal{D}(-A)=\mathcal{D}(A)$ and $\sigma_{p}\left(P_{1} \mathcal{H}\right)=-\sigma_{p}\left(-P_{1} \mathcal{H}\right)$ it follows that $-A$ generates a $C_{0}$-semigroup if and only if $\widetilde{W}_{1} \mathcal{H}(1) Z^{-}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{+}(0)=\mathbb{C}^{d}$.

### 5.3.3 Proof of the implication 2) $\Rightarrow 1$ ) of Theorem 5.3 .3

The following lemma will be useful.
Lemma 5.3.5. Let $s \in \mathbb{C}$ and $P_{1}, P_{0}$ and $\mathcal{H}$ fulfil the conditions for a portHamiltonian operator in Assumption 3.1.1. Then the solutions of the system of ordinary differential equations

$$
\begin{equation*}
s x(\zeta)=\left(P_{1} \frac{d}{d \zeta}+P_{0}\right)(\mathcal{H} x)(\zeta), \quad \zeta \in[0,1] \tag{5.16}
\end{equation*}
$$

denoted by $x(\zeta)=\Psi^{s}(\zeta) x(0)$, satisfy

$$
\widetilde{M} e^{-|\operatorname{Re} s| \tilde{c}_{0} \zeta}\|v\| \leqslant\left\|\Psi^{s}(\zeta) v\right\| \leqslant M e^{|\operatorname{Re} s| c_{0} \zeta}\|v\|, \quad v \in \mathbb{C}^{d}, \zeta \in[0,1]
$$

with constants $M, \widetilde{M}>0$, and $\tilde{c}_{0}, c_{0} \geqslant 0$ independent of $s$ and $\zeta$.
Proof: Writing $\tilde{x}=\mathcal{H} x$, 5.16) can be equivalently formulated as

$$
\mathcal{H}(\zeta) P_{1} \tilde{x}^{\prime}(\zeta)=s \tilde{x}(\zeta)-\mathcal{H}(\zeta) P_{0} \tilde{x}(\zeta)
$$

We write $s=i \omega+r$ with $\omega, r \in \mathbb{R}$ and diagonalize $P_{1} \mathcal{H}(\zeta)=S^{-1}(\zeta) \Delta(\zeta) S(\zeta)$, see Assumption 3.1.1. Thus, $\mathcal{H}(\zeta) P_{1}=S^{*}(\zeta) \Delta(\zeta) S^{-*}(\zeta)$ and we get

$$
\Delta(\zeta) S^{-*}(\zeta) \tilde{x}^{\prime}(\zeta)=i \omega S^{-*}(\zeta) \tilde{x}(\zeta)+\left(r I-S^{-*}(\zeta) \mathcal{H}(\zeta) P_{0} S^{*}(\zeta)\right) S^{-*}(\zeta) \tilde{x}(\zeta)
$$

Using the substitution $z=S^{-*} \tilde{x}$ gives the equivalent differential equation

$$
\begin{equation*}
z^{\prime}(\zeta)=i \omega \Delta^{-1}(\zeta) z(\zeta)+\left(r \Delta^{-1}(\zeta)+Q(\zeta)\right) z(\zeta) \tag{5.17}
\end{equation*}
$$

where

$$
Q(\zeta):=-\Delta^{-1}(\zeta) S^{-*}(\zeta) \mathcal{H}(\zeta) P_{0} S^{*}(\zeta)-\left(S^{-*}\right)^{\prime}(\zeta) S^{*}(\zeta)
$$

Thus, equation (5.16) is equivalent to equation (5.17). Due to the fact that $P_{1} \mathcal{H}(\zeta)$ has real eigenvalues, $\Delta(\zeta)$ is a diagonal, real matrix and $i \omega \Delta^{-1}(\zeta)$ is a diagonal, purely imaginary matrix. We write $\Delta^{-1}(\zeta)=\operatorname{diag}_{k=1, \ldots, n}\left(\alpha_{k}(\zeta)\right)$ with $\alpha_{k}(\zeta):[0,1] \rightarrow \mathbb{R}$ and define $\Phi_{\omega}(\zeta)=\operatorname{diag}\left(\exp \left(-i \omega \int_{0}^{\zeta} \alpha_{k}(\tau) d \tau\right)\right)$ which satisfies

$$
\left\|\Phi_{\omega}(\zeta)\right\|_{\mathcal{L}\left(\mathbb{C}^{d}\right)}=1, \quad \zeta \in[0,1] .
$$

Multiplying (5.17) with $\Phi_{\omega}(\zeta)$, we get

$$
\Phi_{\omega}(\zeta) z^{\prime}(\zeta)-i \omega \Delta^{-1}(\zeta) \Phi_{\omega}(\zeta) z(\zeta)=\Phi_{\omega}(\zeta)\left(r \Delta^{-1}(\zeta)+Q(\zeta)\right) z(\zeta)
$$

or equivalently

$$
\left(\Phi_{\omega}(\zeta) z(\zeta)\right)^{\prime}=\left(r \Delta^{-1}(\zeta)+\Phi_{\omega}(\zeta) Q(\zeta) \Phi_{\omega}^{-1}(\zeta)\right) \Phi_{\omega}(\zeta) z(\zeta)
$$

Using the substitution $y=\Phi_{\omega} z$, this ordinary differential equation becomes

$$
\begin{equation*}
y^{\prime}=\left(r \Delta^{-1}(\zeta)+Q_{\omega}(\zeta)\right) y(\zeta), \tag{5.18}
\end{equation*}
$$

where $Q_{\omega}(\zeta):=\Phi_{\omega}(\zeta) Q(\zeta) \Phi_{\omega}^{-1}(\zeta)$. There exist constants $c_{0}, c_{1} \geqslant 0$, independent of $\omega$, such that

$$
\begin{equation*}
2 \max _{\zeta \in[0,1]}\left\|r \Delta^{-1}(\zeta)+Q_{\omega}(\zeta)\right\| \leqslant|r| c_{0}+c_{1} \tag{5.19}
\end{equation*}
$$

The solution $y$ of (5.18) satisfies

$$
\begin{equation*}
\frac{d}{d \zeta}\|y(\zeta)\|^{2}=y(\zeta)^{*}\left[\left(r \Delta^{-1}(\zeta)+Q_{\omega}(\zeta)\right)+\left(r \Delta^{-1}(\zeta)+Q_{\omega}(\zeta)\right)^{*}\right] y(\zeta) \tag{5.20}
\end{equation*}
$$

This together with (5.19) implies

$$
-\left(|r| c_{0}+c_{1}\right)\|y(\zeta)\|^{2} \leqslant \frac{d}{d \zeta}\|y(\zeta)\|^{2} \leqslant\left(|r| c_{0}+c_{1}\right)\|y(\zeta)\|^{2}
$$

and thus

$$
e^{-\left(|r| c_{0}+c_{1}\right) \zeta}\|y(0)\|^{2} \leqslant\|y(\zeta)\|^{2} \leqslant e^{\left(|r| c_{0}+c_{1}\right) \zeta}\|y(0)\|^{2}
$$

As the mapping $x \mapsto y$ is boundedly invertible on $L^{2}\left(0,1 ; \mathbb{C}^{d}\right)$, with norm independent on $\omega$, the statement follows.
Next we state some results from complex analysis.
Definition 5.3.6. ([AI95, II.1.27]) An entire function $f$ is called an entire function of exponential type, if there exist constants $C$ and $T$ such that $|f(s)| \leqslant$ $C e^{T|s|}$ for all $s \in \mathbb{C}$. Further, an entire function $f$ of exponential type is said to be of sine type, if

1. the zeros of $f$ lie in a strip $\{s \in \mathbb{C}||\operatorname{Im} s| \leqslant h\}$ for some $h \geqslant 0$ and
2. there exist $\hat{\omega} \in \mathbb{R}$ and positive constants $c$ and $C$ such that

$$
c \leqslant|f(r+i \hat{\omega})| \leqslant C
$$

for every $r \in \mathbb{R}$ holds.
Proposition 5.3.7. ([AI95, Proposition II.1.28] (Levin 1961)) If $f$ is of sine type, then its set of zeros counted with algebraic multiplicity is a finite unification of sets each having a uniform gap.
Lemma 5.3.8. A complex number $s \in \mathbb{C}$ is an eigenvalue of a port-Hamiltonian operator $A$ if and only if

$$
\operatorname{det}\left[\widetilde{W}_{1} \mathcal{H}(1) \Psi^{s}(1)+\widetilde{W}_{0} \mathcal{H}(0)\right]=0
$$

where $\Psi^{s}$ is described in Lemma 5.3.5.
Proof: For every $x(0) \in \mathbb{C}^{d}$ there exists a solution of the differential equation $s x(\zeta)=\left(P_{1} \frac{d}{d \zeta}+P_{0}\right)(\mathcal{H} x)(\zeta), \zeta \in[0,1]$. The complex number $s$ is an eigenvalue of $A$ if and only if $x \in \mathcal{D}(A)$ and $A x=s x$. Using Lemma 5.3.5 this is equivalent to

$$
x(\zeta)=\Psi^{s}(\zeta) x(0) \text { and }\left[\begin{array}{ll}
\widetilde{W}_{1} & \widetilde{W}_{0}
\end{array}\right]\left[\begin{array}{l}
(\mathcal{H} x)(1) \\
(\mathcal{H} x)(0)
\end{array}\right]=0 .
$$

Inserting the first equation in the second, we get that $s$ is an eigenvalue of $A$ if and only if $\operatorname{det}\left[\widetilde{W}_{1} \mathcal{H}(1) \Psi^{s}(1)+\widetilde{W}_{0} \mathcal{H}(0)\right]=0$.
Now we are in the situation to give the proof of the implication 2$) \Rightarrow 1$ ).
Proof of the implication 2) $\Rightarrow 1$ ) of Theorem 5.3.3; $\quad$ Assertion 2) implies that the eigenvalues lie in a strip parallel to the imaginary axis. Thus, thanks to Proposition 5.3.2 and Proposition 5.2.10 it suffices to show that the eigenvalues $\left(s_{n}\right)_{n \in \mathbb{N}}$ of $A$ (counted according to the algebraic multiplicity) can be decomposed into finitely many sets having each a uniform gap.
Using Proposition 5.3.7, this is implied by the existence of an entire function $g$ of exponential type with
i) $g$ has exactly the zeros $\left(s_{n}\right)_{n \in \mathbb{N}}$ and
ii) there exist $r, c, C>0$ such that for every $\omega \in \mathbb{R}: c \leqslant|g(r+i \omega)| \leqslant C$.

Note that $f: \mathbb{C} \rightarrow \mathbb{C}$, defined by $f(s):=g(i s)$ is a sine-type function if $g$ is an entire function of exponential type and satisfying the above conditions i) and ii). We define $g: \mathbb{C} \rightarrow \mathbb{C}$ by

$$
\begin{equation*}
g(s):=\operatorname{det}\left[\widetilde{W}_{1} \mathcal{H}(1) \Psi^{s}(1)+\widetilde{W}_{0} \mathcal{H}(0)\right] . \tag{5.21}
\end{equation*}
$$

By Lemma 5.3.8, a complex number $s \in \mathbb{C}$ is an eigenvalue of the operator $A$ if and only if $g(s)=0$.
$\Psi^{s}$ described in Lemma 5.3.5 is an entire function, cf. Was87, Theorem 24.1] and thus $g$ is an entire function as well. Clearly, $g$ has the zeros $\left(s_{n}\right)_{n \in \mathbb{N}}$. Since the determinant of a matrix equals the product of its eigenvalues and every eigenvalue is smaller or equal the norm of the matrix, it yields

$$
|g(s)| \leqslant\left\|\left[\widetilde{W}_{1} \mathcal{H}(1) \Psi^{s}(1)+\widetilde{W}_{0} \mathcal{H}(0)\right]\right\|^{n} .
$$

Using Lemma 5.3.5 it holds $|g(s)| \leqslant c_{2} e^{|\operatorname{Re} s| c_{3}}$ for some constants $c_{2}, c_{3} \geqslant 0$, and thus, $g$ is bounded on lines parallel to the imaginary axis and grows at most exponentially.
Next, we show that $g$ is bounded away from zero on some line parallel to the imaginary axis. Since the control operator $B$ of the port-Hamiltonian system (5.15) is admissible, see Lemma 3.3.12, it yields that for $\omega>\omega_{0}(A)$ exists a constant $M_{\omega}>0$ such that

$$
\begin{equation*}
\left\|\left(s-A_{-1}\right)^{-1} B\right\|_{\mathcal{L}\left(\mathbb{C}^{d}, X\right)} \leqslant \frac{M_{\omega}}{\sqrt{\operatorname{Re} s-\omega}} \quad \text { for } \operatorname{Re} s \geqslant \omega, \tag{5.22}
\end{equation*}
$$

see Proposition 2.2.5. Let $r>\omega_{0}(A)$ and we assume that $g$ is not bounded away from zero on $r+i \mathbb{R}$, i.e., there exists a sequence $\omega_{k} \in \mathbb{R}$ such that $g\left(r+i \omega_{k}\right) \rightarrow 0$. Since all zeros of $g$ have real part less or equal to the growth bound $\omega_{0}(T)$ of the $C_{0}$-semigroup generated by $A$, it holds true that $g\left(r+i \omega_{k}\right) \neq 0$. Let $u_{0}$ be an arbitrary vector in $\mathbb{C}^{d}$. By Proposition 2.2 .20 the solution $x_{u_{0}}^{r+i \omega_{k}}$ of

$$
\begin{array}{r}
\left(r+i \omega_{k}\right) x(\zeta)=\left(P_{1} \frac{d}{d \zeta}+P_{0}\right)(\mathcal{H} x)(\zeta) \\
{\left[\begin{array}{ll}
\widetilde{W}_{1} & \widetilde{W}_{0}
\end{array}\right]\left[\begin{array}{l}
(\mathcal{H} x)(1) \\
(\mathcal{H} x)(0)
\end{array}\right]=u_{0},}
\end{array}
$$

is given by

$$
\begin{aligned}
x_{u_{0}}^{r+i \omega_{k}} & =\left(\left(\left(r+i \omega_{k}\right)-A_{-1}\right)^{-1}\left(\mathfrak{A} \widetilde{B}-\left(r+i \omega_{k}\right) \widetilde{B}\right)+\widetilde{B}\right) u_{0} \\
& =\left(\left(r+i \omega_{k}\right)-A_{-1}\right)^{-1} B u_{0}
\end{aligned}
$$

and $x_{u_{0}}^{r+i \omega_{k}}(0)$ fulfils

$$
\left[\widetilde{W}_{1} \mathcal{H}(1) \Psi^{r+i \omega_{k}}(1)+\widetilde{W}_{0} \mathcal{H}(0)\right] x_{u_{0}}^{r+i \omega_{k}}(0)=u_{0} .
$$

Let $M_{k}:=\left[\widetilde{W}_{1} \mathcal{H}(1) \Psi^{r+i \omega_{k}}(1)+\widetilde{W}_{0} \mathcal{H}(0)\right]$. Since $g\left(r+i \omega_{k}\right)=\operatorname{det}\left(M_{k}\right) \rightarrow 0$ and $1=\operatorname{det}(I)=\operatorname{det}\left(M_{k}\right) \operatorname{det}\left(M_{k}^{-1}\right)$, we have $\operatorname{det}\left(M_{k}^{-1}\right) \rightarrow \infty$. Hence, $M_{k}^{-1}$ has an eigenvalue $\nu_{k}$ with $\left|\nu_{k}\right| \rightarrow \infty$. Choose $u_{k, 0}$ as a normalized eigenvector to $\nu_{k}$. Then it yields

$$
x_{u_{k, 0}}^{r+i \omega_{k}}(0)=M_{k}^{-1} u_{k, 0}=\nu_{k} u_{k, 0},
$$

which implies $\left\|x_{u_{k, 0}}^{r+i \omega_{k}}(0)\right\|_{\mathbb{C}^{d}} \rightarrow \infty$. We note that the function $x_{u_{k, 0}}^{r+i \omega_{k}}$ is given by

$$
x_{u_{k, 0}}^{r+i \omega_{k}}(\zeta)=\Psi^{r+i \omega_{k}}(\zeta) x_{u_{k, 0}}^{r+i \omega_{k}}(0) .
$$

Using that the inverse of $\Psi^{r+i \omega_{k}}$ is a bounded function, see Lemma 5.3.5, we get

$$
\begin{equation*}
\left\|x_{u_{k, 0}}^{r+i \omega_{k}}\right\|_{L^{2}\left((0,1) ; \mathbb{C}^{d}\right)} \rightarrow \infty \tag{5.23}
\end{equation*}
$$

However, since we also have $x_{u_{k, 0}}^{r+i \omega_{k}}=\left(\left(r+i \omega_{k}\right)-A_{-1}\right)^{-1} B u_{k, 0}$, equation (5.23) is in contradiction with the uniform boundedness of $\left(\left(r+i \omega_{k}\right)-A_{-1}\right)^{-1} B u_{0, k}$, see equation (5.22). Thus the entire function $g$ is of exponential type and satisfies condition i) and ii). This concludes the proof.

### 5.3.4 Proof of the implication 1$) \Rightarrow 2$ ) of Theorem 5.3 .3

We start with some characterizations of the resolvent and the spectrum of portHamiltonian operators.
Lemma 5.3.9. Let $\Lambda \in C\left([0,1] ; \mathbb{C}^{d}\right)$ with $\Lambda(\zeta)$ diagonal, invertible and positive definite for every $\zeta \in[0,1], Q \in \mathbb{C}^{d \times d}$ singular and $A: \mathcal{D}(A) \subset L^{2}\left(0,1 ; \mathbb{C}^{d}\right) \rightarrow$ $L^{2}\left(0,1 ; \mathbb{C}^{d}\right)$ defined by $A x=\Lambda x^{\prime}$ and $\mathcal{D}(A)=\left\{x \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{d}\right) \mid x(1)+\right.$ $Q x(0)=0\}$. Then there exist real constants $\gamma<0$ and $a, b>0$ such that $\left\|(s-A)^{-1}\right\| \geqslant a e^{b|s|}$ for real $s \in \rho(A)$ with $s<\gamma$.
Proof: Let $0 \neq x(0) \in \operatorname{ker} Q$. We define $F_{s}(\zeta):=s \int_{0}^{\zeta} \Lambda^{-1}(\tau) d \tau$ for $s \in \mathbb{R}$ and $\zeta \in[0,1]$. Note that $F_{s}$ and $\Lambda$ commute as both are diagonal. There exists $\gamma_{0}<0$ such that $I-e^{2 F_{s}(1)}$ is invertible if $s<\gamma_{0}$. Further, let $s \in \rho(A)$ with $s<\gamma_{0}$ and define

$$
\begin{equation*}
g(\zeta)=2 s e^{F_{s}(1)-F_{s}(\zeta)} g_{0}=2 s e^{s \int_{\zeta}^{1} \Lambda^{-1}(\tau) d \tau} g_{0} \tag{5.24}
\end{equation*}
$$

with $g_{0}:=e^{F_{s}(1)}\left(I-e^{2 F_{s}(1)}\right)^{-1} x(0)$. Then the solution of

$$
\begin{equation*}
\Lambda x^{\prime}=s x+g \tag{5.25}
\end{equation*}
$$

is given by

$$
\begin{aligned}
x(\zeta) & =e^{F_{s}(\zeta)} x(0)+\int_{0}^{\zeta} e^{F_{s}(\zeta)-F_{s}(\tau)} \Lambda^{-1}(\tau) g(\tau) d \tau \\
& =e^{F_{s}(\zeta)} x(0)-e^{F_{s}(\zeta)+F_{s}(1)} \int_{0}^{\zeta}\left(-2 s \Lambda^{-1}(\tau)\right) e^{-2 F_{s}(\tau)} d \tau g_{0} \\
& =e^{F_{s}(\zeta)} x(0)-e^{F_{s}(\zeta)+F_{s}(1)}\left(e^{-2 F_{s}(\zeta)}-I\right) g_{0} \\
& =e^{F_{s}(\zeta)}\left(I-e^{2 F_{s}(1)}\right) e^{-F_{s}(\zeta)} g_{0}-e^{-F_{s}(\zeta)+F_{s}(1)} g_{0}+e^{-F_{s}(\zeta)+F_{s}(1)} g_{0} \\
& =e^{F_{s}(\zeta)-F_{s}(1)} g_{0}-\frac{1}{2 s} g(\zeta) .
\end{aligned}
$$

In particular $x(1)=0$ and thus $(s-A)^{-1} g=x$. As $\Lambda \in C\left([0,1] ; \mathbb{C}^{d}\right)$ is a diagonal and invertible matrix-valued function with positive entries on its diagonal, there exists $\lambda_{0}>0$ such that the diagonal elements of $\Lambda$ are bounded by $\lambda_{0}$. First using (5.24) we can estimate

$$
\|g\|_{L^{2}\left((0,1) ; \mathbb{C}^{d}\right)} \leqslant c \sqrt{|s|}\left\|g_{0}\right\|, c>0
$$

and

$$
\begin{aligned}
\|x\|_{L^{2}\left(0,1 ; \mathbb{C}^{d}\right)} & \geqslant\left\|e^{F_{s}(\cdot)-F_{s}(1)} g_{0}\right\|_{L^{2}\left(0,1 ; \mathbb{C}^{d}\right)}-\frac{1}{2|s|}\|g\|_{L^{2}\left((0,1) ; \mathbb{C}^{d}\right)} \\
& =\left(\int_{0}^{1}\left\|e^{-s \int_{\tau}^{1} \Lambda^{-1}(\sigma) d \sigma} g_{0}\right\|^{2} d \tau\right)^{\frac{1}{2}}-\frac{1}{2|s|}\|g\|_{L^{2}\left((0,1) ; \mathbb{C}^{d}\right)} \\
& \geqslant\left(\int_{0}^{1} e^{2|s|(1-\tau) \lambda_{0}}\left\|g_{0}\right\|^{2} d \tau\right)^{\frac{1}{2}}-\frac{1}{2|s|}\|g\|_{L^{2}\left((0,1) ; \mathbb{C}^{d}\right)} \\
& \geqslant \frac{1}{4|s|^{2} \lambda_{0}}\left(e^{2|s| \lambda_{0}}-1\right)\|g\|_{L^{2}\left((0,1) ; \mathbb{C}^{d}\right)}-\frac{1}{2|s|}\|g\|_{L^{2}\left((0,1) ; \mathbb{C}^{d}\right)} .
\end{aligned}
$$

This completes the proof of lemma.
Lemma 5.3.10. Let $A$ be a port-Hamiltonian operator, which generates a $C_{0}-$ semigroup. Furthermore, let $A$ be a discrete Riesz spectral operator and let $\left(s_{n}\right)_{n \in \mathbb{N}}$ denote its eigenvalues. Then there exists a constant $K>0$ such that for every $n \in \mathbb{N}$ within the ball $\left\{\left.s \in \mathbb{C}\left|\left|s-s_{n}\right| \leqslant K\right| \operatorname{Re} s_{n}\right|^{2}\right\}$ there lie at most $d$ eigenvalues.

Proof: Without lost of generality we assume that $A$ generates an exponentially stable $C_{0}$-semigroup. By Proposition 4.1.5 the corresponding portHamiltonian control system (5.15) with control operator $B$ is exactly controllable in finite time. Then the dual system, described by $A^{*}$ and $B^{*}$, is exactly observable and it yields due to the Hautus Test, cf. Theorem 2.2.10, that there exists a positive constant $m$ such that

$$
\begin{equation*}
\left\|\left(s-A^{*}\right) x\right\|^{2}+|\operatorname{Re} s|\left\|B^{*} x\right\|^{2} \geqslant m|\operatorname{Re} s|^{2}\|x\|^{2}, \quad \operatorname{Re} s<0, x \in \mathcal{D}\left(A^{*}\right) \tag{5.26}
\end{equation*}
$$

No matter that there may exist generalized eigenvectors, we consider only eigenvectors corresponding to different eigenvalues of $A^{*}$. As $\sigma(A)=\overline{\sigma\left(A^{*}\right)}$, it suffices to prove the statement for $A^{*}$. Choose arbitrary $e_{1}, \ldots e_{d+1}$ normed eigenvectors of the operator $A^{*}$ to the eigenvalues $\lambda_{1}, \ldots, \lambda_{d+1}$ with $\lambda_{n} \neq \lambda_{m}$ for $n \neq m \in 1, \ldots, d+1$.
Since $B^{*} \in \mathcal{L}\left(\mathcal{D}\left(A^{*}\right), \mathbb{C}^{d}\right)$ the $d+1$ vectors $B^{*} e_{n}$ are linearly dependent in $\mathbb{C}^{d}$, i.e, there exists scalars $a_{1}, \ldots, a_{d+1} \in \mathbb{C}$ with $\sum_{n=1}^{d+1}\left|a_{n}\right|^{2}=1$ such that

$$
\begin{equation*}
a_{1} B^{*} e_{1}+\ldots a_{d+1} B^{*} e_{d+1}=0 . \tag{5.27}
\end{equation*}
$$

Consider $x=\sum_{n=1}^{d+1} a_{n} e_{n}$. Then $x \in \mathcal{D}\left(A^{*}\right)$ with $B^{*} x=0$, and

$$
\|x\|^{2}=\left\|\sum_{n=1}^{d+1} a_{n} e_{n}\right\| \geqslant m_{1}>0
$$

by Proposition 5.2.6. It follows with the Hautus Test (5.26) at the point $s=$ $\lambda_{d+1}$

$$
\begin{aligned}
m m_{1}\left|\operatorname{Re} \lambda_{d+1}\right|^{2} & \leqslant\left\|\left(\lambda_{d+1}-A^{*}\right) x\right\|^{2}=\left\|\sum_{n=1}^{d} a_{n}\left(\lambda_{d+1}-\lambda_{n}\right) e_{n}\right\|^{2} \\
& \leqslant m_{2} \sum_{n=1}^{d}\left|a_{n}\right|^{2}\left|\lambda_{d+1}-\lambda_{n}\right|^{2}
\end{aligned}
$$

Thus,

$$
\frac{d+1}{d+1} \frac{m m_{1}}{m_{2}}\left|\operatorname{Re} \lambda_{d+1}\right|^{2} \leqslant \sum_{n=1}^{d}\left|\lambda_{d+1}-\lambda_{n}\right|^{2} .
$$

Since the eigenvalues are arbitrary chosen, this implies that in the ball with radius $\frac{m m_{1}}{(d+1) m_{2}}\left|\operatorname{Re} \lambda_{d+1}\right|^{2}$ around $\lambda_{d+1}$ lies at most $d$ eigenvalues. Hence, the statement follows.
Now we are in the position to prove the implication 1$) \Rightarrow 2$ ).
Proof of the implication 1) $\Rightarrow 2$ ) of Theorem 5.3.3: We assume that $A$ does not generate a $C_{0}$-group.
Since $A$ is a port-Hamiltonian operator, we denote by $S$ the matrix-valued function such that $P_{1} \mathcal{H}(\zeta)=S^{-1}(\zeta) \Delta(\zeta) S(\zeta)$, see Assumption 3.1.1. Since the eigenvalues of $P_{1} \mathcal{H}(\zeta)$ and $\mathcal{H}(\zeta)^{\frac{1}{2}} P_{1} \mathcal{H}(\zeta)^{\frac{1}{2}}$ are the same, it follows by Sylvester's law of inertia that the number of positive and negative eigenvalues of $P_{1} \mathcal{H}(\zeta)$ equal those of $P_{1}$. Let $d_{1}$ denote the number of positive and $d_{2}=d-d_{1}$ the number of negative eigenvalues of $P_{1}$. Thus without loss of generality $\Delta$ can be written as $\Delta(\zeta)=\left[\begin{array}{cc}\Lambda(\zeta) & 0 \\ 0 & \Theta(\zeta)\end{array}\right]$, where $\Lambda(\zeta) \in \mathbb{C}^{d_{1} \times d_{1}}$ correspond to the positive eigenvalues and $\Theta(\zeta) \in \mathbb{C}^{d_{2} \times d_{2}}$ to the negative ones.
Let $\mathcal{S} \in \mathcal{L}(X)$ be the multiplication operator $(\mathcal{S} x)(\zeta):=S(\zeta) x(\zeta)$. By assumption $\mathcal{S}$ is invertible and we obtain

$$
\begin{aligned}
\mathcal{S} A \mathcal{S}^{-1} z(\zeta)= & \Delta(\zeta)(z(\zeta))^{\prime} \\
& +\Delta^{\prime}(\zeta) z(\zeta)+S(\zeta)\left(S^{-1}\right)^{\prime} \Delta(\zeta) z(\zeta)+S(\zeta) P_{0} \mathcal{H}(\zeta) S^{-1}(\zeta) z(\zeta) \\
\mathcal{D}\left(\mathcal{S} A \mathcal{S}^{-1}\right)= & \left\{z \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{d}\right) \left\lvert\,\left[\begin{array}{ll}
\widetilde{W}_{1} & \widetilde{W}_{0}
\end{array}\right]\left[\begin{array}{l}
\left(\mathcal{H} S^{-1} z\right)(1) \\
\left(\mathcal{H} S^{-1} z\right)(0)
\end{array}\right]=0\right.\right\} .
\end{aligned}
$$

The operator $\mathcal{S} A \mathcal{S}^{-1}$ generates a $C_{0}$-semigroup, too. We split the variable $z(\zeta)=\left[\begin{array}{l}z^{+}(\zeta) \\ z^{-}(\zeta)\end{array}\right] \in \mathbb{C}^{d}$ with $z^{+}(\zeta) \in \mathbb{C}^{d_{1}}$ and $z^{-}(\zeta) \in \mathbb{C}^{d_{2}}$, and we define $\widetilde{W}_{1} \mathcal{H}(1) S^{-1}(1)=:\left[\begin{array}{ll}V_{1} & V_{2}\end{array}\right]$ and $\widetilde{W}_{0} \mathcal{H}(0) S^{-1}(0)=:\left[\begin{array}{ll}U_{1} & U_{2}\end{array}\right]$, where $U_{1}, V_{1} \in$ $\mathbb{C}^{d \times d_{1}}$ and $U_{2}, V_{2} \in \mathbb{C}^{d \times d_{2}}$. Then, $z \in \mathcal{D}\left(\mathcal{S} A \mathcal{S}^{-1}\right)$ if and only if $z \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{d}\right)$
and

$$
\begin{aligned}
0 & =\left[\begin{array}{ll}
\widetilde{W}_{1} & \widetilde{W}_{0}
\end{array}\right]\left[\begin{array}{l}
\left(\mathcal{H} S^{-1} z\right)(1) \\
\left(\mathcal{H} S^{-1} z\right)(0)
\end{array}\right] \\
& =\left[\begin{array}{ll}
V_{1} & V_{2}
\end{array}\right]\left[\begin{array}{l}
z^{+}(1) \\
z^{-}(1)
\end{array}\right]+\left[\begin{array}{ll}
U_{1} & U_{2}
\end{array}\right]\left[\begin{array}{l}
z^{+}(0) \\
z^{-}(0)
\end{array}\right] \\
& =\left[\begin{array}{ll}
V_{1} & U_{2}
\end{array}\right]\left[\begin{array}{l}
z^{+}(1) \\
z^{-}(0)
\end{array}\right]+\left[\begin{array}{ll}
U_{1} & V_{2}
\end{array}\right]\left[\begin{array}{l}
z^{+}(0) \\
z^{-}(1)
\end{array}\right] \\
& =K\left[\begin{array}{l}
\Lambda(1) z^{+}(1) \\
\Theta(0) z^{-}(0)
\end{array}\right]+Q\left[\begin{array}{l}
\Lambda(0) z^{+}(0) \\
\Theta(1) z^{-}(1)
\end{array}\right],
\end{aligned}
$$

where $K:=\left[\begin{array}{ll}V_{1} & U_{2}\end{array}\right]\left[\begin{array}{cc}\Lambda(1)^{-1} & 0 \\ 0 & \Theta(0)^{-1}\end{array}\right]$ and $Q:=\left[\begin{array}{ll}U_{1} & V_{2}\end{array}\right]\left[\begin{array}{cc}\Lambda(0)^{-1} & 0 \\ 0 & \Theta(1)^{-1}\end{array}\right]$. Since $\mathcal{S} A \mathcal{S}^{-1}$ is the generator of a $C_{0}$-semigroup and this property is invariant under bounded perturbations, $K$ is invertible, see [JZ12, Theorem 13.3.1].
Let $\mathcal{T} \in \mathcal{L}(X)$ be defined by $\mathcal{T}\left[\begin{array}{c}z^{+}(\zeta) \\ z^{-}(\zeta)\end{array}\right]:=\left[\begin{array}{c}z^{+}(\zeta) \\ z^{-}(1-\zeta)\end{array}\right]$. Clearly, $\mathcal{T}$ is invertible. Then the operator $\mathcal{A}:=\mathcal{T} A \mathcal{S}^{-1} \mathcal{T}^{-1}$ on $X$ is given by

$$
\begin{aligned}
& \mathcal{A} z(\zeta)=\left[\begin{array}{cc}
\Lambda(\zeta) & 0 \\
0 & -\Theta(1-\zeta)
\end{array}\right] z^{\prime}(\zeta)+R(\zeta) z(\zeta) \\
& \mathcal{D}(\mathcal{A})=\left\{z \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{d}\right) \mid\right. \\
&\left.\quad K\left[\begin{array}{cc}
\Lambda(1) & 0 \\
0 & \Theta(0)
\end{array}\right] z(1)+Q\left[\begin{array}{cc}
\Lambda(1) & 0 \\
0 & \Theta(0)
\end{array}\right] z(0)=0\right\}
\end{aligned}
$$

where $z \mapsto R(\cdot) z(\cdot)$ is a bounded multiplication operator on $X$. Let $\widetilde{K}:=$ $K\left[\begin{array}{cc}\Lambda(1) & 0 \\ 0 & \Theta(0)\end{array}\right]$ and $\widetilde{Q}:=Q\left[\begin{array}{cc}\Lambda(1) & 0 \\ 0 & \Theta(0)\end{array}\right]$. By assumption, the matrix $\widetilde{K}$ is invertible. As a bounded perturbation of a generator of $C_{0}$-group generates again a $C_{0}{ }^{-}$ group, we obtain that the operator

$$
\begin{aligned}
& \widetilde{A} z(\zeta)=\left[\begin{array}{cc}
\Lambda(\zeta) & 0 \\
0 & -\Theta(1-\zeta)
\end{array}\right] z^{\prime}(\zeta) \\
& \mathcal{D}(\widetilde{A})=\left(z \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{d}\right) \mid z(1)=\widetilde{K}^{-1} \widetilde{Q} z(0)\right\}
\end{aligned}
$$

generates a $C_{0}$-semigroup, but not a $C_{0}$-group. In particular, $Q_{1}:=\widetilde{K}^{-1} \widetilde{Q}$ is singular.
Since $\mathcal{A}$ is a discrete Riesz spectral operator, due to Lemma 5.3.10, there is a $K>0$ such that in the ball with radius $K|\operatorname{Re} s|^{2}$ around every eigenvalue $s$ of $A$ lie at most $d$ eigenvalues. Thus there exist sequences $\left(t_{n}\right)_{n \in \mathbb{N}} \subset \mathbb{R}$ and $\left(r_{n}\right)_{n \in \mathbb{N}} \subset(0, \infty)$ with $t_{n} \rightarrow-\infty$ and $r_{n} \rightarrow \infty$ such that the ball with center $t_{n}$ and radius $r_{n}$ lie in $\rho(A)$. By Lemma 5.2.8 we get

$$
\left\|\left(t_{n}-\widetilde{A}\right)^{-1}\right\| \rightarrow 0
$$

However, for the operator $\widetilde{A}$ the Lemma 5.3.9, is applicable which implies that $\left\|\left(t_{n}-\widetilde{A}\right)^{-1}\right\| \rightarrow \infty$. This leads to a contradiction.

### 5.4 Examples

### 5.4.1 Wave equation with boundary feedback

We consider the one-dimensional wave equation with boundary feedback as in XW11, but we allow for spatial dependent mass density and Young's modulus, given by

$$
\begin{align*}
w_{t t}(\zeta, t) & =\frac{1}{\rho(\zeta)} \frac{\partial}{\partial \zeta}\left(T(\zeta) w_{\zeta}(\zeta, t)\right), \quad \zeta \in[0,1], t \geqslant 0 \\
w(0, t) & =0 \\
u(t) & =T(1) w_{\zeta}(1, t),  \tag{5.28}\\
y(t) & =w_{t}(1, t), \\
u(t) & =-\kappa y(t), \kappa>0,
\end{align*}
$$

where $\zeta \in[0,1]$ is the spatial variable, $w(\zeta, t)$ describes the displacement of the point $\zeta$ of the string at time $t, T(\zeta)>0$ is the Young's modulus of the string, $\rho(\zeta)>0$ is the mass density, and $\kappa>0$. We model this system as a portHamiltonian system. Therefore we introduce the state variable $x=\left[\begin{array}{l}x_{1}(\zeta, t) \\ x_{2}(\zeta, t)\end{array}\right]$ with $x_{1}=\rho(\zeta) \frac{\partial w}{\partial t}$ (momentum), $x_{2}=\frac{\partial w}{\partial \zeta}$ (strain) and the state space $X=$ $L^{2}\left(0,1 ; \mathbb{C}^{d}\right)$. Since the meaning of $w(0, t)=0$ is that in the point $\zeta=0$ the string is fixed for all times, we model this boundary condition as $\frac{\partial w}{\partial t} w(0, t)=0$. Then the closed-loop system (5.28) can equivalently be written as

$$
\begin{align*}
\frac{\partial}{\partial t}\left[\begin{array}{l}
x_{1}(\zeta, t) \\
x_{2}(\zeta, t)
\end{array}\right] & =P_{1} \frac{\partial}{\partial \zeta}\left(\mathcal{H}(\zeta)\left[\begin{array}{l}
x_{1}(\zeta, t) \\
x_{2}(\zeta, t)
\end{array}\right]\right), \\
0 & =\left[\begin{array}{llll}
0 & 0 & 1 & 0 \\
\kappa & 1 & 0 & 0
\end{array}\right]\left[\begin{array}{l}
\mathcal{H}(1) x(1, t) \\
\mathcal{H}(0) x(0, t)
\end{array}\right], \tag{5.29}
\end{align*}
$$

where $P_{1}=\left[\begin{array}{ll}0 & 0 \\ 0 & 1\end{array}\right], \mathcal{H}(\zeta)=\left[\begin{array}{cc}\frac{1}{\rho(\zeta)} & 0 \\ 0 & T(\zeta)\end{array}\right], \widetilde{W}_{1}=\left[\begin{array}{ll}0 & 0 \\ \kappa & 1\end{array}\right], \widetilde{W}_{0}=\left[\begin{array}{ll}1 & 0 \\ 0 & 0\end{array}\right]$ and $\kappa>0$. We define the corresponding port-Hamiltonian operator $A$

$$
\begin{aligned}
A x & :=P_{1} \frac{\partial}{\partial \zeta}(\mathcal{H} x), \\
\mathcal{D}(A) & =\left\{x \in X \mid \mathcal{H} x \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{2}\right) \text { and }\left[\widetilde{W}_{1} \quad \widetilde{W}_{0}\right]\left[\begin{array}{l}
(\mathcal{H} x)(1) \\
(\mathcal{H} x)(0)
\end{array}\right]=0\right\}
\end{aligned}
$$

and remind that due to Theorem 3.3.5 the system has a unique mild solution if $A$ generates a $C_{0}$-semigroup. To show that the port-Hamiltonian operator $A$ is a discrete Riesz spectral operator, it is sufficient due to Theorem 5.3.3 to prove that $A$ generates a $C_{0}$-group. So we have only to check that $\widetilde{W}_{1} \mathcal{H}(1) Z^{+}(1) \oplus$ $\widetilde{W}_{0} \mathcal{H}(0) Z^{-}(0)=\mathbb{C}^{d}$ and $\widetilde{W}_{1} \mathcal{H}(1) Z^{-}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{+}(0)=\mathbb{C}^{d}$, where $Z^{+}(\zeta)$ denotes the span of the eigenvectors of $P_{1} \mathcal{H}(\zeta)$ corresponding to the positive eigenvalues of $P_{1} \mathcal{H}(\zeta)$ and $Z^{-}(\zeta)$ is the span of the eigenvectors of $P_{1} \mathcal{H}(\zeta)$ corresponding to the negative eigenvalues of $P_{1} \mathcal{H}(\zeta)$. Defining $\gamma=\sqrt{T(\zeta) / \rho(\zeta)}$,
the matrix function $P_{1} \mathcal{H}$ can be factorized as

$$
P_{1} \mathcal{H}=\underbrace{\left[\begin{array}{cc}
\gamma & -\gamma \\
\rho^{-1} & \rho^{-1}
\end{array}\right]}_{S^{-1}} \underbrace{\left[\begin{array}{cc}
\gamma & 0 \\
0 & -\gamma
\end{array}\right]}_{\Delta} \underbrace{\left[\begin{array}{cc}
(2 \gamma)^{-1} & \rho / 2 \\
(2 \gamma)^{-1} & \rho / 2
\end{array}\right]}_{S} .
$$

It is easy to see that $Z^{+}(\zeta)=\operatorname{span}\left[\begin{array}{c}T(\zeta) \\ \gamma(\zeta)\end{array}\right]$ and $Z^{-}(\zeta)=\operatorname{span}\left[\begin{array}{c}-T(\zeta) \\ \gamma(\zeta)\end{array}\right]$. Then it holds for $\kappa \neq-\frac{T(1)}{\gamma(1)}$ and $\kappa \neq \frac{T(1)}{\gamma(1)}$

$$
\begin{aligned}
& \widetilde{W}_{1} \mathcal{H}(1) Z^{+}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{-}(0)=\left[\begin{array}{c}
0 \\
\kappa \gamma(1)+T(1)
\end{array}\right] \oplus\left[\begin{array}{c}
-\gamma(0) \\
0
\end{array}\right]=\mathbb{C}^{2}, \\
& \widetilde{W}_{1} \mathcal{H}(1) Z^{-}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{+}(0)=\left[\begin{array}{c}
0 \\
-\kappa \gamma(1)+T(1)
\end{array}\right] \oplus\left[\begin{array}{c}
\gamma(0) \\
0
\end{array}\right]=\mathbb{C}^{2} .
\end{aligned}
$$

Thus, $A$ is a discrete Riesz spectral operator for $\kappa \neq-\frac{T(1)}{\gamma(1)}$ and $\kappa \neq \frac{T(1)}{\gamma(1)}$. In contrast to [XW11] we do not need determine the eigenvalues exactly, which is only possible if $\rho$ and $T$ are constant. For spacial varying coefficients like $\rho(\zeta)=e^{\zeta}$ and $T(\zeta)=\zeta+1$ the problem is not analytically solvable, see JZ12, Exercise 12.1].

### 5.4.2 Timoshenko beam with boundary damping

In Example 3.1.8 and 3.2 .9 it is shown that the Timoshenko beam with boundary damping can be formulated as port-Hamiltonian system: It can be written in the form of (5.15), with

$$
P_{1}=\left[\begin{array}{llll}
0 & 1 & 0 & 0 \\
1 & 0 & 0 & 0 \\
0 & 0 & 0 & 1 \\
0 & 0 & 1 & 0
\end{array}\right], \mathcal{H}(\zeta)=\left[\begin{array}{cccc}
K(\zeta) & 0 & 0 & 0 \\
0 & \frac{1}{\rho(\zeta)} & 0 & 0 \\
0 & 0 & E I(\zeta) & 1 \\
0 & 0 & 0 & \frac{1}{I_{\rho}(\zeta)}
\end{array}\right]
$$

and

$$
P_{0}=\left[\begin{array}{cccc}
0 & 0 & 0 & -1 \\
0 & 0 & 0 & 0 \\
0 & 0 & 0 & 0 \\
1 & 0 & 0 & 0
\end{array}\right],
$$

where the $K(\zeta)$ denotes the shear modulus, $E I(\zeta)$ is the product of Young's modulus of elasticity and the moment of inertia of a cross section, $\rho(\zeta)$ is the mass per unit length and $I_{\rho}(\zeta)$ denotes the rotary moment of inertia of a cross section. All these physical parameters are positive and continuously differentiable functions of $\zeta$. To model the fact that the beam is clamped in at $\zeta=0$ and controlled at $\zeta=1$ by the force and moment feedback, we add the boundary condition

$$
\left[\begin{array}{l}
0 \\
0 \\
0 \\
0
\end{array}\right]=\left[\widetilde{W}_{1} \widetilde{W}_{0}\right]\left[\begin{array}{l}
(\mathcal{H} x)(1, t) \\
(\mathcal{H} x)(0, t)
\end{array}\right] \text { with }\left[\widetilde{W}_{1} \widetilde{W}_{0}\right]=\left[\begin{array}{cccccccc}
0 & 0 & 0 & 0 & 0 & 1 & 0 & 0 \\
0 & 0 & 0 & 0 & 0 \\
1 & \alpha_{1} & 0 & 0 & 0 & 0 & 0 & 1 \\
0 & 0 & 1 & \alpha_{2} & 0 & 0 & 0 & 0 \\
0 & 0
\end{array}\right]
$$

and $\alpha_{1}, \alpha_{2}$ are given positive gain feedback constants.

For shortness, we define the $\zeta$-depending functions $\gamma_{1}=\frac{1}{\sqrt{\rho(\zeta) K(\zeta)}}$ and $\gamma_{2}=$ $\frac{1}{\sqrt{I_{\rho}(\zeta) E I(\zeta)}}$. Then we diagonalize $P_{1} \mathcal{H}(\zeta)$ and it holds

$$
\begin{aligned}
P_{1} \mathcal{H}(\zeta) & =\left[\begin{array}{cccc}
0 & \frac{1}{\rho(\zeta)} & 0 & 0 \\
K(\zeta) & 0 & 0 & 0 \\
0 & 0 & 0 & \frac{1}{I_{\rho}(\zeta)} \\
0 & 0 & E I(\zeta) & 0
\end{array}\right] \\
& =\left[\begin{array}{cccc}
0 & \gamma_{1} & 0 & -\gamma_{1} \\
0 & 1 & 0 & 1 \\
\gamma_{2} & 0 & -\gamma_{2} & 0 \\
1 & 0 & 1 & 0
\end{array}\right]\left[\begin{array}{cccc}
\sqrt{\frac{E I(\zeta)}{I_{\rho}(\zeta)}} & 0 & 0 & 0 \\
0 & \sqrt{\frac{K(\zeta)}{\rho(\zeta)}} & 0 & 0 \\
0 & 0 & -\sqrt{\frac{E I(\zeta)}{I_{\rho}(\zeta)}} & 0 \\
0 & 0 & 0 & -\sqrt{\frac{K(\zeta)}{\rho(\zeta)}}
\end{array}\right]\left[\begin{array}{cccc}
0 & 0 & \frac{1}{2 \gamma_{2}} & \frac{1}{2} \\
\frac{1}{2 \gamma_{1}} & \frac{1}{2} & 0 & 0 \\
0 & 0 & -\frac{1}{2 \gamma_{2}} & \frac{1}{2} \\
-\frac{1}{2 \gamma_{1}} & \frac{1}{2} & 0 & 0
\end{array}\right] .
\end{aligned}
$$

Thus,

$$
\begin{aligned}
& Z^{+}(\zeta)=\operatorname{span}\left\{\left[\begin{array}{llll}
0 & 0 & \gamma_{2}(\zeta) & 1
\end{array}\right]^{\top},\left[\begin{array}{llll}
\gamma_{1}(\zeta) & 1 & 0 & 0
\end{array}\right]^{\top}\right\} \text { and } \\
& Z^{-}(\zeta)=\operatorname{span}\left\{\left[\begin{array}{llll}
0 & 0 & -\gamma_{2}(\zeta) & 1
\end{array}\right]^{\top},\left[\begin{array}{llll}
-\gamma_{1}(\zeta) & 1 & 0 & 0
\end{array}\right]^{\top}\right\} . \\
& \text { Since } \widetilde{W}_{1} \mathcal{H}(1)=\left[\begin{array}{cccc}
0 & 0 & 0 & 0 \\
0 & 0_{1} & 0 & 0 \\
K(1) & \frac{\alpha_{1}}{\rho(1)} & 0 & 0 \\
0 & 0 & E I(1) & \frac{\alpha_{2}}{I_{\rho}(1)}
\end{array}\right] \text { and } \widetilde{W}_{0} \mathcal{H}(0)=\left[\begin{array}{cccc}
0 & \frac{1}{\rho(0)} & 0 & 0 \\
0 & 0 & 0 & \frac{1}{I_{\rho}(0)} \\
0 & 0 & 0 \\
0 & 0 & 0 & 0
\end{array}\right] \text {, }
\end{aligned}
$$

we have

$$
\begin{aligned}
& \widetilde{W}_{1} \mathcal{H}(1) Z^{+}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{-}(0) \\
& =\operatorname{span}\left\{\left[\begin{array}{c}
0 \\
0 \\
0 \\
\frac{\gamma_{2}^{-1}(1)+\alpha_{2}}{I_{\rho}(1)}
\end{array}\right],\left[\begin{array}{c}
0 \\
\frac{\gamma_{1}^{-1}(1)+\alpha_{1}}{\rho(1)} \\
0
\end{array}\right],\left[\begin{array}{c}
0 \\
\frac{1}{I_{\rho}(0)} \\
0 \\
0
\end{array}\right],\left[\begin{array}{c}
\frac{1}{\rho(0)} \\
0 \\
0 \\
0
\end{array}\right]\right\}=\mathbb{C}^{4}
\end{aligned}
$$

and

$$
\begin{aligned}
& \widetilde{W}_{1} \mathcal{H}(1) Z^{-}(1) \oplus \widetilde{W}_{0} \mathcal{H}(0) Z^{+}(0)= \\
& \operatorname{span}\left\{\left[\begin{array}{c}
0 \\
0 \\
\frac{-\gamma_{2}^{-1}(1)+\alpha_{2}}{I \rho(1)}
\end{array}\right],\left[\begin{array}{c}
0 \\
\frac{-\gamma_{1}^{-1}(1)+\alpha_{1}}{\rho(1)} \\
0
\end{array}\right],\left[\begin{array}{c}
0 \\
\frac{1}{I \rho(0)} \\
0 \\
0
\end{array}\right],\left[\begin{array}{c}
\frac{1}{\rho(0)} \\
0 \\
0 \\
0
\end{array}\right]\right\}=\mathbb{C}^{4} .
\end{aligned}
$$

Recall that $W_{B}=\left[\begin{array}{ll}\widetilde{W}_{1} & \widetilde{W}_{0}\end{array}\right]\left[\begin{array}{cc}P_{1} & -P_{1} \\ I & I\end{array}\right]^{-1}$, cf. Equation 3.13. Since $W_{B} \Sigma W_{B} \neq 0, A$ does not generate a unitary $C_{0}$-group, but nevertheless $A$ generates a $C_{0}$-group and is by Theorem 5.3.3 a discrete Riesz spectral operator.
Xu and Feng dedicate the paper XF02 to this example and they proved under the extra assumption that all physical constants are independent of $\zeta$ that the eigenvectors and generalized eigenvectors of the operator form a Riesz basis. This example is also revisited in Vil07 using another approach. Using our main theorem, we can easy verify that the associated system operator is a discrete Riesz spectral operator.

### 5.5 Closing remarks and open problems

We have shown that a port-Hamiltonian system of the form (5.15) is a Riesz spectral operator if and only if it generates a $C_{0}$-group. Many (hyperbolic) systems can be written into this form, with as main exception the Euler-Bernoulli beam equation. Of course the basis property of this equation is well-studied, and many results are known, see e.g. GW19. However, we assert that the main result of this chapter does not hold for the Euler-Bernoulli beam equation.
In Theorem 5.3.3 we have shown that if the port-Hamiltonian systems (5.15) is a Riesz spectral operator, then the eigenvalues (counted according to the algebraic multiplicity) can be decomposed into finitely many sets each having a uniform gap. If we count the eigenvalues without multiplicity, then JZ01a, Theorem 2] shows that they can be decomposed into at most $d$ sets each having a uniform gap. We claim that this results holds true if we count the eigenvalues according to the algebraic multiplicity.
One may ask whether the main theorem of this chapter (Theorem 5.3.3) holds if we drop the assumption that $P_{0}$ is skew-symmetric. Our proof uses the fact that every well-posed port-Hamiltonian control system (5.15) is exactly controllable in finite time and this property is only known in the case that $P_{0}$ is skew-symmetric. Thus, if $P_{0}$ is an arbitrary $d \times d$-matrix, then our proof carries over to this more general case, provided we add the assumption that the corresponding port-Hamiltonian system is exactly controllable in finite-time. However, we assert that even when $P_{0}$ is not skew-symmetric, the system (5.15) is exactly controllable in finite time, and thus this extra assumption would not be needed.
Guo and Wang [GW19] studied the Riesz basis property for a closely related class of systems, that is, hyperbolic systems of the form $\frac{\partial x}{\partial t}=K(\zeta) \frac{\partial x}{\partial \zeta}+C(\zeta) x$ with $K$ and $C$ diagonal. They showed that for their class of systems the state space can be split into two parts, one part generated by a $C_{0}$-group and the other generated by an operator without spectrum, see [GW19, Theorem 4.10]. In particular, the spectrum of their operators always lies in a strip parallel to the imaginary axis. This result does not generalize to our system class as the following lemma shows.

Lemma 5.5.1. We consider the port-Hamiltonian system

$$
\begin{align*}
\frac{\partial}{\partial t}\left[\begin{array}{l}
x_{1}(\zeta, t) \\
x_{2}(\zeta, t)
\end{array}\right] & =\left[\begin{array}{ll}
1 & 0 \\
0 & \frac{1}{2}
\end{array}\right] \frac{\partial}{\partial \zeta}\left[\begin{array}{l}
x_{1}(\zeta, t) \\
x_{2}(\zeta, t)
\end{array}\right]+\left[\begin{array}{cc}
0 & 1 \\
-1 & 0
\end{array}\right]\left[\begin{array}{l}
x_{1}(\zeta, t) \\
x_{2}(\zeta, t)
\end{array}\right]  \tag{5.30}\\
0 & =\left[\begin{array}{llll}
1 & 0 & 0 & 1 \\
0 & 1 & 0 & 0
\end{array}\right]\left[\begin{array}{l}
x(1, t) \\
x(0, t)
\end{array}\right] \tag{5.31}
\end{align*}
$$

The system operator associated to (5.30-(5.31) generates a $C_{0}$-semigroup, but there exists a sequence of eigenvalues which real parts converge to $-\infty$.

Proof: By Theorem 3.2 .5 it is easy to see that the operator associated to (5.30)-(5.31) generates a $C_{0}$-semigroup. To show that there exists a sequence of eigenvalues which real parts converge to $-\infty$, we characterize the eigenvalues
of the port-Hamiltonian operator. If a complex number $s$ is an eigenvalue of the port-Hamiltonian operator associated to the system (5.30)-(5.31), then

$$
\begin{aligned}
& s x=\left[\begin{array}{ll}
1 & 0 \\
0 & \frac{1}{2}
\end{array}\right] x^{\prime}+\left[\begin{array}{cc}
0 & 1 \\
-1 & 0
\end{array}\right] x \\
& \Leftrightarrow x^{\prime}=s\left[\begin{array}{ll}
1 & 0 \\
0 & 2
\end{array}\right] x-\left[\begin{array}{cc}
1 & 0 \\
0 & 2
\end{array}\right]\left[\begin{array}{cc}
0 & 1 \\
-1 & 0
\end{array}\right] x \\
& \Leftrightarrow x^{\prime}=s\left[\begin{array}{ll}
1 & 0 \\
0 & 2
\end{array}\right] x-\left[\begin{array}{cc}
0 & 1 \\
-2 & 0
\end{array}\right] x \\
& \Leftrightarrow x^{\prime}=\left(s\left[\begin{array}{ll}
1 & 0 \\
0 & 2
\end{array}\right]+\left[\begin{array}{cc}
0 & -1 \\
2 & 0
\end{array}\right]\right) x .
\end{aligned}
$$

Due to Lemma 5.3.8, $s \in \mathbb{C}$ is an eigenvalue if and only if

$$
\operatorname{det}\left(\left[\begin{array}{ll}
1 & 0  \tag{5.32}\\
0 & 1
\end{array}\right] \Psi^{s}(1)+\left[\begin{array}{ll}
0 & 1 \\
0 & 0
\end{array}\right]\right)=0
$$

where

$$
\Psi^{s}(1)=\exp \left(s\left[\begin{array}{ll}
1 & 0 \\
0 & 2
\end{array}\right]+\left[\begin{array}{cc}
0 & -1 \\
2 & 0
\end{array}\right]\right)
$$

To obtain the eigenvalues of the matrix $s\left[\begin{array}{ll}1 & 0 \\ 0 & 2\end{array}\right]+\left[\begin{array}{cc}0 & -1 \\ 2 & 0\end{array}\right]=\left[\begin{array}{cc}s & -1 \\ 2 & 2 s\end{array}\right]$ we determine the zeros of

$$
\begin{align*}
\operatorname{det}\left(\left[\begin{array}{cc}
s-\mu & -1 \\
2 & 2 s-\mu
\end{array}\right]\right) & =(s-\mu)(2 s-\mu)+2  \tag{5.33}\\
& =\mu^{2}-3 s \mu+2\left(s^{2}+1\right) \tag{5.34}
\end{align*}
$$

Thus the determinant 5.33 has the zeros

$$
\begin{equation*}
\mu_{1,2}=\frac{3 s}{2} \pm \sqrt{\left(\frac{-3 s}{2}\right)^{2}-2\left(s^{2}+1\right)}=\frac{3 s \pm \sqrt{s^{2}-8}}{2} \tag{5.35}
\end{equation*}
$$

These are the eigenvalues of $\left[\begin{array}{cc}s & -1 \\ 2 & 2 s\end{array}\right]$ to the eigenvectors $v_{1}=\left[\begin{array}{c}1 \\ s-\mu_{1}\end{array}\right]$ and $v_{2}=$ $\left[{ }_{s}{ }^{1}-\mu_{2}\right]$. Note that $\mu_{1,2}$ are $s$-dependent. Thus, we get

$$
\exp \left(\left[\begin{array}{cc}
s & -1 \\
2 & 2 s
\end{array}\right]\right)=V(s)\left[\begin{array}{cc}
\mathrm{e}^{\mu_{1}} & 0 \\
0 & \mathrm{e}^{\mu_{2}}
\end{array}\right] V(s)^{-1},
$$

where $V(s)$ consists of the eigenvectors $v_{1}$ and $v_{2}$, i.e., $V(s)=\left[\begin{array}{cc}1 & 1 \\ s-\mu_{1} & s-\mu_{2}\end{array}\right]$.

Thus, 5.32 is equivalent to

$$
\begin{align*}
& \operatorname{det}\left(\left[\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right] V(s)\left[\begin{array}{cc}
\mathrm{e}^{\mu_{1}} & 0 \\
0 & \mathrm{e}^{\mu_{2}}
\end{array}\right] V(s)^{-1}+\left[\begin{array}{ll}
0 & 1 \\
0 & 0
\end{array}\right]\right)=0 \\
\Leftrightarrow & \operatorname{det}\left(\left(V(s)\left[\begin{array}{cc}
\mathrm{e}^{\mu_{1}} & 0 \\
0 & \mathrm{e}^{\mu_{2}}
\end{array}\right]+\left[\begin{array}{ll}
0 & 1 \\
0 & 0
\end{array}\right] V(s)\right) V(s)^{-1}\right)=0 \\
\Leftrightarrow & \operatorname{det}\left(V(s)\left[\begin{array}{cc}
\mathrm{e}^{\mu_{1}} & 0 \\
0 & \mathrm{e}^{\mu_{2}}
\end{array}\right]+\left[\begin{array}{ll}
0 & 1 \\
0 & 0
\end{array}\right] V(s)\right)=0 \\
\Leftrightarrow & \operatorname{det}\left(\left[\begin{array}{cc}
1 & 1 \\
s-\mu_{1} & s-\mu_{2}
\end{array}\right]\left[\begin{array}{cc}
\mathrm{e}^{\mu_{1}} & 0 \\
0 & \mathrm{e}^{\mu_{2}}
\end{array}\right]+\left[\begin{array}{ll}
0 & 1 \\
0 & 0
\end{array}\right]\left[\begin{array}{cc}
1 & 1 \\
s-\mu_{1} & s-\mu_{2}
\end{array}\right]\right)=0 \\
\Leftrightarrow & \operatorname{det}\left(\left[\begin{array}{cc}
\mathrm{e}^{\mu_{1}} & \mathrm{e}^{\mu_{2}} \\
\left(s-\mu_{1}\right) \mathrm{e}^{\mu_{1}} & \left(s-\mu_{2}\right) \mathrm{e}^{\mu_{2}}
\end{array}\right]+\left[\begin{array}{cc}
s-\mu_{1} & s-\mu_{2} \\
0 & 0
\end{array}\right]\right)=0 \\
\Leftrightarrow & \operatorname{det}\left[\begin{array}{c}
\mathrm{e}^{\mu_{1}}+s-\mu_{1} \\
\left(s-\mathrm{e}^{\mu_{2}}+s-\mu_{2}\right. \\
\left(s \mathrm{e}^{\mu_{1}}\right. \\
\left(s-\mu_{2}\right) \mathrm{e}^{\mu_{2}}
\end{array}\right]=0 \\
\Leftrightarrow & \left(\mathrm{e}^{\mu_{1}}+s-\mu_{1}\right)\left(s-\mu_{2}\right) \mathrm{e}^{\mu_{2}}-\left(\mathrm{e}^{\mu_{2}}+s-\mu_{2}\right)\left(s-\mu_{1}\right) \mathrm{e}^{\mu_{1}}=0 \\
\Leftrightarrow & \left(\mu_{1}-\mu_{2}\right) \mathrm{e}^{\mu_{2}+\mu_{1}}+\left(s-\mu_{1}\right)\left(s-\mu_{2}\right) \mathrm{e}^{\mu_{2}}-\left(s-\mu_{1}\right)\left(s-\mu_{2}\right) \mathrm{e}^{\mu_{1}}=0 . \tag{5.36}
\end{align*}
$$

By equation 5.33 and 5.35 the determinant of $\left[\begin{array}{cc}s-\mu & -1 \\ 2 & 2 s-\mu\end{array}\right]$ is described by a polynomial $p(\mu)=\left(\mu-\mu_{1}\right)\left(\mu-\mu_{2}\right)$. Evaluating $p(\mu)$ at $\mu=s$, we get, using again (5.33),

$$
\begin{equation*}
p(s)=\left(s-\mu_{1}\right)\left(s-\mu_{2}\right)=s^{2}-3 s^{2}+2 s^{2}+2=2 \tag{5.37}
\end{equation*}
$$

Using (5.37), equation (5.36) is equivalent to

$$
\begin{equation*}
\left(\mu_{1}-\mu_{2}\right) \mathrm{e}^{\mu_{2}+\mu_{1}}+2 \mathrm{e}^{\mu_{2}}-2 \mathrm{e}^{\mu_{1}}=0 \tag{5.38}
\end{equation*}
$$

Now, we consider the asymptotic behaviour of the zeros of 5.33 . Since it holds

$$
\sqrt{s^{2}-8}-\sqrt{s^{2}}=\frac{8}{s\left(\sqrt{1-\frac{8}{s^{2}}}+1\right)}
$$

it holds for $s=x+i y$

$$
\lim _{x \rightarrow-\infty} \frac{8}{(x+i y)\left(\sqrt{1-\frac{8}{(x+i y)^{2}}}+1\right)}=0
$$

In the following, the $o(1)$-notation is used for Re $s \rightarrow-\infty$. Thus, $\mu_{1}-2 s=o(1)$, i.e., $\lim _{\operatorname{Re} s \rightarrow-\infty} \mu_{1}-2 s=0$. Analogously, it holds $\mu_{2}-s=o(1)$. This implies

$$
\begin{aligned}
\mathrm{e}^{\mu_{1}} & =\mathrm{e}^{2 s+o(1)}=\mathrm{e}^{2 s} \mathrm{e}^{o(1)} \\
\mathrm{e}^{\mu_{2}} & =\mathrm{e}^{s+o(1)}=\mathrm{e}^{s} \mathrm{e}^{o(1)}, \\
\mu_{1}-s & =o(1)+s \\
\mu_{2}-\mu_{1} & =-s+o(1)
\end{aligned}
$$

Using equation 5.37) it holds

$$
\begin{aligned}
\mu_{2}-s & =\frac{2}{\mu_{1}-s}=\frac{2}{s+o(1)} \text { and } \\
\mu_{1}-\mu_{2} & =\left(\mu_{1}-s\right)-\left(\mu_{2}-s\right)=s+o(1)-\frac{2}{s+o(1)} .
\end{aligned}
$$

We aim to apply the Theorem of Rouché, c.f. BC63, Theorem 12.2].
For a closed contour $C$ in the left halfplane with Res very small, it holds

$$
\begin{aligned}
g_{\mu}(s) & =\mathrm{e}^{\mu_{1}}\left[\left(\mu_{1}-\mu_{2}\right) \mathrm{e}^{\mu_{2}}+2 \mathrm{e}^{\mu_{2}-\mu_{1}}-2\right] \\
& =\mathrm{e}^{2 s} \mathrm{e}^{o(1)}\left[\left(s+o(1)-\frac{2}{s+o(1)}\right) \mathrm{e}^{s} \mathrm{e}^{o(1)}+2 \mathrm{e}^{-s} \mathrm{e}^{o(1)}-2\right] .
\end{aligned}
$$

We have to show that the approximation of $g_{\mu}(s)$ by

$$
g(s)=\mathrm{e}^{2 s}\left[\left(s-\frac{2}{s}\right) \mathrm{e}^{s}+2 \mathrm{e}^{-s}-2\right]
$$

is good, i.e.,

$$
\frac{\left|g_{\mu}(s)-g(s)\right|}{|g(s)|}<1 .
$$

It holds for Re $s$ very small

$$
\begin{aligned}
& \frac{\left|g_{\mu}(s)-g(s)\right|}{|g(s)|} \\
& =\frac{\left|\mathrm{e}^{\mu_{1}}\left[\left(\mu_{1}-\mu_{2}\right) \mathrm{e}^{\mu_{2}}+2 \mathrm{e}^{\mu_{2}-\mu_{1}}-2\right]-\mathrm{e}^{2 s}\left[\left(s-\frac{2}{s}\right) \mathrm{e}^{s}+2 \mathrm{e}^{-s}-2\right]\right|}{\left|\mathrm{e}^{2 s}\left[\left(s-\frac{2}{s}\right) \mathrm{e}^{s}+2 \mathrm{e}^{-s}-2\right]\right|} \\
& =\frac{\left|\mathrm{e}^{2 s} \mathrm{e}^{o(1)}\left[\left(s+o(1)-\frac{2}{s+o(1)}\right) \mathrm{e}^{s} \mathrm{e}^{o(1)}+2 \mathrm{e}^{-s} \mathrm{e}^{o(1)}-2\right]-\mathrm{e}^{2 s}\left[\left(s-\frac{2}{s}\right) \mathrm{e}^{s}+2 \mathrm{e}^{-s}-2\right]\right|}{\left|\mathrm{e}^{2 s}\left[\left(s-\frac{2}{s}\right) \mathrm{e}^{s}+2 \mathrm{e}^{-s}-2\right]\right|} \\
& =\frac{\left|\left(s+o(1)-\frac{2}{s+o(1)}\right) \mathrm{e}^{o(1)} \mathrm{e}^{o(1)}+2 \mathrm{e}^{-2 s} \mathrm{e}^{o(1)}-2 \mathrm{e}^{-s} \mathrm{e}^{o(1)}-\left[\left(s-\frac{2}{s}\right)+2 \mathrm{e}^{-2 s}-2 \mathrm{e}^{-s}\right]\right|}{\left|\left(s-\frac{2}{s}\right)+2 \mathrm{e}^{-2 s}-2 \mathrm{e}^{-s}\right|} \\
& =\frac{\left|\left(s-\frac{2}{s}\right)(1+o(1))+2 \mathrm{e}^{-2 s}(1+o(1))-2 \mathrm{e}^{-s}(1+o(1))-\left(s-\frac{2}{s}\right)-2 \mathrm{e}^{-2 s}+2 \mathrm{e}^{-s}\right|}{\left|\left(s-\frac{2}{s}\right)+2 \mathrm{e}^{-2 s}-2 \mathrm{e}^{-s}\right|} \\
& =\frac{\left|\left(\left(s-\frac{2}{s}\right)+2 \mathrm{e}^{-2 s}-2 \mathrm{e}^{-s}\right)(1+o(1))-\left(s-\frac{2}{s}\right)-2 \mathrm{e}^{-2 s}+2 \mathrm{e}^{-s}\right|}{\left|\left(s-\frac{2}{s}\right)+2 \mathrm{e}^{-2 s}-2 \mathrm{e}^{-s}\right|} \\
& =\frac{\left|\left(\left(s-\frac{2}{s}\right)+2 \mathrm{e}^{-2 s}-2 \mathrm{e}^{-s}\right) o(1)\right|}{\left|\left(s-\frac{2}{s}\right)+2 \mathrm{e}^{-2 s}-2 \mathrm{e}^{-s}\right|} \\
& =o(1) .
\end{aligned}
$$

Thus, it is sufficient to get information about the zeros of

$$
\mathrm{e}^{2 s}\left[\left(s-\frac{2}{s}\right) \mathrm{e}^{s}+2 \mathrm{e}^{-s}-2\right] .
$$

To get some information about the asymptotic behaviour of its zeros we bring the polynomial of exponentials in the standard form introduced in BC63, Chapter 12]. It holds

$$
\begin{align*}
\mathrm{e}^{2 s}\left[\left(s-\frac{2}{s}\right) \mathrm{e}^{s}+2 \mathrm{e}^{-s}-2\right] & =\frac{1}{s} e^{s}\left[\left(s^{2}-2\right) \mathrm{e}^{2 s}-2 s \mathrm{e}^{s}+2 s\right] \\
& =\frac{1}{s} e^{s} \sum_{j=0}^{n} p_{j}(s) \mathrm{e}^{\beta_{j} s}, \quad 0=\beta_{0}<\beta_{1}<\cdots<\beta_{n} \tag{5.39}
\end{align*}
$$

Let $m_{j}$ denotes the degree of the polynomial in $p_{j}(s)$. We aim to draw the distributional diagram. Thus, we need the points with coordinates $\left(\beta_{j}, m_{j}\right)$, draw the upper boundary part of the convex hull of $\left(\beta_{j}, m_{j}\right)$ and get a polygonal graph. Then it is not possible that points lie above the polygonal graph, but some may lie below it. The points below the polygonal graph does not effect the asymptotics of the zeros, see [BC63, Chapter 12.8]. Thus, due to equation (5.39) we draw the points $(2,2),(0,1),(1,1)$ and get the following distributional diagram with only one line segment with slope $\frac{1}{2}$.


Figure 5.1: Distributional Diagram
Applying Theorem 12.10.d in BC63], we get that the zeros of (5.39) lie asymptotically along a curve $\left|s^{\frac{1}{2}} \mathrm{e}^{s}\right|=c$, where $c \in \mathbb{R}$ denotes a constant and the slope $\frac{1}{2}$ is taken into account. It holds

$$
\begin{aligned}
\left|s^{\frac{1}{2}} \mathrm{e}^{s}\right|=c & \left.\Leftrightarrow\left||s|^{\frac{1}{2}} \mathrm{e}^{\frac{1}{2} \cdot i(\arg (s)+2 k \pi)} \mathrm{e}^{s}\right||=c \Leftrightarrow| s\right|^{\frac{1}{2}} \mathrm{e}^{\operatorname{Re} s}=c \\
& \Leftrightarrow \operatorname{Re} s+\frac{1}{2} \ln (|s|)=\ln (c) \Leftrightarrow \operatorname{Re}\left(s+\frac{1}{2} \ln (|s|)+i \arg (s)\right)=\tilde{c} \\
& \Leftrightarrow \operatorname{Re}\left(s+\frac{1}{2} \ln s\right)=\tilde{c},
\end{aligned}
$$

where $\tilde{c}:=\ln (c)$. We define $s=x+i y$ and by Lemma 12.3 in BC63 we get that the curve is asymptotic to the curve $x+\frac{1}{2} \ln (|y|)=\tilde{c}$. Then the zeros lies along a curve with $x=\tilde{c}-\frac{1}{2} \ln (|y|)$ and thus, there exists a sequence of eigenvalues which real parts converge to $-\infty$.

The following example shows that the equivalence 1) $\Leftrightarrow 2$ ) in Theorem 5.3.3 does not hold for generators $A$ of $C_{0}$-semigroups $(T(t))_{t \geqslant 0}$ on Hilbert spaces even if we additionally assume that there exists a admissible control operator $B \in \mathcal{L}\left(\mathbb{C}^{d}, X_{-1}\right)$ for $(T(t))_{t \geqslant 0}$ such that the control system $\dot{x}(t)=A x(t)+B u(t)$ is exactly controllable in finite time.
Example 5.5.2. Let $A: \mathcal{D}(A) \subset \ell^{2} \rightarrow \ell^{2}$ be defined by $(A x)_{n}=\left(s_{n} x_{n}\right)_{n}$, $\left(s_{n}\right)_{n \in \mathbb{N}}=\left(-2^{n}\right)_{n \in \mathbb{N}}$, and $\mathcal{D}(A)=\left\{\left.x \in \ell^{2}(\mathbb{N})\left|\sum_{n \in \mathbb{N}}\left(1+\left|s_{n}\right|^{2}\right)\right| x_{n}\right|^{2}<\infty\right\}$. Clearly, $A$ is a discrete Riesz spectral operator, generates a $C_{0}$-semigroup, but not a $C_{0}$-group. Here

$$
X_{-1}=\ell_{-1}^{2}=\left\{\left(x_{n}\right)_{n \in \mathbb{N}} \left\lvert\, \sum_{n \in \mathbb{N}} \frac{\left|x_{n}\right|^{2}}{\left(1+\left|s_{n}\right|^{2}\right)}<\infty\right.\right\} .
$$

Hence, we can identify $B \in \mathcal{L}\left(\mathbb{C}, \ell_{-1}^{2}\right)$ with a sequence $\left(b_{n}\right) \in \ell_{-1}^{2}$. Let $\left(b_{n}\right)_{n \in \mathbb{N}}=$ $\left(\sqrt{2^{n}}\right)_{n \in \mathbb{N}}$. Then it holds $\left(b_{n}\right)_{n \in \mathbb{N}} \in \ell_{-1}^{2}$, since

$$
\begin{equation*}
\sum_{n \in \mathbb{N}} \frac{\left(\sqrt{2^{n}}\right)^{2}}{1+\left(2^{n}\right)^{2}}=\sum_{n \in \mathbb{N}} \frac{2^{n}}{1+2^{2 n}}<\sum_{n \in \mathbb{N}} \frac{2^{n}}{2^{2 n}}<\infty . \tag{5.40}
\end{equation*}
$$

To proof the admissibility of $B$ we use the Carleson measure criterion by Weiss, Wei88]. Since the eigenvalues $s_{n}=-2^{n}$ are real, we only have to check that there exists a constant $M>0$ independent of $h$ such that

$$
\begin{equation*}
\sum_{-s_{n} \in R(h, 0)}\left|b_{n}\right|^{2} \leqslant M h \text { for any } h>0, \tag{5.41}
\end{equation*}
$$

where $R(h, 0):=\{z \in \mathbb{C} \mid 0 \leqslant \operatorname{Re} z \leqslant h$ and $|\operatorname{Im} z| \leqslant h\}$ denotes a rectangle in the right complex half plane at the imaginary axis. It holds

$$
\begin{equation*}
\sum_{-s_{n} \in R(h, 0)}\left|\sqrt{2^{n}}\right|^{2}=\sum_{n=1}^{k} 2^{n} \leqslant 2 \cdot 2^{k}, \tag{5.42}
\end{equation*}
$$

where $k:=\max _{i} h-2^{i}>0$. To verify that $(A, B)$ is exactly controllable in finite time, we use one of the dual equivalences in [JZ01b, Theorem 2] for onedimensional output operators and formulate the assertion for diagonal systems with one-dimensional input operators.

Theorem 5.5.3. The diagonal system $(A, B)$ is exactly controllable in infinite time if and only if the following two conditions hold:

1. The eigenvalues $s_{n}$ are properly spaced, i.e.,

$$
\begin{equation*}
\inf _{n \neq m}\left|\frac{s_{n}-s_{m}}{\operatorname{Re} s_{n}}\right|>0 \text { and } \tag{5.43}
\end{equation*}
$$

2. there exists a constant $C>0$ such that $C \mid$ Res $_{m}\left|\leqslant\left|b_{m}\right|^{2}\right.$.

Here, we use the above theorem for an exponential stable semigroup. Thus, exact controllability in infinite time is equivalent to exact controllability in finite time. It holds

$$
\begin{equation*}
\left|\frac{2^{n}-2^{m}}{2^{n}}\right| \geqslant\left|\frac{2^{n}-2^{m-1}}{2^{n}}\right|=\left|1-2^{-1}\right|>0 \tag{5.44}
\end{equation*}
$$

Thus, we see that the sequence of eigenvalues is properly spaced and it holds $\left|\operatorname{Re} s_{m}\right| \leqslant\left|b_{m}\right|^{2}$ as well and therefore, exact controllability in finite time follows.

## Chapter 6

## Generalization of port-Hamiltonian systems

So far, we have only considered port-Hamiltonian systems of order 1. In the following section we consider port-Hamiltonian systems of order $N$. Then not only the Timoschenko beam but also the Euler-Bernoulli beam can be modelled as a port-Hamiltonian system, namely a port-Hamiltonian system of order 2. Port-Hamiltonian systems of $N$-th order on a bounded interval are well-studied, see for example (Vil07], LGZM05], [AJ14] and Aug16].
We consider the well-posedness of a class of hyperbolic partial differential equations on a one dimensional spatial domain, i.e., whether the associated operator generates a contraction $C_{0}$-semigroup. This class includes coupled wave and beam equations and in particular infinite networks of these equations, that means networks with an infinite number of edges.


Figure 6.1: Arbitrary infinitedimensional network


Figure 6.2: Infinite-dimensional tree

In this chapter equivalent conditions for contraction $C_{0}$-semigroup generation are derived. We consider these equations on a finite interval as well as on a semi-axis. In particular, contraction $C_{0}$-semigroup generation has been studied in Chapter 3 and LGZM05, JZ12, AJ14, Aug16, JMZ15. In this chapter we aim to generalize these results to the infinite-dimensional situation and to the semi-axis.
The results of this chapter are published in (JK19b].

### 6.1 Port-Hamiltonian systems in the infinite-dimensional setting

We consider on the interval $[0,1]$ a system of partial differential equations of the form

$$
\begin{align*}
\frac{\partial x}{\partial t}(\zeta, t) & =\left(\sum_{k=0}^{N} P_{k} \frac{\partial^{k}}{\partial \zeta^{k}}\right)(\mathcal{H}(\zeta) x(\zeta, t)), \quad \zeta \in(0,1), t \geqslant 0,  \tag{6.1}\\
x(\zeta, 0) & =x_{0}(\zeta)
\end{align*}
$$

where $P_{N}$ is an invertible operator on a Hilbert space $H$ and $P_{k} \in \mathcal{L}(H)$, $k=0, \cdots, N$, with $P_{k}^{*}=(-1)^{k+1} P_{k}, k=1, \cdots, N$. Here $\mathcal{L}(H)$ denotes the set of linear bounded operators on $H . \mathcal{H}(\zeta)$ is a positive operator on $H$ for a.e. $\zeta \in(0,1)$ satisfying $\mathcal{H}, \mathcal{H}^{-1} \in L^{\infty}(0,1 ; \mathcal{L}(H))$.

To give an example of a port-Hamiltonian system of order 2 we consider the Euler-Bernoulli beam.
The equation of the Euler-Bernoulli beam models the transversal vibration of an elastic beam where the cross section of the beam is also vertical to the neutral axis after the bending. An extension of the Euler-Bernoulli beam is the Timoshenko beam model, which takes shear and rotational inertia effects into account, see Example 3.1.8.
Example 6.1.1. The Euler-Bernoulli beam is described by the partial differential equation

$$
\begin{equation*}
\rho(\zeta) \frac{\partial^{2} \omega}{\partial t^{2}}(\zeta, t)+\frac{\partial^{2}}{\partial \zeta^{2}}\left(E I(\zeta) \frac{\partial^{2} \omega}{\partial \zeta^{2}}(\zeta, t)\right)=0, \quad t \geqslant 0 \tag{6.2}
\end{equation*}
$$

where $\omega(\zeta, t)$ describes the transverse displacement of the beam. All physical parameters are positive and continuously differentiable functions of $\zeta$. Here $\rho(\zeta)$ denotes the mass per unit length and $E I(\zeta)$ is the product of Young's modulus of elasticity and the moment of inertia of the cross section. Using the state variables

$$
\begin{aligned}
x_{1}(\zeta, t) & :=\rho(\zeta) \frac{\partial \omega}{\partial t}(\zeta, t) \\
x_{2}(\zeta, t) & :=\frac{\partial^{2} \omega}{\partial \zeta^{2}}(\zeta, t)
\end{aligned}
$$

we model the Euler-Bernoulli beam equation (6.2) as a port-Hamiltonian system of order 2 .

$$
\frac{\partial}{\partial t}\left[\begin{array}{l}
x_{1}(\zeta, t)  \tag{6.3}\\
x_{2}(\zeta, t)
\end{array}\right]=\left[\begin{array}{cc}
0 & -1 \\
1 & 0
\end{array}\right] \frac{\partial^{2}}{\partial \zeta^{2}}\left(\left[\begin{array}{cc}
\frac{1}{\rho(\zeta)} & 0 \\
0 & E I(\zeta)
\end{array}\right]\left[\begin{array}{l}
x_{1}(\zeta, t) \\
x_{2}(\zeta, t)
\end{array}\right]\right)
$$

There are also examples of port-Hamiltonian systems of higher order. An example for a port-Hamiltonian system of order 3 is the Airy equation, which is described in (MNS18].
Example 6.1.2. The Airy equation is the linear part of the Korteweg-de Vries equation, which describes waves on shallow water. The Airy equation on a one-dimensional spacial domain is given by

$$
\begin{equation*}
\frac{\partial \omega}{\partial t}(t, \zeta)+\frac{\partial^{3} \omega}{\partial \zeta^{3}}(\zeta, t)=0, t \geqslant 0, \zeta \in(0,1) . \tag{6.4}
\end{equation*}
$$

This equation can be written as a port-Hamiltonian system of order 3 with $\mathcal{H}=1, P_{0}=P_{1}=P_{2}=0$, and $P_{3}=-1$.
In order to guarantee unique solutions of equation (6.1), we have to impose boundary conditions, which will be of the form

$$
\begin{equation*}
\widetilde{W}_{B}(\Phi(\mathcal{H} x))(\cdot, t)=0 . \tag{6.5}
\end{equation*}
$$

We assume $\widetilde{W}_{B} \in \mathcal{L}\left(H^{2 N}, H^{N}\right)$ and that the operator $\Phi$ is given by

$$
\Phi: \mathcal{W}^{N, 2}(0,1 ; H) \rightarrow H^{2 N}, \quad \Phi(x):=\left[\Phi_{1}(x) \Phi_{0}(x)\right]^{T}
$$

where $\Phi_{i}(x):=\left[x(i) \ldots \frac{d^{N-1} x}{d \zeta^{N-1}}(i)\right]^{T}$ for $i \in\{0,1\}$ and $\mathcal{W}^{N, 2}(0,1 ; H)$ denotes the Sobolev space of order $N$, cf. Definition 2.0.4. Clearly, whether or not equation (6.1) possesses unique and non-increasing solutions depend on the boundary conditions, or equivalently on the operator $\widetilde{W}_{B}$. The partial differential equation (6.1) with the boundary conditions (6.5) can be equivalently written as an abstract Cauchy problem

$$
\begin{aligned}
\dot{x}(t) & =A x(t), \\
x(0) & =x_{0},
\end{aligned}
$$

where $A$ is the linear operator on the Hilbert space $X:=L^{2}(0,1 ; H)$ given by

$$
\begin{gather*}
A x:=\sum_{k=0}^{N} P_{k} \frac{\partial^{k}}{\partial \zeta^{k}}(\mathcal{H} x), \quad x \in \mathcal{D}(A),  \tag{6.6}\\
\mathcal{D}(A)=\left\{x \in X \mid \mathcal{H} x \in \mathcal{W}^{N, 2}(0,1 ; H) \text { and } \widetilde{W}_{B} \Phi(\mathcal{H} x)=0\right\} . \tag{6.7}
\end{gather*}
$$

We denote $A$ as port-Hamiltonian operator of order $N$. Again, as in Chapter 3. we equip $X$ not with the standard scalar product of $L^{2}(0,1 ; H)$ but with the inner product $\langle f, \mathcal{H} g\rangle$.
We define

$$
Q=\left(Q_{i j}\right)_{\substack{1 \leqslant i \leqslant N  \tag{6.8}\\ 1 \leqslant j \leqslant N}}= \begin{cases}(-1)^{i-1} P_{i+j-1} & \text { if } i+j \leqslant N+1 \\ 0 & \text { else. }\end{cases}
$$

Clearly, $Q_{i j} \in \mathcal{L}(H)$, i.e. $Q \in \mathcal{L}\left(H^{N}\right)$ and

$$
Q=\left[\begin{array}{cccccc}
P_{1} & P_{2} & P_{3} & \cdots & P_{N-1} & P_{N} \\
-P_{2} & -P_{3} & -P_{4} & \cdots & -P_{N} & 0 \\
P_{3} & P_{4} & . \cdot & . & 0 & 0 \\
-P_{4} & . \cdot & . \cdot & . & & \vdots \\
\vdots & . & . & & & \vdots \\
(-1)^{N-1} P_{N} & 0 & \cdots & \cdots & \cdots & 0
\end{array}\right] .
$$

Thus, $Q \in \mathcal{L}\left(H^{N}\right)$ is a selfadjoint block operator matrix and invertible due to the fact that $P_{N}$ is invertible. Let

$$
W_{B}:=\left[\begin{array}{ll}
W_{1} & W_{0}
\end{array}\right]:=\widetilde{W}_{B}\left[\begin{array}{cc}
Q & -Q \\
I & I
\end{array}\right]^{-1} \quad \text { and } \quad \Sigma:=\left[\begin{array}{ll}
0 & I \\
I & 0
\end{array}\right] \in \mathcal{L}\left(H^{N} \times H^{N}\right),
$$

where $W_{1}, W_{0} \in \mathcal{L}\left(H^{N}\right)$. Let $P \in \mathcal{L}(H)$. We call $P$ negative semi-definite, in short $P \leqslant 0$, if $\langle x, P x\rangle_{H} \leqslant 0$ for all $x \in H$. We define $\operatorname{Re} P=\frac{1}{2}\left(P+P^{*}\right)$ and $\operatorname{Im} P=\frac{1}{2 i}\left(P-P^{*}\right)$. Thus, $P=\operatorname{Re} P+i \operatorname{Im} P$ and $\operatorname{Re} P \leqslant 0$ if and only if $\langle x, \operatorname{Re} P x\rangle_{H}=\operatorname{Re}\langle x, P x\rangle_{H} \leqslant 0$.
The aim of this section is to give equivalent conditions for the fact that $A$ generates a contraction $C_{0}$-semigroup on $X$. Under a weak condition, we show that $A \mathcal{H}$ generates a contraction $C_{0}$-semigroup if and only if the operator $A$ is dissipative. Moreover, equivalent conditions in terms of the operator $\widetilde{W}_{B}$ are presented. We note that the mentioned weak condition is in particular satisfied if the Hilbert space $H$ is finite-dimensional. However, even if $H$ is finitedimensional, our result contains new equivalent conditions for the contraction $C_{0}$-semigroup characterization in Vil07, LGZM05 and AJ14].
Thus, we consider the operator $A$ on the Hilbert space $X=L^{2}(0,1 ; H)$, where $H$ is a (possibly infinite-dimensional) Hilbert space.
We start to collect all assertions, before we introduce some technical definitions and lemmas, and give the proofs of the following theorems and corollaries at the end of this section.

Theorem 6.1.3. Let $A$ be given by (3.9)-(3.10). Further, assume

$$
\begin{equation*}
\operatorname{ran}\left(W_{1}-W_{0}\right) \subseteq \operatorname{ran}\left(W_{1}+W_{0}\right) \tag{6.9}
\end{equation*}
$$

Then the following statements are equivalent:

1. The operator $A$ generates a contraction $C_{0}$-semigroup on $X$;
2. $A$ is dissipative, that is, $\operatorname{Re}\langle A x, x\rangle \leqslant 0$ for every $x \in \mathcal{D}(A)$;
3. $\operatorname{Re} P_{0} \leqslant 0, W_{1}+W_{0}$ is injective and $W_{B} \Sigma W_{B}^{*} \geqslant 0$;
4. $\operatorname{Re} P_{0} \leqslant 0, W_{1}+W_{0}$ is injective and there exists $V \in \mathcal{L}(H)$ with $\|V\| \leqslant 1$ such that $W_{B}=\frac{1}{2}\left(W_{1}+W_{0}\right)[I+V \quad I-V]$;
5. Re $P_{0} \leqslant 0$ and $u^{*} Q u-y^{*} Q y \leqslant 0$ for every $\left[\begin{array}{l}u \\ y\end{array}\right] \in \operatorname{ker} \widetilde{W}_{B}$.

Remark 6.1.4. 1. Condition (6.9) is in general not satisfied: Let $N=1$, $H=\ell^{2}$ and $W_{B}=\left[\begin{array}{ll}W_{1} & W_{0}\end{array}\right] \in \mathcal{L}\left(\ell^{2} \times \ell^{2}, \ell^{2}\right)$ with $W_{1} e_{i}:=e_{i+1}+e_{i}$ and $W_{0} e_{i}:=e_{i+1}-e_{i}$, where $\left\{e_{i}\right\}_{i \in \mathbb{N}}$ is an orthonormal basis of $\ell^{2}$. Then $\operatorname{ran}\left(W_{1}-W_{0}\right)=\ell^{2}$ whereas $e_{1} \notin \operatorname{ran}\left(W_{1}+W_{0}\right)$.
2. We point out that the implications $1 \Rightarrow 2,4 \Rightarrow 3$, and the equivalence $2 \Leftrightarrow$ 5 hold even without the additional condition (6.9). Moreover, condition (6.9) is not needed for the fact that 2 implies the injectivity of $W_{1}+W_{0}$.
3. We note that $W_{B}$ is not uniquely determined, only the kernel of $W_{B}$ is. However, if $W_{B}$ does not satisfy condition (6.9), then in general it is not possible to choose another operator instead of $W_{B}$ with the same kernel such that condition $\sqrt{6.9}$ ) holds.
4. If $H$ is finite-dimensional, then $A$ has a compact resolvent, see AJ14, Theorem 2.3]. However, in general, $A$ does not have a compact resolvent. Take for example $N=1, P_{1}=1, P_{0}=0, H=\ell^{2}, \widetilde{W}_{B}=\left[\begin{array}{ll}I & L\end{array}\right]$ and $\mathcal{H}(\zeta)=I_{\ell^{2}}$. Here $L$ denotes the left shift on $H$, that is, $L e_{i}=e_{i+1}$. Thus, $A$ generates the left shift semigroup on $X=L^{2}\left(0,1 ; \ell^{2}\right)$, which is isometric isomorph to the left shift on $X=L^{2}(0, \infty ; \mathbb{C})$. However, 0 is a spectral point of $A$, but not in the point spectrum.

As a corollary of Theorem 6.1.3 we obtain the well-known contraction $C_{0-}$ semigroup characterization for the case of a finite-dimensional Hilbert space $H$, see AJ14]. However, we remark that Conditions 3 and 4 are new even in the finite-dimensional situation.

Corollary 6.1.5. Let $A$ be given by (3.9)-(3.10) and assume that $H$ is finitedimensional. Then, assertions 1 to 5 in Theorem 6.1.3 are equivalent, and, moreover, they are equivalent to
6. $\operatorname{Re} P_{0} \leqslant 0, W_{B}$ surjective and $W_{B} \Sigma W_{B}^{*} \geqslant 0$;
7. $\operatorname{Re} P_{0} \leqslant 0, W_{B}$ surjective and there exists $V \in \mathcal{L}(H)$ with $\|V\| \leqslant 1$ such that $W_{B}=\frac{1}{2}\left(W_{1}+W_{0}\right)\left[\begin{array}{ll}I+V & I-V\end{array}\right]$.
Remark 6.1.6. If $H$ is infinite-dimensional, then in general Conditions 6 and 7 of the previous corollary are not equivalent to the fact that $A$ generates a contraction $C_{0}$-semigroup. In the following we give two counterexamples.
Let $H=\ell^{2}(\mathbb{N}), N \in \mathbb{N}$, and $P_{i}$ and $\mathcal{H}$ are operators satisfying the general assumptions. First, we consider $W_{B}=\left[\begin{array}{ll}W_{1} & W_{0}\end{array}\right]$ with $W_{1}:=\frac{3}{2} R$ and $W_{0}:=$ $\frac{1}{2} R$, where $R$ denotes the right shift on $\ell^{2}(\mathbb{N})$. Then $\operatorname{ran}\left(W_{1}-W_{0}\right)=\operatorname{ran}\left(W_{1}+\right.$ $\left.W_{0}\right), W_{1}+W_{0}$ is injective and $W_{B} \Sigma W_{B}^{*} \geqslant 0$ but $W_{B}$ is not surjective. Thus, $A$ generates a contraction $C_{0}$-semigroup on $X$, but Conditions 6 and 7 are not satisfied. Conversely, for the choice $W_{B}=\left[\begin{array}{ll}I-L & -I-L\end{array}\right]$, where $L$ denotes the left shift on $\ell^{2}(\mathbb{N})$, surjectivity of $W_{B}$ holds, $\operatorname{ran}\left(W_{1}-W_{0}\right) \subseteq \operatorname{ran}\left(W_{1}+W_{0}\right)$ and $W_{B} \Sigma W_{B}^{*} \geqslant 0$, but $W_{1}+W_{0}$ is not injective. Thus, for these boundary conditions the Conditions 6 and 7 of the previous corollary are satisfied, but $A$ does not generate a contraction $C_{0}$-semigroup on $X$.
Next, we characterize the property of unitary group generation of $A$.
Theorem 6.1.7. Let $A$ be given by (3.9)-(3.10). Further assume

$$
\begin{equation*}
\operatorname{ran}\left(W_{1}-W_{0}\right)=\operatorname{ran}\left(W_{1}+W_{0}\right) \tag{6.10}
\end{equation*}
$$

Then the following statements are equivalent:

1. A generates a unitary $C_{0}$-group on $X$;
2. $\operatorname{Re}\langle A x, x\rangle=0$ for every $x \in \mathcal{D}(A)$;
3. $\operatorname{Re} P_{0}=0, W_{1}+W_{0}$ and $-W_{1}+W_{0}$ are injective and $W_{B} \Sigma W_{B}^{*}=0$;
4. $\operatorname{Re} P_{0}=0, W_{1}+W_{0}$ and $-W_{1}+W_{0}$ are injective and there exists $V \in \mathcal{L}(H)$ with $\|V\|=1$ such that $W_{B}=\frac{1}{2}\left(W_{1}+W_{0}\right)[I+V \quad I-V]$;
5. Re $P_{0}=0$ and $u^{*} Q u-y^{*} Q y=0$ for every $\binom{u}{v} \in \operatorname{ker} \widetilde{W}_{B}$.

Corollary 6.1.8. Let $A$ be given by (3.9)-(3.10) and assume that $H$ is finitedimensional. Then, assertions 1 to 5 in Theorem 6.1.7 are equivalent, and, moreover, they are equivalent to
6. $\operatorname{Re} P_{0}=0, W_{B}$ surjective and $W_{B} \Sigma W_{B}^{*}=0$;
7. $\operatorname{Re} P_{0}=0, W_{B}$ surjective and there exists $V \in \mathcal{L}(H)$ with $\|V\|=1$ such that $W_{B}=\frac{1}{2}\left(W_{1}+W_{2}\right)\left[\begin{array}{ll}I+V & I-V\end{array}\right]$.

In order to prove these statements it is convenient to introduce the following linear combinations of the boundary values [LGZM05, which is a generalization of Definition 3.2.2.

Definition 6.1.9. For $x \in \mathcal{H}^{-1} \mathcal{W}^{N, 2}(0,1 ; H)$ we define so called boundary port variables, namely boundary flow and boundary effort, by

$$
\left[\begin{array}{l}
f_{\partial, \mathcal{H} x}  \tag{6.11}\\
e_{\partial, \mathcal{H} x}
\end{array}\right]:=\frac{1}{\sqrt{2}}\left[\begin{array}{cc}
Q & -Q \\
I & I
\end{array}\right] \Phi(\mathcal{H} x)=R_{e x t} \Phi(\mathcal{H} x),
$$

where $Q$ is defined by (6.8) and $R_{\text {ext }}:=\frac{1}{\sqrt{2}}\left[\begin{array}{cc}Q & -Q \\ I & I\end{array}\right] \in \mathcal{L}\left(H^{2 N}\right)$.
Remark 6.1.10. Thanks to the invertibility of $Q$, the operator $R_{\text {ext }}$ is invertible. As well as in Chapter 3 we use the boundary port variables to reformulate the domain of the operator $A$ :

$$
\begin{aligned}
\mathcal{D}(A) & =\left\{x \in X \mid \mathcal{H} x \in \mathcal{W}^{N, 2}(0,1 ; H) \text { and } \widetilde{W}_{B} \Phi(\mathcal{H} x)=0\right\} \\
& =\left\{x \in X \mid \mathcal{H} x \in \mathcal{W}^{N, 2}(0,1 ; H) \text { and } W_{B}\left[\begin{array}{l}
f_{\partial, \mathcal{H} x} \\
e_{\partial, \mathcal{H} x}
\end{array}\right]=0\right\}
\end{aligned}
$$

where $W_{B}=\widetilde{W}_{B} R_{e x t}^{-1}$.
Next, we determine the adjoint operator of $A$. We define $\widetilde{Q}=-Q$ and

$$
\left[\begin{array}{l}
\tilde{f}_{2, \mathcal{H} x} \\
\tilde{e}_{\partial, \mathcal{H} x}
\end{array}\right]=\widetilde{R}_{e x t} \Phi(\mathcal{H} x) \text { with } \widetilde{R}_{e x t}=\frac{1}{\sqrt{2}}\left[\begin{array}{cc}
\widetilde{Q} & -\widetilde{Q} \\
I & I
\end{array}\right] .
$$

Lemma 6.1.11. The adjoint operator of the operator $A$ defined in (6.6) with domain (6.7) and a boundary operator $W_{B}$ of the form $W_{B}=S[I+V \quad I-V]$ where $S, V \in \mathcal{L}\left(H^{N}\right)$ and $S$ is injective, is given by

$$
\begin{align*}
A^{*} y & =P_{0}^{*} y-\sum_{k=1}^{N} P_{k} \frac{d^{k}}{d \zeta^{k}} y, \quad y \in \mathcal{D}\left(A^{*}\right),  \tag{6.12}\\
\mathcal{D}\left(A^{*}\right) & =\left\{y \in \mathcal{W}^{N, 2}(0,1 ; H): S\left[I+V^{*} \quad I-V^{*}\right]\left[\begin{array}{l}
\tilde{f}_{\partial, y} \\
\tilde{e}_{\partial, y}
\end{array}\right]=0\right\} . \tag{6.13}
\end{align*}
$$

Proof: The statement can be proved in a similar manner as Proposition 3.4.3 in Aug16], where the statement is shown for finite-dimensional Hilbert spaces $H$.

Definition 6.1.12. We define the operators $A_{0}: \mathcal{D}\left(A_{0}\right) \subseteq X \rightarrow X$ and $\left(A^{*}\right)_{0}:$ $\mathcal{D}\left(\left(A^{*}\right)_{0}\right) \subseteq X \rightarrow X$ by

$$
\begin{aligned}
A_{0} x & :=\sum_{k=0}^{N} P_{k} \frac{\partial^{k}}{\partial \zeta^{k}} x, \quad\left(A^{*}\right)_{0} y:=P_{0}^{*} y-\sum_{k=1}^{N} P_{k} \frac{d^{k}}{d \zeta^{k}} y \\
\mathcal{D}\left(A_{0}\right) & =\mathcal{D}\left(A_{0}^{*}\right)=\mathcal{W}^{N, 2}(0,1 ; H) .
\end{aligned}
$$

Remark, that $A_{0}$ and $\left(A^{*}\right)_{0}$ are extensions of $A$ and $A^{*}$, respectively. Integration by parts yields the following lemma.

Lemma 6.1.13. We have for $x \in \mathcal{W}^{N, 2}(0,1 ; H)$

$$
\begin{aligned}
\operatorname{Re}\left\langle A_{0} x, x\right\rangle & =\operatorname{Re}\left\langle f_{\partial, x}, e_{\partial, x}\right\rangle_{H^{N}}+\operatorname{Re}\left\langle P_{0} x, x\right\rangle \\
& =\Phi_{1}(x)^{*} Q \Phi_{1}(x)-\Phi_{0}(x)^{*} Q \Phi_{0}(x)+\operatorname{Re}\left\langle P_{0} x, x\right\rangle, \\
\operatorname{Re}\left\langle\left(A^{*}\right)_{0} x, x\right\rangle & =\operatorname{Re}\left\langle\tilde{f}_{\partial, x}, \tilde{e}_{\partial, x}\right\rangle_{H^{N}}+\operatorname{Re}\left\langle P_{0} x, x\right\rangle \\
& =\Phi_{1}(x)^{*} \widetilde{Q} \Phi_{1}(x)-\Phi_{0}(x)^{*} \widetilde{Q} \Phi_{0}(x)+\operatorname{Re}\left\langle P_{0} x, x\right\rangle .
\end{aligned}
$$

Furthermore, we need some technical results. First, we give a generalization of the technical Lemma 7.3.2 in JZ12] for $N \geqslant 1$ and arbitrary Banach spaces $Z$.

Lemma 6.1.14. Let $Z$ be a Banach space and $V \in \mathcal{L}(Z)$. Then it holds

$$
\operatorname{ker}\left[\begin{array}{ll}
I+V & I-V
\end{array}\right]=\operatorname{ran}\left[\begin{array}{c}
I-V \\
-I-V
\end{array}\right]
$$

where $\left[\begin{array}{ll}I+V & I-V\end{array}\right] \in \mathcal{L}(Z \times Z, Z)$ and $\binom{I-V}{-I-V} \in \mathcal{L}(Z, Z \times Z)$.
Proof: Assume $\binom{x}{y} \in \operatorname{ker}\left[\begin{array}{ll}I+V & I-V\end{array}\right]$. Thus, it holds

$$
x+V x+y-V y=0 .
$$

For $l:=\frac{1}{2}(x-y) \in Z$ we get
$(I-V) l=\frac{1}{2}(x-y)-\frac{1}{2} V(x-y)=x$ and $(-I-V) l=-\frac{1}{2}(x-y)-\frac{1}{2} V(x-y)=y$.
Thus, it follows $\binom{x}{y} \in \operatorname{ran}\left[\begin{array}{c}I-V \\ -I-V\end{array}\right]$. Conversely, assume $\binom{x}{y} \in \operatorname{ran}\left[\begin{array}{c}I-V \\ -I-V\end{array}\right]$. Then, we have

$$
\left[\begin{array}{ll}
I+V & I-V
\end{array}\right]\binom{x}{y}=\left[\begin{array}{ll}
I+V & I-V
\end{array}\right]\left[\begin{array}{c}
I-V \\
-I-V
\end{array}\right] l=0
$$

for some $l \in Z$ and the lemma is proved.

Lemma 6.1.15. ( $(\overline{\text { KZ15 }}$, Lemma 2.4 $])$ Let $W=\left[\begin{array}{ll}W_{1} & W_{0}\end{array}\right] \in \mathcal{L}\left(H^{2 N}, H^{N}\right)$ such that $W_{1}+W_{0}$ is injective and

$$
\operatorname{ran}\left(W_{1}-W_{0}\right) \subseteq \operatorname{ran}\left(W_{1}+W_{0}\right)
$$

Then there exist an unique operator $V \in \mathcal{L}\left(H^{N}\right)$ such that

$$
W=\left[\begin{array}{ll}
W_{1} & W_{0}
\end{array}\right]=\frac{1}{2}\left(W_{1}+W_{0}\right)\left[\begin{array}{ll}
I+V & I-V \tag{6.14}
\end{array}\right] .
$$

Moreover,

$$
\operatorname{ker}\left[\begin{array}{ll}
W_{1} & W_{0}
\end{array}\right]=\operatorname{ker}\left[\begin{array}{ll}
I+V & I-V
\end{array}\right],
$$

and

$$
\left[\begin{array}{ll}
W_{1} & W_{0}
\end{array}\right]\left[\begin{array}{ll}
0 & I \\
I & 0
\end{array}\right]\left[\begin{array}{ll}
W_{1} & W_{0}
\end{array}\right]^{*} \geqslant 0 \Leftrightarrow V V^{*} \leqslant I .
$$

Lemma 6.1.16. Let $A_{0}$ be defined as in Definition 6.1.12. For an arbitrary element $\binom{u}{v} \in H^{N} \times H^{N}$ there exists a function $x \in \mathcal{D}\left(A_{0}\right)$ such that $\Phi(x)=$ $\binom{u}{v}$.

Proof: We give a constructive proof: Consider $\binom{u}{v} \in H^{N} \times H^{N}$ where

$$
u=\left[\begin{array}{c}
u_{1} \\
\vdots \\
u_{N}
\end{array}\right] \text { and } v=\left[\begin{array}{c}
v_{1} \\
\vdots \\
v_{N}
\end{array}\right],
$$

with entries $u_{1}, \ldots, u_{N}, v_{1}, \ldots, v_{N} \in H$. To construct a proper function $x$, we define two polynomials, $P_{u}(\zeta)$ and $P_{v}(\zeta)$, by

$$
P_{u}(\zeta):=\sum_{i=0}^{N} \frac{u_{i+1}}{i!}(\zeta-1)^{i} \text { and } P_{v}(\zeta):=\sum_{i=0}^{N} \frac{v_{i+1}}{i!} \zeta^{i} .
$$

Furthermore, we define the functions $\varphi_{0} \in \mathcal{C}^{\infty}[0,1]$ and $\varphi_{1} \in \mathcal{C}^{\infty}[0,1]$ such that $\left.\varphi_{0}\right|_{[0, \varepsilon]}=0$ and $\left.\varphi_{0}\right|_{[1-\varepsilon, 1]}=1$ and analogously $\left.\varphi_{1}\right|_{[0, \varepsilon]}=1$ and $\left.\varphi_{1}\right|_{[1-\varepsilon, 1]}=0$ hold. Thus, for

$$
x:=\left(\varphi_{0} \cdot P_{u}+\varphi_{1} \cdot P_{v}\right) I_{H^{N}} \in \mathcal{C}^{\infty}\left([0,1] ; H^{N}\right) \subseteq \mathcal{D}\left(A_{0}\right)
$$

we get $\Phi(x)=\binom{u}{v}$.
Lemma 6.1.17. Let $A$ be defined by (3.9)-(3.10). Then $A$ is dissipative if and only if $A-P_{0}$ is dissipative and it holds $\operatorname{Re} P_{0} \leqslant 0$.

Proof: " $\Rightarrow$ ": Let $A$ be dissipative. Hence, the operator $A-P_{0}$ is dissipative if $\operatorname{Re} P_{0} \leqslant 0$ holds. We will prove $\operatorname{Re}\left\langle P_{0} z, z\right\rangle \leqslant 0$ for all $z \in H$ : Let $z \in H$ and $\Psi(\zeta) \in \mathcal{C}_{c}^{\infty}(0,1)$ with $\zeta \in[0,1]$ an arbitrary, scalar-valued function with $\Psi \not \equiv 0$. We define

$$
x:=\Psi(\zeta) z \in \mathcal{C}_{c}^{\infty}((0,1) ; H) \subseteq \mathcal{D}(A)
$$

and it yields, since the derivation equals zero at the boundary,

$$
\begin{aligned}
0 \geqslant \operatorname{Re}\langle A x, x\rangle & =\operatorname{Re}\left\langle P_{0} x, x\right\rangle=\operatorname{Re}\left\langle P_{0} \Psi z, \Psi z\right\rangle \\
& =\operatorname{Re} \int_{0}^{1}|\Psi(\zeta)|^{2}\left\langle P_{0} z, z\right\rangle_{H} d \zeta \\
& =\|\Psi\|^{2} \operatorname{Re}\left\langle P_{0} z, z\right\rangle_{H} .
\end{aligned}
$$

$" \Leftarrow ":$ We assume $\operatorname{Re} P_{0} \leqslant 0$ and $\operatorname{Re}\left\langle\left(A-P_{0}\right) x, x\right\rangle \leqslant 0$ for all $x \in \mathcal{D}(A)$. Thus, we get for $x \in \mathcal{D}(A)$

$$
\operatorname{Re}\langle A x, x\rangle=\operatorname{Re}\left\langle\left(A-P_{0}\right) x, x\right\rangle+\operatorname{Re}\left\langle P_{0} x, x\right\rangle \leqslant 0
$$

Thus, we get the assertion of the lemma.
We are now in the position to prove the main results of this section.
Proof of Theorem 6.1.3; $\quad$ The implication $1 \Rightarrow 2$ follows by the LumerPhillips Theorem, cf. Theorem 2.1.14, and the equivalence $3 \Leftrightarrow 4$ has been shown in Lemma 6.1.15,
Next, we prove the equivalence $2 \Leftrightarrow 5$ : Lemma 6.1.13 implies for $x \in \mathcal{D}(A)$

$$
\operatorname{Re}\langle A x, x\rangle=\Phi_{1}(x)^{*} Q \Phi_{1}(x)-\Phi_{0}(x)^{*} Q \Phi_{0}(x)+\operatorname{Re}\left\langle P_{0} x, x\right\rangle .
$$

Note that $x \in \mathcal{W}^{N, 2}(0,1 ; H)$ satisfies $x \in \mathcal{D}(A)$ if and only if $\binom{\Phi_{1}(x)}{\Phi_{0}(x)} \in \operatorname{ker} \widetilde{W}_{B}$. This proves the implication $5 \Rightarrow 2$.
We now assume that 2 holds. Then Lemma 6.1.17 shows that $\operatorname{Re} P_{0} \leqslant 0$ and that $A-P_{0}$ is dissipative, that is,

$$
\Phi_{1}(x)^{*} Q \Phi_{1}(x)-\Phi_{0}(x)^{*} Q \Phi_{0}(x) \leqslant 0
$$

for every $x \in \mathcal{W}^{N, 2}(0,1 ; H)$ satisfying $\binom{\Phi_{1}(x)}{\Phi_{0}(x)} \in \operatorname{ker} \widetilde{W}_{B}$. Further, by Lemma 6.1.16. for an arbitrary element $\binom{u}{v} \in \operatorname{ker} \widetilde{W}_{B}$ there exists a function $x \in \mathcal{D}(A)$ such that $\binom{\Phi_{1}(x)}{\Phi_{0}(x)}=\binom{u}{v}$. This proves 5 .
Next, we prove the implication $2 \Rightarrow 4$ : Lemma 6.1.17 shows that $\operatorname{Re} P_{0} \leqslant 0$ and that $A-P_{0}$ is dissipative, that is, using Lemma 6.1.13

$$
\begin{equation*}
\operatorname{Re}\left\langle f_{\partial, x}, e_{\partial, x}\right\rangle_{H^{N}} \leqslant 0, \quad x \in \mathcal{D}(A) \tag{6.15}
\end{equation*}
$$

For an arbitrary element $\binom{f}{e} \in \operatorname{ker} W_{B} \subseteq H^{N} \times H^{N}$ a function $x \in \mathcal{D}(A)$ exists due to Lemma 6.1.16 such that $R_{e x t} \Phi(x)=\left(\begin{array}{c}f \\ f_{2, x} \\ e, x\end{array}\right)=\binom{f}{e}$. By equation 6.15) we get $e^{*} f+f^{*} e \leqslant 0$ for all $\binom{f}{e} \in \operatorname{ker} W_{B}$, where $W_{B}:=\left[\begin{array}{ll}W_{1} & W_{0}\end{array}\right]$. For $y \in \operatorname{ker}\left(W_{1}+W_{0}\right)$ we have $W_{B}\binom{y}{y}=0$ and thus $y^{*} y+y y^{*} \leqslant 0$. Since the norm of an element is non negative, it follows $y=0$ and therefore $\operatorname{ker}\left(W_{1}+W_{0}\right)=\{0\}$, which shows the injectivity of $W_{1}+W_{0}$. Due to this fact, by Lemma 6.1.15 there exists an operator $V$ satisfying (6.14). It remains to show that $\|V\| \leqslant 1$. Let $l \in H^{N}$ be arbitrarily. By Lemma 6.1.14 we obtain $\binom{I-V}{-I-V} l \in \operatorname{ker} W_{B}$.

From Lemma 6.1.16 we follow that a function $x \in \mathcal{D}\left(A_{0}\right)$ exists, such that $R_{\text {ext }} \Phi(x)=\binom{f_{\partial, x}}{e_{\partial, x}}=\binom{I-V}{-I-V} l$. Therefore, $\binom{f_{\partial, x}}{e_{\partial, x}} \in \operatorname{ker} W_{B}$ and even $x \in$ $\mathcal{D}(A)$. In conclusion, we obtain using (6.15)

$$
\begin{align*}
2 \operatorname{Re}\left\langle f_{\partial, x}, e_{\partial, x}\right\rangle_{H^{N}} & =\left\langle f_{\partial, x}, e_{\partial, x}\right\rangle_{H^{N}}+\left\langle e_{\partial, x}, f_{\partial, x}\right\rangle_{H^{N}} \\
& =\langle(I-V) l,(-I-V) l\rangle_{H^{N}}+\langle(-I-V) l,(I-V) l\rangle_{H^{N}} \\
& =2\left\langle l,\left(-I+V^{*} V\right) l\right\rangle_{H^{N}} \leqslant 0 \tag{6.16}
\end{align*}
$$

and therefore $\|V\| \leqslant 1$.
Finally, we show the implication $4 \Rightarrow 1: A$ is a closed operator, see Aug16, Lemma 3.2.2]. To prove that $A$ generates a contraction $C_{0}$-semigroup, it is sufficient to verify that $A$ and $A^{*}$ are dissipative, cf. Theorem 2.1.15. Let $x \in \mathcal{D}(A)$. Then, we have $\binom{f_{\partial, x}}{e_{\partial, x}} \in \operatorname{ker} W_{B}$ and from Lemma 6.1.14 it follows that there exists an $l \in H^{N}$ such that $\binom{f_{\partial, x}}{e, x, x}=\binom{I-V}{-I-V} l$. Using Lemma 6.1.13 and Lemma 6.1.15, we obtain

$$
\begin{aligned}
2 \operatorname{Re}\langle A x, x\rangle_{L^{2}} & =2 \operatorname{Re}\left\langle f_{\partial, x}, e_{\partial, x}\right\rangle_{H^{N}}+2\left\langle P_{0} x, x\right\rangle \\
& \leqslant 2\left\langle l,\left(-I+V^{*} V\right) l\right\rangle_{H^{N}} \leqslant 0 .
\end{aligned}
$$

Now we consider the adjoint operator $A^{*}$ : Let $y \in \mathcal{D}\left(A^{*}\right)$. By Lemma 6.1.11, we obtain $\binom{\tilde{f}_{\partial, y}}{\tilde{e}_{2, y}} \in \operatorname{ker} S\left[\begin{array}{ll}I+V^{*} & I-V^{*}\end{array}\right]$. Applying Lemma 6.1.14 and Lemma 6.1.15 to the operator $V^{*}$, there exists $m \in H^{N}$ such that $\binom{f \overrightarrow{f_{, x, x}}}{\tilde{e}, x}=\binom{I-V^{*}}{-I-V^{*}} m$. Using again Lemma 6.1.13 we get

$$
\begin{equation*}
2 \operatorname{Re}\left\langle A^{*} y, y\right\rangle_{L^{2}} \leqslant 2\left\langle m,\left(-I+V V^{*}\right) m\right\rangle_{H^{N}} \leqslant 0, \tag{6.17}
\end{equation*}
$$

which concludes the proof.
Proof of Corollary 6.1.5: We want to apply Theorem6.1.3 for the proof of Corollary 6.1.5. Therefore, we have to check condition 6.9).
If $\operatorname{dim} H<\infty$, then $W_{1}+W_{0}$ injective implies the surjectivity of $W_{1}+W_{0}$ and hence condition (6.9). Due to this and Remark 6.1.4. 2 assertions 1, 2, 3, 4 and 5 of Theorem 6.1.5 are equivalent. The implications $3 \Rightarrow 6$ and $4 \Rightarrow 7$ follows, since we have $W_{1}+W_{0}$ injective, and thus, $W_{1}+W_{2}$ is also surjective. Clearly, it follows that $W_{B}$ is surjective. A straightforward calculation shows the implication $7 \Rightarrow 6$. In order to show $6 \Rightarrow 3$ we prove that in the finite-dimensional setting the surjectivity of $W_{B}$ and $W_{B} \Sigma W_{B}^{*} \geqslant 0$ implies the injectivity of $W_{1}+W_{0}$. From

$$
W_{B} \Sigma W_{B}^{*} \geqslant 0 \Leftrightarrow W_{0} W_{1}^{*}+W_{1} W_{0}^{*} \geqslant 0,
$$

we obtain

$$
\begin{aligned}
& W_{1} W_{1}^{*}+W_{0} W_{1}^{*}+W_{0} W_{0}^{*}+W_{1} W_{0}^{*} \\
& =\left(W_{1}+W_{0}\right)\left(W_{1}+W_{0}\right)^{*} \geqslant\left(W_{1}-W_{0}\right)\left(W_{1}-W_{0}\right)^{*} \geqslant 0 .
\end{aligned}
$$

Let $x$ be in $\operatorname{ker}\left(W_{1}+W_{0}\right)^{*}$. This yields $x \in \operatorname{ker}\left(W_{1}-W_{0}\right)\left(W_{1}-W_{0}\right)^{*}$. With

$$
\begin{aligned}
\left\|\left(W_{1}-W_{0}\right)^{*} x\right\|^{2} & =\left\langle\left(W_{1}-W_{0}\right)^{*} x,\left(W_{1}-W_{0}\right)^{*} x\right\rangle \\
& =\left\langle x,\left(W_{1}-W_{0}\right)\left(W_{1}-W_{0}\right)^{*} x\right\rangle=\langle x, 0\rangle=0
\end{aligned}
$$

we get $x \in \operatorname{ker}\left(W_{1}-W_{0}\right)^{*}$ and thus, $x \in \operatorname{ker} W_{1}^{*} \cap W_{0}^{*}$. Since $W_{B}$ is surjective, $W_{B}^{*}$ is injective and thus $x=0$. This implies that $W_{1}+W_{0}$ is injective.
Proof of Theorem 6.1.7: Without loss of generality again we consider just the case $\mathcal{H}=I$. In the following proof we will often apply Theorem 6.1.3 to the operators $A$ and $-A$. So, first of all, we have to verify, that also the boundary condition operator $\bar{W}_{B}$ of $-A$ satisfies the condition (6.9).
We define analogously to (6.11) the boundary flow and the boundary effort for $-A$ :

$$
\left[\begin{array}{l}
\bar{f}_{\partial, x}  \tag{6.18}\\
\bar{e}_{\partial, x}
\end{array}\right]:=\frac{1}{\sqrt{2}}\left[\begin{array}{cc}
-Q & Q \\
I & I
\end{array}\right] \Phi(\mathcal{H} x)
$$

Therefore, it yields $\bar{f}_{\partial, x}=-f_{\partial, x}$ and $\bar{e}_{\partial, x}=e_{\partial, x}$. Due to $\mathcal{D}(A)=\mathcal{D}(-A)$, we get

$$
\begin{aligned}
\mathcal{D}(A) & =\left\{x \in \mathcal{W}^{N, 2}(0,1 ; H) \left\lvert\, W_{B}\left[\begin{array}{l}
f_{\partial, x} \\
e_{\partial, x}
\end{array}\right]=0\right.\right\} \\
& =\left\{x \in \mathcal{W}^{N, 2}(0,1 ; H) \left\lvert\, \bar{W}_{B}\left[\begin{array}{l}
\bar{f}_{\partial, x} \\
\bar{e}_{\partial, x}
\end{array}\right]=0\right.\right\} \\
& =\left\{x \in \mathcal{W}^{N, 2}(0,1 ; H) \left\lvert\, \bar{W}_{B}\left[\begin{array}{c}
-f_{\partial, x} \\
e_{\partial, x}
\end{array}\right]=0\right.\right\}
\end{aligned}
$$

and thus,

$$
\bar{W}_{B}=\left[\begin{array}{ll}
-W_{1} & W_{0} \tag{6.19}
\end{array}\right] .
$$

It is easy to check that under condition (6.10) the operator $\bar{W}_{B}$ satisfies 6.9). Then the equivalences $1 \Leftrightarrow 2 \Leftrightarrow 5$ follow by Theorem 6.1 .3 applied for $A$ and $-A$.
$1 \Rightarrow 4$ : Let $A$ be the generator of a unitary group. Then, due to Theorem [JZ12, Theorem 6.2.5] $A$ and $-A$ are generators of contraction $C_{0}$-semigroups. It follows $\operatorname{Re} P_{0}=0, W_{1}+W_{0}$ and $-W_{1}+W_{0}$ are injective and $\operatorname{Re}\langle A x, x\rangle=0$ for all $x \in \mathcal{D}(A)$ by Theorem 6.1.3. Thus, we get with the estimation 6.16)

$$
\begin{equation*}
\left.0=2 \operatorname{Re}\langle A x, x\rangle=2\left\langle l,\left(-I+V^{*} V\right) l\right)\right\rangle_{H^{N}} \text { for all } l \in H^{N} \tag{6.20}
\end{equation*}
$$

and therefore $\|V\|=1$.
$4 \Rightarrow 3$ : Let $\operatorname{Re} P_{0}=0,\|V\|=1, W_{1}+W_{0}$ and $-W_{1}+W_{0}$ injective. Define $S:=\frac{1}{2}\left(W_{1}+W_{0}\right)$ and with the technical Lemma 6.1.15 (Lemma 2.4 in KZ15) it yields

$$
\begin{aligned}
W_{B} \Sigma W_{B}^{*} & =S\left[\begin{array}{ll}
I+V & I-V
\end{array}\right]\left[\begin{array}{ll}
0 & I \\
I & 0
\end{array}\right]\left(S\left[\begin{array}{ll}
I+V & I-V
\end{array}\right]\right)^{*} \\
& =S\left(2 I-2 V V^{*}\right) S^{*}=0 .
\end{aligned}
$$

The implication $3 \Rightarrow 1$ follows analogously to the proof of $3 \Rightarrow 1$ in Theorem 6.1 .3 for the operator $-A$. However, instead of the boundary effort and the boundary flow for $A$ we need to consider them for $-A$ and have to determine the boundary condition operator $\bar{W}_{B}$ for $-A$.

### 6.1.1 Examples for port-Hamiltonian systems in the infinitedimensional setting

In this section we now illustrate our results by a number of examples. Networks of discrete partial differential equations on infinite networks are also considered in (Mug14).
Further examples on the interval $(0,1)$ with a finite-dimensional Hilbert space $H$ can be found in [JZ12] and Aug16]. In the following we consider examples on the bounded interval $(0,1)$ with an infinite-dimensional Hilbert space $H$.
Example 6.1.18. Choose $H=\ell^{2}(\mathbb{N})$ and consider the operator $A$ given by

$$
\begin{equation*}
A f=\frac{\partial}{\partial \zeta} f \tag{6.21}
\end{equation*}
$$

on the domain

$$
\mathcal{D}(A)=\left\{f \in \mathcal{W}^{1,2}\left(0,1 ; \ell^{2}(\mathbb{N})\right) \left\lvert\,\left[\begin{array}{ll}
I & -L \tag{6.22}
\end{array}\right] \Phi(f)=0\right.\right\}
$$

This means that the network is a path graph, see Figure 6.3.

Figure 6.3: Path graph
Clearly, $A$ denotes a port-Hamiltonian operator with $N=1, P_{1}=I, P_{0}=0$ and $W_{B}=\left[\begin{array}{ll}W_{1} & W_{0}\end{array}\right]=\frac{1}{2}\left[\begin{array}{ll}I+L & I-L\end{array}\right]$. Here $L$ denotes the left shift and $L^{*}=R$ the right shift, i.e., $L: \ell^{2}(\mathbb{N}) \rightarrow \ell^{2}(\mathbb{N})$ is defined by $L\left(x_{1}, x_{2}, \ldots\right) \mapsto\left(x_{2}, x_{3}, \ldots\right)$ and $R: \ell^{2}(\mathbb{N}) \rightarrow \ell^{2}(\mathbb{N})$ is given as $R\left(x_{1}, x_{2}, \ldots\right) \mapsto\left(0, x_{1}, x_{2}, \ldots\right)$. Clearly, it yields $W_{1}+W_{0}=I$, and thus, condition $(6.9)$ is fulfilled. Therefore, we can apply Theorem 6.1.3 and check assertion 3: $W_{1}+W_{0}$ is injective and

$$
\begin{aligned}
W_{B} \Sigma W_{B}^{*} & =\frac{1}{4}\left[\begin{array}{ll}
I+L & I-L
\end{array}\right] \Sigma\left[\begin{array}{ll}
I+L & I-L
\end{array}\right]^{*} \\
& =\frac{1}{4}\left[\begin{array}{ll}
I-L & I+L
\end{array}\right]\left[\begin{array}{ll}
I+L^{*} & I-L^{*}
\end{array}\right] \\
& =\frac{1}{4}\left((I-L)\left(I+L^{*}\right)+(I+L)\left(I-L^{*}\right)\right)=\frac{1}{4}\left(2 I-2 L L^{*}\right)=0 .
\end{aligned}
$$

Hence, $A$ generates a contraction $C_{0}$-semigroup. In the finite-dimensional setting we would expect that $A$ also generates a unitary $C_{0}$-group, since $W_{B} \Sigma W_{B}^{*}=$ 0 . However, we can apply Theorem 6.1.7, since condition (6.10) is fulfilled: $\operatorname{ran}(L)=\operatorname{ran}(I)$, because the left shift is surjective. Using assertion 3 of Theorem 6.1.7, we can conclude that $A$ does not generate a unitary $C_{0}$-group, since $-W_{1}+W_{0}=-L$ and the left shift is not injective.

Example 6.1.19. We choose again $H=\ell^{2}(\mathbb{N})$ and consider the operator $A$ given by

$$
\begin{equation*}
A f=\frac{\partial}{\partial \zeta} f \tag{6.23}
\end{equation*}
$$

on the domain

$$
\mathcal{D}(A)=\left\{f \in \mathcal{W}^{1,2}\left(0,1 ; \ell^{2}(\mathbb{N})\right) \left\lvert\,\left[\begin{array}{ll}
I & T \tag{6.24}
\end{array}\right] \Phi(f)=0\right.\right\}
$$

where $T: \ell^{2}(\mathbb{N}) \rightarrow \ell^{2}(\mathbb{N})$ is defined by

$$
T\left(x_{1}, x_{2}, \ldots\right) \mapsto \frac{1}{2}\left(-x_{3}-x_{4},-x_{5}-x_{6},-x_{7}-x_{8}, \ldots\right)
$$

These boundary conditions imply that the network is a binary tree, see Figure 6.4.


Figure 6.4: Binary tree
We write $f \in \mathcal{W}^{1,2}\left(0,1 ; \ell^{2}(\mathbb{N})\right)$ as $f=\left(f_{1}, f_{2}, \ldots\right)^{T}$, where $f_{i} \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{d}\right)$ denotes a function on the $i$-th edge of the binary tree. Clearly, $A$ denotes a portHamiltonian operator with $N=1, P_{1}=I, P_{0}=0$ and $W_{B}=\frac{1}{2}[I-T \quad I+T]$. It yields $W_{1}+W_{0}=I$, and thus, condition (6.9) is fulfilled.
$W_{1}+W_{0}$ is injective and $T^{*}: \ell^{2}(\mathbb{N}) \rightarrow \ell^{2}(\mathbb{N})$ is given by

$$
T^{*}\left(x_{1}, x_{2}, \ldots\right) \mapsto \frac{1}{2}\left(0,0,-x_{1},-x_{1},-x_{2},-x_{2}, \ldots\right)
$$

We obtain

$$
W_{B} \Sigma W_{B}^{*}=\frac{1}{4}\left(2 I-2 T T^{*}\right)=\frac{1}{4} I
$$

Hence, $A$ generates a contraction $C_{0}$-semigroup.

### 6.2 Port-Hamiltonian systems on the semi-axis

In this section, we consider port-Hamiltonian systems on the semi-axis, i.e., systems of the form

$$
\begin{align*}
\frac{\partial x}{\partial t}(\zeta, t) & =\left(P_{1} \frac{\partial}{\partial \zeta}+P_{0}\right)(\mathcal{H}(\zeta) x(\zeta, t)) \\
x(\zeta, 0) & =x_{0}(\zeta), \zeta \in(0, \infty), t \geqslant 0  \tag{6.25}\\
0 & =\widetilde{W}_{B}(\Phi(\mathcal{H} x))(\cdot, t) \tag{6.26}
\end{align*}
$$

where $P_{1}$ is an invertible Hermitian $d \times d$-matrix, $P_{0} \in \mathbb{C}^{d \times d}, \widetilde{W}_{B} \in \mathbb{C}^{k \times d}$ with $k \in\{0,1, \cdots, d\}$ and $\mathcal{H}(\zeta) \in \mathbb{C}^{d \times d}$ is positive definite for a.e. $\zeta \in[0, \infty)$ satisfying $\mathcal{H}, \mathcal{H}^{-1} \in L^{\infty}\left(0, \infty ; \mathbb{C}^{d \times d}\right)$. Since $P_{1}$ is an invertible Hermitian matrix, its eigenvalues are real and nonzero.
Here $\widetilde{W}_{B} \in \mathcal{L}\left(H^{N}, \widetilde{H}^{N}\right)$, where $\widetilde{H}$ is a subspace of $H$, and $\Phi$ is given by

$$
\Phi: \mathcal{W}^{N, 2}(0, \infty ; H) \rightarrow H^{N}, \quad \Phi(x):=\Phi_{0}(x)
$$

The contraction $C_{0}$-semigroup property has been shown for some specific examples [EN00, I.4.16], MNS18], and related results can be found in BK13, EKF19], KPS08], KS99 and [SSWW15]. In the following we provide a characterization of the contraction $C_{0}$-semigroup property of the operator $A$. Again $A$ generates a contraction $C_{0}$-semigroup if and only if the operator $A$ is dissipative. The main difference to the port-Hamiltonian systems on a bounded interval is that the number of boundary conditions depends on $P_{1}$.
We consider the port-Hamiltonian operator $A$, associated to the system (6.25),

$$
\begin{align*}
A x & =P_{0} \mathcal{H} x+P_{1} \frac{\partial}{\partial \zeta}(\mathcal{H} x) \text { with }  \tag{6.27}\\
\mathcal{D}(A) & =\left\{x \in L^{2}\left(0, \infty ; \mathbb{C}^{d}\right) \mid \mathcal{H} x \in \mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d}\right), \widetilde{W}_{B}(\mathcal{H} x(0))=0\right\} \tag{6.28}
\end{align*}
$$

on the space $X=L^{2}\left(0, \infty ; \mathbb{C}^{d}\right)$.
We denote by $d_{1}$ the number of positive and by $d_{2}=d-d_{1}$ the number of negative eigenvalues of $P_{1}$ and write

$$
P_{1}=S^{-1} \Delta S=S^{-1}\left[\begin{array}{cc}
\Lambda & 0  \tag{6.29}\\
0 & \Theta
\end{array}\right] S
$$

with a unitary matrix $S \in \mathbb{C}^{d \times d}$, a positive definite diagonal matrix $\Lambda \in \mathbb{R}^{d_{1} \times d_{1}}$, and a negative definite diagonal matrix $\Theta \in \mathbb{R}^{d_{2} \times d_{2}}$. We define $\Delta=\left(\begin{array}{cc}\Lambda & 0 \\ 0 & \Theta\end{array}\right)$.
In the following, first we formulate the main results of this section, then we give a technical lemma and finally the proof of the main result.

Theorem 6.2.1. Assume $A$ is given by (6.27)-(6.28), $\widetilde{W}_{B} \in \mathbb{C}^{k \times d}$ with $k \leqslant d_{2}$ has full row rank. Then the following statements are equivalent:

1. A generates a contraction $C_{0}$-semigroup on $X$;
2. $\operatorname{Re}\langle A x, x\rangle \leqslant 0$ for every $x \in \mathcal{D}(A)$;
3. $\operatorname{Re} P_{0} \leqslant 0$ and $y^{*} P_{1} y \geqslant 0$ for every $y \in \operatorname{ker} \widetilde{W}_{B}$;
4. $\operatorname{Re} P_{0} \leqslant 0, k=d_{2}$ and $\widetilde{W}_{B}=B\left[\begin{array}{ll}U & I\end{array}\right] S$, with $B \in \mathbb{C}^{d_{2} \times d_{2}}$ invertible, $U \in \mathbb{C}^{d_{2} \times d_{1}}$ and $\Lambda+U^{*} \Theta U \geqslant 0$.

Further, we are able to characterize the property of unitary group generation for port-Hamiltonian operators on the semi-axis.

Theorem 6.2.2. Let $A$ be given by 6.27)-6.28, $\widetilde{W}_{B} \in \mathbb{C}^{k \times d}$ with $k \leqslant$ $\min \left\{d_{1}, d_{2}\right\}$ has full row rank. Then the following statements are equivalent:

1. A generates a unitary $C_{0}$-group on $X$;
2. $\operatorname{Re}\langle A x, x\rangle=0$ for every $x \in \mathcal{D}(A)$;
3. $\operatorname{Re} P_{0}=0$ and $y^{*} P_{1} y=0$ for every $y \in \operatorname{ker} \widetilde{W}_{B}$;
4. $k=d_{1}=d_{2}$, Re $P_{0}=0$ and $\widetilde{W}_{B}=\left[\begin{array}{ll}U_{1} & U_{2}\end{array}\right] S$, where $U_{1}, U_{2} \in \mathbb{C}^{d_{1} \times d_{1}}$ are invertible with $\Lambda+U_{1}^{*} U_{2}^{-*} \Theta U_{2}^{-1} U_{1}=0$.

For the proof of the main statements we need the following technical assertions.
Lemma 6.2.3. 1. Assume $\Lambda \in \mathbb{R}^{d_{1} \times d_{1}}$ is a positive, invertible diagonal matrix and $y \in L^{2}\left(0, \infty ; \mathbb{C}^{d_{1}}\right)$. Then the function

$$
\begin{equation*}
x(t):=\int_{0}^{\infty} e^{-s \Lambda^{-1}} \Lambda^{-1} y(s+t) d s, \quad t \geqslant 0, \tag{6.30}
\end{equation*}
$$

satisfies $x \in \mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d_{1}}\right)$ and $x-\Lambda x^{\prime}=y$.
2. Let $\Theta \in \mathbb{R}^{d_{2} \times d_{2}}$ be a negative, invertible diagonal matrix, $y \in L^{2}\left(0, \infty ; \mathbb{C}^{d_{2}}\right)$ and $x_{0} \in \mathbb{C}^{d_{2}}$. Then the differential equation

$$
\begin{equation*}
x-\Theta x^{\prime}=y, \quad x(0)=x_{0}, \tag{6.31}
\end{equation*}
$$

has a unique solution satisfying $x \in \mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d_{2}}\right)$.
Proof Part 1: $\Lambda>0$ and $y \in L^{2}\left(0, \infty ; \mathbb{C}^{d_{1}}\right)$ imply that $x(t)$ is well defined for every $t \geqslant 0$. Minkowski's integral inequality shows $x \in L^{2}\left(0, \infty ; \mathbb{C}^{d_{1}}\right)$. Further, the solution of $x-\Lambda x^{\prime}=y$, or equivalently, of $x^{\prime}=\Lambda^{-1} x-\Lambda^{-1} y$ is given by

$$
x(t)=e^{t \Lambda^{-1}} x(0)-\int_{0}^{t} e^{(t-s) \Lambda^{-1}} \Lambda^{-1} y(s) d s, \quad t \geqslant 0 .
$$

The choice of $x(0)=\int_{0}^{\infty} e^{-s \Lambda^{-1}} \Lambda^{-1} y(s) d s$, implies 6.30). Moreover, $x^{\prime}=$ $\Lambda^{-1} x-\Lambda^{-1} y$ and hence $x \in \mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d_{1}}\right)$.
Part 2: We first note that (6.31) is equivalent to $x^{\prime}=\Theta^{-1} x-\Theta^{-1} y$. Now the statement of the lemma follows from ODE-Theory for linear stable systems, since $\Theta<0$ and $y \in L^{2}\left(0, \infty ; \mathbb{C}^{d_{2}}\right)$, see HP05, Proposition 3.3.22].

Proof of Theorem 6.2.1: $\quad$ Thanks to the Theorem of Lumer-Phillips, cf. Theorem 2.1.14, it holds 1 implies 2.
Next, we show the implication $2 \Rightarrow 3$. Using integration by parts and $P_{1}^{*}=$ $P_{1}$, it yields $2 \operatorname{Re}\left\langle P_{1} \frac{d}{d \zeta} x, x\right\rangle=-x(0)^{*} P_{1} x(0)$, since $\lim _{\zeta \rightarrow \infty} x(\zeta)=0$ for $x \in$ $\mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d}\right)$. Thus, for $x \in D(A)$ we have
$2 \operatorname{Re}\langle A x, x\rangle=2 \operatorname{Re}\left\langle P_{1} \frac{d}{d \zeta} x+P_{0} x, x\right\rangle=-x(0)^{*} P_{1} x(0)+2 \operatorname{Re} \int_{0}^{\infty} x(\zeta)^{*} P_{0} x(\zeta) d \zeta$.
Choosing $x \in \mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d}\right) \backslash\{0\}$ with $x(0)=0$, we obtain $\operatorname{Re} P_{0} \leqslant 0$. For every $y \in \mathbb{C}^{d}$ and every $\varepsilon>0$ there exists a function $x \in \mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d}\right)$ such that $x(0)=y$ and the $L^{2}$-norm of $x$ is less than $\varepsilon$. Choosing this function in
equation (6.32) and letting $\varepsilon$ go to zero implies the second assertion in 3 . In order to prove the implication $3 \Rightarrow 4$, for $x \in D(A)$ we define $\left[\begin{array}{l}f_{1} \\ f_{2}\end{array}\right]:=S x(0)$. Using (6.29), the second condition in 3 can be written as

$$
\left[\begin{array}{ll}
f_{1}^{*} & f_{2}^{*}
\end{array}\right]\left[\begin{array}{cc}
\Lambda & 0  \tag{6.33}\\
0 & \Theta
\end{array}\right]\left[\begin{array}{l}
f_{1} \\
f_{2}
\end{array}\right] \geqslant 0, \quad \text { for }\left[\begin{array}{c}
f_{1} \\
f_{2}
\end{array}\right] \in \operatorname{ker} \widetilde{W}_{B} S^{-1}
$$

Since $\widetilde{W}_{B} S^{-1}$ is a full row rank $k \times d$-matrix with $k \leqslant d_{2}$, its kernel has dimension $d-k$. By the assumptions on $\Lambda$ and $\Theta$, we have $d-k \leqslant d_{1}$, or equivalently, $k \geqslant d_{2}$. Thus $k=d_{2}$.
We write $\widetilde{W}_{B} S^{-1}=\left[\begin{array}{ll}U_{1} & U_{2}\end{array}\right]$ with $U_{1} \in \mathbb{C}^{d_{2} \times d_{1}}$ and $U_{2} \in \mathbb{C}^{d_{2} \times d_{2}}$. Assuming $U_{2}$ is not invertible, there exists $u \in \mathbb{C}^{d_{2}}$ such that $\left[\begin{array}{l}0 \\ u\end{array}\right] \in \operatorname{ker} \widetilde{W}_{B} S^{-1}$ which is in contradiction to (6.33), since $\Theta<0$. Thus, the matrix $\widetilde{W}_{B} S^{-1}$ is of the form $B\left[\begin{array}{ll}U & I\end{array}\right]$, with $U \in \mathbb{C}^{d_{2} \times d_{1}}$ and $B \in \mathbb{C}^{d_{2} \times d_{2}}$ invertible. Hence, (6.33) is equivalent to

$$
\left[\begin{array}{ll}
f_{1}^{*} & f_{2}^{*}
\end{array}\right]\left[\begin{array}{cc}
\Lambda & 0  \tag{6.34}\\
0 & \Theta
\end{array}\right]\left[\begin{array}{l}
f_{1} \\
f_{2}
\end{array}\right] \geqslant 0 \text { and } U f_{1}+f_{2}=0 \text {, for }\left[\begin{array}{l}
f_{1} \\
f_{2}
\end{array}\right] \in \mathbb{C}^{d_{1}+d_{2}}
$$

which is equivalent to $\Lambda+U^{*} \Theta U \geqslant 0$. This shows 4 .
It remains to show that 4 implies 1 . Due to the fact that $\operatorname{Re} P_{0} \leqslant 0$, and bounded, dissipative perturbations of generators of contraction $C_{0}$-semigroups, again generate a contraction $C_{0}$-semigroup, see EN00, Theorem III.2.7], without loss of generality we may assume $P_{0}=0$.
First, we prove the dissipativity of the operator $A$. Let $x \in \mathcal{D}(A)$ and define $\left[\begin{array}{l}f_{1} \\ f_{2}\end{array}\right]:=S x(0)$, where the unitary matrix $S$ is given by (6.29). This implies $U f_{1}+f_{2}=0$ as $\widetilde{W}_{B}=B\left[\begin{array}{ll}U & I\end{array}\right] S$.
Thus, it yields

$$
\begin{aligned}
\operatorname{Re}\langle A x, x\rangle & =-\left\langle x(0), P_{1} x(0)\right\rangle_{\mathbb{C}^{d}}=-\left\langle x(0), S^{-1}\left[\begin{array}{cc}
\Lambda & 0 \\
0 & \Theta
\end{array}\right] S x(0)\right\rangle_{\mathbb{C}^{d}} \\
& =-\left\langle S x(0),\left[\begin{array}{cc}
\Lambda & 0 \\
0 & \Theta
\end{array}\right] S x(0)\right\rangle_{\mathbb{C}^{d}}=-\left(f_{1}^{*} \Lambda f_{1}+f_{2}^{*} \Theta f_{2}\right) \\
& =-\left(f_{1}^{*} \Lambda f_{1}+f_{1}^{*} U^{*} \Theta U f_{1}\right) \leqslant 0
\end{aligned}
$$

by the last assertion of 4 .
Further, thanks to the Theorem of Lumer-Phillips, cf. Theorem 2.1.14, it remains to show that for every $y \in L^{2}\left(0, \infty ; \mathbb{C}^{d}\right)$ there exists $x \in D(A)$ such that $x-A x=y$. Equivalently, by (6.29) it is sufficient to show that for every $y_{1} \in$ $L^{2}\left(0, \infty ; \mathbb{C}^{d_{1}}\right)$ and $y_{2} \in L^{2}\left(0, \infty ; \mathbb{C}^{d_{2}}\right)$ there exist functions $x_{1} \in \mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d_{1}}\right)$ and $x_{2} \in \mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d_{2}}\right)$ such that

$$
x_{1}-\Lambda x_{1}^{\prime}=y_{1}, \quad x_{2}-\Theta x_{2}^{\prime}=y_{2} \quad \text { and } \quad U x_{1}(0)+x_{2}(0)=0 .
$$

Let $y_{1} \in L^{2}\left(0, \infty ; \mathbb{C}^{d_{1}}\right)$ and $y_{2} \in L^{2}\left(0, \infty ; \mathbb{C}^{d_{2}}\right)$ be arbitrarily. Lemma 6.2.3 1 1 im plies the existence of $x_{1} \in \mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d_{1}}\right)$ with $x_{1}(0)=\int_{0}^{\infty} e^{-s \Lambda^{-1}} \Lambda^{-1} y_{1}(s) d s$ and $x_{1}-\Lambda x_{1}^{\prime}=y_{1}$. Finally, Lemma 6.2.3|2 shows that there exists a function
$x_{2} \in \mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d_{1}}\right)$ with $x_{2}(0)=-U x_{1}(0)$ and $x_{2}-\Theta x_{2}^{\prime}=y_{2}$. This concludes the proof.
Proof of Theorem 6.2.2; $\quad$ Since $A$ generates a unitary $C_{0}$-group if and only if $A$ and $-A$ generate contraction $C_{0}$-semigroups, cf. Theorem 2.1.16, the equivalence of assertions 1,2 , and 3 follows directly from Theorem 6.2 .1 for $-A$ and $A$.
Formulating assertion 4 of Theorem 6.2 .1 for $-A$, we get $\operatorname{Re}\left(-P_{0}\right) \leqslant 0, k=d_{1}$,

$$
\widetilde{W}_{B}=\bar{B}\left[\begin{array}{ll}
I & \bar{U}
\end{array}\right] S
$$

and $\Theta+\bar{U}^{*} \Lambda \bar{U} \leqslant 0$, where $\bar{B} \in K^{d_{1} \times d_{1}}$ is invertible. Thus, assertion 4 of Theorem 6.2.1 for $-A$ and $A$ is equivalent to $\operatorname{Re} P_{0}=0, k=d_{1}=d_{2}$ and $\widetilde{W}_{B}=\bar{B}\left[\begin{array}{cc}I & \bar{U}\end{array}\right] S=B\left[\begin{array}{ll}U & I\end{array}\right] S$ with $B$ and $\bar{B}$ invertible. It yields $\bar{B}=B U$ and $B=\bar{B} \bar{U}$ with $B, \bar{B}$ invertible. Therefore, we get $\bar{U} U=I$ and $\bar{U}, U$ invertible. Thus, we have $\Theta+\bar{U}^{*} \Lambda \bar{U} \leqslant 0 \Leftrightarrow U^{*} \Theta U+\Lambda \leqslant 0$. Choosing $U_{1}=B U$ and $U_{2}=B$ we get the assertion.

### 6.2.1 Examples for port-Hamiltonian systems on the semi-axis

Example 6.2.4. Let $A$ be given by (6.27)- (6.28) on the semi-axis $(0, \infty)$.

1. Let $P_{1}<0$, that is, $d_{2}=d$, and $\operatorname{Re} P_{0} \leqslant 0$. In this situation $A$ with domain

$$
\mathcal{D}(A)=\left\{x \in X \mid \mathcal{H} x \in \mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d}\right) \text { and }(\mathcal{H} x)(0)=0\right\}
$$

generates a contraction $C_{0}$-semigroup on $X$.
2. Let $P_{1}>0$, that is, $d_{2}=0$ and $\operatorname{Re} P_{0} \leqslant 0$. Then $A$ with domain

$$
\mathcal{D}(A)=\left\{x \in X \mid \mathcal{H} x \in \mathcal{W}^{1,2}\left(0, \infty ; \mathbb{C}^{d}\right)\right\}
$$

generates a contraction $C_{0}$-semigroup on $X$.
3. We consider again the wave equation as in Example 3.1.7 but now on the semi-axis. There, an (undamped) vibrating string can be modelled by

$$
\begin{equation*}
\frac{\partial^{2} w}{\partial t^{2}}(\zeta, t)=\frac{1}{\rho(\zeta)} \frac{\partial}{\partial \zeta}\left(T(\zeta) \frac{\partial w}{\partial \zeta}(\zeta, t)\right), \quad t \geqslant 0, \zeta \in[0, \infty) \tag{6.35}
\end{equation*}
$$

where $w(\zeta, t)$ is the vertical position of the string at place $\zeta$ and time $t$, $T(\zeta)>0$ is the Young's modulus of the string, and $\rho(\zeta)>0$ is the mass density, which may vary along the string. We assume that $T$ and $\rho$ are positive functions satisfying $\rho, \rho^{-1}, T, T^{-1} \in L^{\infty}[0, \infty)$. By choosing the state variables $x_{1}=\rho \frac{\partial w}{\partial t}$ (momentum) and $x_{2}=\frac{\partial w}{\partial \zeta}$ (strain), the partial differential equation (6.35) can equivalently be written as

$$
\begin{align*}
\frac{\partial}{\partial t}\left[\begin{array}{l}
x_{1}(\zeta, t) \\
x_{2}(\zeta, t)
\end{array}\right] & =\left[\begin{array}{ll}
0 & 1 \\
1 & 0
\end{array}\right] \frac{\partial}{\partial \zeta}\left(\left[\begin{array}{cc}
\frac{1}{\rho(\zeta)} & 0 \\
0 & T(\zeta)
\end{array}\right]\left[\begin{array}{l}
x_{1}(\zeta, t) \\
x_{2}(\zeta, t)
\end{array}\right]\right) \\
& =P_{1} \frac{\partial}{\partial \zeta}\left(\mathcal{H}(\zeta)\left[\begin{array}{l}
x_{1}(\zeta, t) \\
x_{2}(\zeta, t)
\end{array}\right]\right), \tag{6.36}
\end{align*}
$$

where $P_{1}=\left[\begin{array}{ll}0 & 1 \\ 1 & 0\end{array}\right]$ and $\mathcal{H}(\zeta)=\left[\begin{array}{cc}\frac{1}{\rho(\zeta)} & 0 \\ 0 & T(\zeta)\end{array}\right]$.
The boundary conditions for (6.36) are

$$
\widetilde{W}_{B}(\mathcal{H} x)(0, t)=0,
$$

where $\widetilde{W}_{B}$ is a $k \times 2$-matrix with rank $k \in\{0,1,2\}$, or equivalently, the partial differential equation (6.35) is equipped with the boundary conditions

$$
\widetilde{W}_{B}\left[\begin{array}{c}
\frac{\partial w}{\partial \partial}(0, t) \\
T \frac{\partial w}{\partial \zeta}(0, t)
\end{array}\right]=0 .
$$

The matrix $P_{1}$ can be factorized as

$$
P_{1}=\left[\begin{array}{cc}
1 & -1 \\
1 & 1
\end{array}\right]\left[\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right]\left[\begin{array}{cc}
1 / 2 & 1 / 2 \\
-1 / 2 & 1 / 2
\end{array}\right],
$$

This implies $d_{2}=1$. Thus, by Theorem 6.2.1 the corresponding operator

$$
\begin{aligned}
(A x)(\zeta) & =\left[\begin{array}{ll}
0 & 1 \\
1 & 0
\end{array}\right] \frac{\partial}{\partial \zeta}\left(\left[\begin{array}{cc}
\frac{1}{\rho(\zeta)} & 0 \\
0 & T(\zeta)
\end{array}\right] x(\zeta)\right) \\
D(A) & =\left\{x \in \mathcal{W}^{1,2}\left(0,1 ; \mathbb{C}^{2}\right) \mid \widetilde{W}_{B}(\mathcal{H} x)(0, t)=0\right\}
\end{aligned}
$$

generates a contraction $C_{0}$-semigroup on $L^{2}\left(0,1 ; \mathbb{C}^{2}\right)$ if and only if

$$
\widetilde{W}_{B}=\frac{b}{2}\left[\begin{array}{ll}
u-1 & u+1
\end{array}\right]
$$

for $b \in \mathbb{C} \backslash\{0\}$ and $u \in \mathbb{C}$. More precisely, the partial differential equation

$$
\begin{aligned}
& \frac{\partial^{2} w}{\partial t^{2}}(\zeta, t)=\frac{1}{\rho(\zeta)} \frac{\partial}{\partial \zeta}\left(T(\zeta) \frac{\partial w}{\partial \zeta}(\zeta, t)\right), \quad t \geqslant 0, \zeta \in[0, \infty), \\
& \quad(u-1) \frac{\partial w}{\partial t}(0, t)+(u+1) T(0) \frac{\partial w}{\partial \zeta}(0, t)=0, \quad t \geqslant 0, \\
& \rho(\zeta) \frac{\partial w}{\partial t}(\zeta, 0)=z_{0}(\zeta), \quad \zeta \geqslant 0 \\
& \frac{\partial w}{\partial \zeta}(\zeta, 0)=z_{1}(\zeta), \quad \zeta \geqslant 0
\end{aligned}
$$

where $u \in \mathbb{C}$ and $z_{0}, z_{1} \in L^{2}[0, \infty)$, possesses a unique solution satisfying

$$
\int_{0}^{\infty} \rho(\zeta)\left[\frac{\partial w}{\partial t}(\zeta, t)\right]^{2}+T(\zeta)\left[\frac{\partial w}{\partial \zeta}(\zeta, t)\right]^{2} d \zeta \leqslant \int_{0}^{\infty} \frac{z_{0}^{2}(\zeta)}{\rho(\zeta)}+T(\zeta) z_{1}^{2}(\zeta) d \zeta
$$

for $t>0$, which means that the energy of the system is non-increasing.

### 6.3 Closing remarks and open problems

The following example illustrates the connection between port-Hamiltonian systems in the infinite-dimensional setting and on the semi-axis.

Example 6.3.1. We consider the port-Hamiltonian operator

$$
\begin{gathered}
A_{\infty} x=\frac{d}{d \zeta} x \\
\mathcal{D}\left(A_{\infty}\right)=\mathcal{W}^{1,2}(0, \infty ; \mathbb{R})
\end{gathered}
$$

on the semi-axis, which generates a contraction $C_{0}$-semigroup on the space $X=$ $L^{2}(0, \infty ; \mathbb{R})$, namely the translation shift $(T(t))_{t \geqslant 0}$ with $(T(t) f)(\zeta)=f(\zeta+t)$. Since $X \simeq Y:=L^{2}\left(0,1 ; \ell^{2}\right)$ via the isomorphism $J$ given by

$$
J: L^{2}(0, \infty ; \mathbb{R}) \rightarrow L^{2}\left(0,1 ; \ell^{2}\right)
$$

$$
f(x) \mapsto g(x)=\left(\begin{array}{c}
g_{1}(\zeta) \\
\vdots \\
g_{n}(\zeta) \\
\vdots
\end{array}\right) \text { with } g_{n}(\zeta)=f(\zeta+(n-1))
$$



Figure 6.5: Sketch of the isomorphic map $J$
Therefore, we can conclude that $A g=J A_{\infty} J^{-1} g=\frac{d}{d \zeta} g$ with

$$
\begin{aligned}
\mathcal{D}(A)= & \left\{x \in H^{N}\left(0,1 ; \ell^{2}(\mathbb{N})\right) \mid \widetilde{W}_{B} \Phi(x)=0\right\} \\
= & \left\{g \in H^{1}\left(0,1 ; \ell^{2}(\mathbb{N})\right)\right. \text { with } \\
& \left.\left.g(\zeta)=\left(\begin{array}{c}
g_{1}(\zeta) \\
\vdots \\
g_{n}(\zeta) \\
\vdots
\end{array}\right) \right\rvert\, g_{n}(1)=g_{n+1}(0) \forall n \in \mathbb{N}\right\}
\end{aligned}
$$

generates also a contraction semigroup on $L^{2}\left(0,1 ; \ell^{2}\right)$. Since $A g$ is a portHamiltonian operator with $N=1, P_{0}=0$ and $P_{1}=I \in \mathcal{L}\left(\ell^{2}(\mathbb{N})\right)$ we may get the same result via Theorem 6.1.3.
On yields $\widetilde{W}_{B}=\left[\begin{array}{ll}2 I & -2 L\end{array}\right]$ and thus

$$
\begin{aligned}
W_{B} & =\left[\begin{array}{ll}
2 I & -2 L
\end{array}\right]\left[\begin{array}{cc}
Q & -Q \\
I & I
\end{array}\right]^{-1}=\left[\begin{array}{ll}
2 I & -2 L
\end{array}\right]\left[\begin{array}{cc}
I & I \\
-I & I
\end{array}\right]^{-1} \\
& =\left[\begin{array}{ll}
I & -L
\end{array}\right]\left[\begin{array}{cc}
I & I \\
-I & I
\end{array}\right]=\left[\begin{array}{ll}
I+L & I-L
\end{array}\right]=\left[\begin{array}{ll}
W_{1} & W_{2}
\end{array}\right],
\end{aligned}
$$

where
$L: \ell^{2}(\mathbb{N}) \rightarrow \ell^{2}(\mathbb{N}), L\left(x_{1}, x_{2}, \ldots\right) \mapsto\left(x_{2}, x_{3}, x_{4} \ldots\right)$ denotes the left shift and $R: \ell^{2}(\mathbb{N}) \rightarrow \ell^{2}(\mathbb{N}), R\left(x_{1}, x_{2}, \ldots\right) \mapsto\left(0, x_{1}, x_{2}, \ldots\right)$ denotes the right shift. Thus, we are in the situation of Example 6.1.18 and can verify using Theorem 6.1.3 that $A$ generates a contraction $C_{0}$-semigroup.
Remark 6.3.2. Furthermore, Jacob and Wegner give in JW19 a characterization of the generation of $C_{0}$-semigroups on the semi-axis. Results for more general networks of port-Hamiltonian systems are obtained in WW20. An open problem is still the characterization of (contraction) $C_{0}$-semigroups for port-Hamiltonian systems of higher order.

## Bibliography

[Ada75] R. A. Adams. Sobolev spaces. Academic Press, New York-London, 1975. Pure and Applied Mathematics, Vol. 65.
[AI95] S. A. Avdonin and S. A. Ivanov. Families of Exponentials. Cambridge University Press, Cambridge, 1995.
[AJ14] B. Augner and B. Jacob. Stability and stabilization of infinitedimensional linear port-Hamiltonian systems. Evol. Equ. Control Theory, 3:207-229, 2014.
[AK06] F. Albiac and N. J. Kalton. Topics in Banach space theory, volume 233 of Graduate Texts in Mathematics. Springer, New York, 2006.
[Aug16] B. Augner. Stabilisation of Infinite-dimensional Port-Hamiltonian Systems. PhD thesis, University of Wuppertal, 2016.
[BC63] R. Bellman and K. L. Cooke. Differential-difference equations. Academic Press, New York-London, 1963.
[BC16] G. Bastin and J.-M. Coron. Stability and Boundary Stabilization of 1-D Hyperbolic Systems, volume 88 of Progress in Nonlinear Differential Equations and their Applications. Birkhäuser/Springer, Cham, 2016.
[BK13] G. Berkolaiko and P. Kuchment. Introduction to Quantum Graphs, volume 186 of Mathematical Surveys and Monographs. American Mathematical Society, Providence, RI, 2013.
[Bre82] P. C. Breedveld. Thermodynamic bond graphs and the problem of thermal inertance. J. Franklin Inst., 314(1):15-40, 1982.
[CCD81] J. T. Cannon, J. J. Cannon, and S. Dostrovsky. The Evolution of Dynamics - Vibration Theory from 1687 to 1742. Springer New York, Berlin-Heidelberg, 1981.
[Cur84] R. F. Curtain. Spectral systems. Internat. J. Control, 39(4):657666, 1984.
[CZ95] R. F. Curtain and H. Zwart. An Introduction to InfiniteDimensional Linear Systems Theory. Springer-Verlag, New York, 1995.
[CZ20] R. F. Curtain and H. Zwart. Introduction to Infinite-Dimensional Systems Theory. Springer-Verlag, New York, 2020.
[DH20] S. Dubljevic and J.-P. Humaloja. Model predictive control for regular linear systems. Automatica J. IFAC, 119:109066, 9, 2020.
[DMSB09] V. Duindam, A. Macchelli, S. Stramigioli, and H. Bruyninckx, editors. Modeling and Control of Complex Physical Systems. Springer-Verlag, Berlin, 2009.
[DS71] N. Dunford and J. T. Schwartz. Linear Operators. Part III: Spectral Operators. Interscience Publishers [John Wiley \& Sons, Inc.], New York, 1971.
[DvdS99] M. Dalsmo and A. van der Schaft. On representations and integrability of mathematical structures in energy-conserving physical systems. SIAM J. Control Optim., 37(1):54-91, 1999.
[EKF19] K.-J. Engel and M. Kramar Fijavž. Waves and diffusion on metric graphs with general vertex conditions. Evol. Equ. Control Theory, 8(3):633-661, 2019.
[EMvdS07] D. Eberard, B. M. Maschke, and A. J. van der Schaft. An extension of Hamiltonian systems to the thermodynamic phase space: towards a geometry of nonreversible processes. Rep. Math. Phys., 60(2):175-198, 2007.
[EN00] K.-J. Engel and R. Nagel. One-Parameter Semigroups for Linear Evolution Equations, volume 194 of Graduate Texts in Mathematics. Springer-Verlag, New York, 2000.
[Eng13] K.-J. Engel. Generator property and stability for generalized difference operators. J. Evol. Equ., 13(2):311-334, 2013.
[GGK90] I. Gohberg, S. Goldberg, and M. A. Kaashoek. Classes of Linear Operators. Vol. I, volume 49 of Operator Theory: Advances and Applications. Birkhäuser Verlag, Basel, 1990.
[Gol85] J. A. Goldstein. Semigroups of linear operators and applications. Oxford Mathematical Monographs. The Clarendon Press, Oxford University Press, New York, 1985.
[GW19] B.-Z. Guo and J.-M. Wang. Control of Wave and Beam PDEs. Communications and Control Engineering Series. Springer, Cham, 2019. The Riesz Basis Approach.
[GX04] B.-Z. Guo and G.-Q. Xu. Riesz bases and exact controllability of $C_{0}$-groups with one-dimensional input operators. Systems Control Lett., 52(3-4):221-232, 2004.
[GZ01] B. Guo and H. Zwart. Riesz spectral systems. MEMORANDUM, No. 1594, 2001.
[Hei11] C. Heil. A basis theory primer. Applied and Numerical Harmonic Analysis. Birkhäuser/Springer, New York, expanded edition, 2011.
[Hil48] E. Hille. Functional Analysis and Semi-Groups. American Mathematical Society Colloquium Publications, vol. 31. American Mathematical Society, New York, 1948.
[HP05] D. Hinrichsen and A. J. Pritchard. Mathematical Systems Theory. I, volume 48 of Texts in Applied Mathematics. Springer-Verlag, Berlin, 2005.
[HP18] J.-P. Humaloja and L. Paunonen. Robust regulation of infinitedimensional port-Hamiltonian systems. IEEE Trans. Automat. Control, 63(5):1480-1486, 2018.
[JK19a] B. Jacob and J. T. Kaiser. On exact controllability of infinitedimensional linear port-Hamiltonian systems. IEEE Control Systems Letters, 3(3):661-666, 2019.
[JK19b] B. Jacob and J. T. Kaiser. Well-posedness of systems of 1-D hyperbolic partial differential equations. J. Evol. Equ., 19(1):91-109, 2019.
[JKZ20] B. Jacob, J. T. Kaiser, and H. Zwart. Riesz bases of portHamiltonian systems. Submitted, available at arXiv: 2009.08521, 2020.
[JMZ15] B. Jacob, K. Morris, and H. Zwart. $C_{0}$-semigroups for hyperbolic partial differential equations on a one-dimensional spatial domain. J. Evol. Equ., 15(2):493-502, 2015.
[JvdS09] D. Jeltsema and A. J. van der Schaft. Lagrangian and Hamiltonian formulation of transmission line systems with boundary energy flow. Rep. Math. Phys., 63(1):55-74, 2009.
[JW19] B. Jacob and S.-A. Wegner. Well-posedness of a class of hyperbolic partial differential equations on the semi-axis. J. Evol. Equ., 19(4):1111-1147, 2019.
[JZ99] B. Jacob and H. Zwart. Equivalent conditions for stabilizability of infinite-dimensional systems with admissible control operators. SIAM J. Control Optim., 37(5):1419-1455, 1999.
[JZ01a] B. Jacob and H. Zwart. Exact observability of diagonal systems with a finite-dimensional output operator. Systems Control Lett., 43(2):101-109, 2001.
[JZ01b] B. Jacob and H. Zwart. Exact observability of diagonal systems with a one-dimensional output operator. Int. J. Appl. Math. Comput. Sci., 11(6):1277-1283, 2001.
[JZ12] B. Jacob and H. J. Zwart. Linear port-Hamiltonian Systems on Infinite-Dimensional Spaces, volume 223. Birkhäuser/Springer Basel AG, Basel, 2012. Linear Operators and Linear Systems.
[JZ18] B. Jacob and H. Zwart. An operator theoretic approach to infinitedimensional control systems. GAMM-Mitt., 41(4):e201800010, 14, 2018.
[JZ21] B. Jacob and H. Zwart. Observability for port-Hamiltonian systems. Submitted at 2021 European Control Conference (ECC), 2021.
[Kat95] T. Kato. Perturbation theory for linear operators. Classics in Mathematics. Springer-Verlag, Berlin, 1995. Reprint of the 1980 edition.
[Kom94] V. Komornik. Exact Controllability and Stabilization. RAM: Research in Applied Mathematics. Masson, Paris; John Wiley \& Sons, Ltd., Chichester, 1994.
[KPS08] V. Kostrykin, J. Potthoff, and R. Schrader. Contraction semigroups on metric graphs. In Analysis on graphs and its applications, volume 77 of Proc. Sympos. Pure Math., pages 423-458. Amer. Math. Soc., Providence, RI, 2008.
[KS99] V. Kostrykin and R. Schrader. Kirchhoff's rule for quantum wires. J. Phys. A, 32(4):595-630, 1999.
[KZ15] M. Kurula and H. Zwart. Linear wave systems on $n$ - $D$ spatial domains. Internat. J. Control, 88(5):1063-1077, 2015.
[Lag11] J. Lagrange. Mécanique analytique. Number 1 in Mécanique analytique. Ve Courcier, 1811.
[Lan12] C. Lanczos. The Variational Principles of Mechanics. University of Toronto Press, New York, 2012.
[LGZM05] Y. Le Gorrec, H. Zwart, and B. Maschke. Dirac structures and boundary control systems associated with skew-symmetric differential operators. SIAM J. Control Optim., 44:1864-1892, 2005.
[LP61] G. Lumer and R. S. Phillips. Dissipative operators in a Banach space. Pacific J. Math., 11:679-698, 1961.
[LW83] J.-C. Louis and D. Wexler. On exact controllability in Hilbert spaces. J. Differential Equations, 49(2):258-269, 1983.
[MM05] A. Macchelli and C. Melchiorri. Control by interconnection of mixed port Hamiltonian systems. IEEE Trans. Automat. Control, 50(11):1839-1844, 2005.
[MNS18] D. Mugnolo, D. Noja, and C. Seifert. Airy-type evolution equations on star graphs. Anal. PDE, 11(7):1625-1652, 2018.
[Mug14] D. Mugnolo. Semigroup Methods for Evolution Equations on Networks. Understanding Complex Systems. Springer, Cham, 2014.
[New87] I. Newton. Philosophiae naturalis principia mathematica. William Dawson \& Sons Ltd., London, 1687.
[Paz83] A. Pazy. Semigroups of linear operators and applications to partial differential equations, volume 44 of Applied Mathematical Sciences. Springer-Verlag, New York, 1983.
[RLGMZ14] H. Ramírez, Y. Le Gorrec, A. Macchelli, and H. Zwart. Exponential stabilization of boundary controlled port-Hamiltonian systems with dynamic feedback. IEEE Trans. Automat. Control, 59(10):2849-2855, 2014.
[RW94] D. L. Russell and G. Weiss. A general necessary condition for exact observability. SIAM J. Control Optim., 32(1):1-23, 1994.
[RW97] R. Rebarber and G. Weiss. An extension of Russell's principle on exact controllability. European Control Conference (ECC), 1997.
[RZLG17] H. Ramírez, H. Zwart, and Y. Le Gorrec. Stabilization of infinite dimensional port-Hamiltonian systems by nonlinear dynamic boundary control. Automatica J. IFAC, 85:61-69, 2017.
[Sch27] J. Schauder. Zur Theorie stetiger Abbildungen in Funktionalräumen. Mathematische Zeitschrift, 26(1):47-65, 1927.
[SSVW15] C. Schubert, C. Seifert, J. Voigt, and M. Waurick. Boundary systems and (skew-)self-adjoint operators on infinite metric graphs. Math. Nachr., 288(14-15):1776-1785, 2015.
[Sta05] O. Staffans. Well-Posed Linear Systems, volume 103 of Encyclopedia of Mathematics and its Applications. Cambridge University Press, Cambridge, 2005.
[SZ18] J. Schmid and H. Zwart. Stabilization of port-Hamiltonian systems by nonlinear boundary control in the presence of disturbances. Proc. Mathematical Theory of Networks and Systems (MTNS 2018), Hong Kong, China, pages 570-575, 2018.
[Tre00a] C. Tretter. Spectral problems for systems of differential equations $y^{\prime}+A_{0} y=\lambda A_{1} y$ with $\lambda$-polynomial boundary conditions. Math. Nachr., 214:129-172, 2000.
[Tre00b] C. Tretter. Linear operator pencils $A-\lambda B$ with discrete spectrum. Integral Equations Operator Theory, 37(3):357-373, 2000.
[Tri91] R. Triggiani. Lack of exact controllability for wave and plate equations with finitely many boundary controls. Differential Integral Equations, 4(4):683-705, 1991.
[TW09] M. Tucsnak and G. Weiss. Observation and Control for Operator Semigroups. Birkhäuser Advanced Texts: Basler Lehrbücher. Birkhäuser Verlag, Basel, 2009.
[TW14] M. Tucsnak and G. Weiss. Well-posed systems - the LTI case and beyond. Automatica J. IFAC, 50(7):1757-1779, 2014.
[vdS06] A. van der Schaft. Port-Hamiltonian systems: an introductory survey. In International Congress of Mathematicians. Vol. III, pages 1339-1365. Eur. Math. Soc., Zürich, 2006.
[vdSM02] A. J. van der Schaft and B. M. Maschke. Hamiltonian formulation of distributed-parameter systems with boundary energy flow. J. Geom. Phys., 42(1-2):166-194, 2002.
[Vil07] J. Villegas. A Port-Hamiltonian Approach to Distributed Parameter Systems. PhD thesis, University of Twente, Netherlands, 2007.
[Was87] W. Wasow. Asymptotic Expansions for Ordinary Differential Equations. Dover Publications, Inc., New York, 1987. Reprint of the 1976 edition.
[Wei88] G. Weiss. Admissibility of input elements for diagonal semigroups on $l^{2}$. Systems Control Lett., 10(1):79-82, 1988.
[Wei94] G. Weiss. Regular linear systems with feedback. Math. Control Signals Systems, 7(1):23-57, 1994.
[Wer00] D. Werner. Funktionalanalysis. Springer-Verlag, Berlin, extended edition, 2000.
[WW20] M. Waurick and S.-A. Wegner. Dissipative extensions and portHamiltonian operators on networks. Submitted, available at arXiv: 2009.08521, 2020.
[XF02] G.-Q. Xu and D.-X. Feng. The Riesz basis property of a Timoshenko beam with boundary feedback and application. IMA J. Appl. Math., 67(4):357-370, 2002.
[XG03] G.-Q. Xu and B.-Z. Guo. Riesz basis property of evolution equations in Hilbert spaces and application to a coupled string equation. SIAM J. Control Optim., 42(3):966-984, 2003.
[XW11] C.-Z. Xu and G. Weiss. Eigenvalues and eigenvectors of semigroup generators obtained from diagonal generators by feedback. Commun. Inf. Syst., 11(1):71-104, 2011.
[Yos48] K. Yosida. On the differentiability and the representation of oneparameter semi-group of linear operators. J. Math. Soc. Japan, 1:15-21, 1948.
[You80] R. M. Young. An Introduction to nonharmonic Fourier Series, volume 93 of Pure and Applied Mathematics. Academic Press, Inc., New York-London, 1980.
[ZLMV10] H. Zwart, Y. Le Gorrec, B. Maschke, and J. Villegas. Wellposedness and regularity of hyperbolic boundary control systems on a one-dimensional spatial domain. ESAIM Control Optim. Calc. Var., 16(4):1077-1093, 2010.
[Zwa10] H. Zwart. Riesz basis for strongly continuous groups. J. Differential Equations, 249(10):2397-2408, 2010.

## Acknowledgement

First of all I would like to express my deepest gratitude to my supervisor Prof. Dr. Birgit Jacob for all her support during my studies and my doctorate. She gave me the opportunity to work in the field of port-Hamiltonian systems. I am much obliged to Prof. Dr. Hans Zwart for all the interesting discussions in the field of Riesz bases, port-Hamiltonian systems and beyond and for being second reviewer of this thesis.

For the very friendly working atmosphere I thank all members of the research group functional analysis, my colleagues and former colleagues, Prof. Dr. Bálint Farkas, Dr. Henrik Kreidler, Dr. Jens Wintermayr, Dr. Christian Wyss, René Hosfeld, Sebastian Möller, Merlin Schmitz, Nathanael Skrepek, Lukas Vorberg, Dr. Vincent Andrieu, Dr. Björn Augner, Dr. Christian Budde, Dr. Waed Dada, Dr. Hafida Laasri, Dr. Robert Nabiullin, Dr. Felix Schwenninger, Dr. Sven-Ake Wegner.

I would like to thank Kerstin Scheibler for her friendly help since my first day in the research group functional analysis.
Furthermore, I thankfully acknowledge partial financial support by the German Research Foundation (DFG) within Grants JA 735/8-1 and JA 735/13-1. Last but not least, I want to thank my family and my friends for supporting me. Finally, I thank my husband Markus for his support, his love, and his patience.

