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# Minimal null control time of some 1D hyperbolic balance laws with constant coefficients and properties of related kernel equations

Long Hu<sup>\*</sup> Guillaume Olive<sup>†</sup>

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#### Abstract

In this work, we study the null controllability by one-sided boundary controls of onedimensional hyperbolic balance laws with constant coefficients. Our first result shows that, when the system has only one negative or positive speed, the minimal null control time 9 of such systems depends on some orthogonality conditions for a particular sequence. This 10 sequence is explicit in function of the coefficients of the system but it is defined by a nonlinear 11 recurrence relation. Our second result then completes the previous one by giving explicit 12 bounds on the number of orthogonality conditions that have to be checked in two nontrivial 13 situations. The proofs rely on a careful analysis of the so-called kernel equations associated 14 with the system, including a new well-posedness result. Our results are also valid for the 15 finite-time stabilization property. 16

Keywords. Hyperbolic systems; Minimal control time; Backstepping method; Nonlinear
 recurrence relation

**2010 Mathematics Subject Classification.** 35L40, 93B05, 93D15, 11B

#### <sup>20</sup> 1 Introduction and main results

#### <sup>21</sup> 1.1 Problem description

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In this paper, we are interested in the null controllability properties of a class of one-dimensional
(1D) hyperbolic system with constant coefficients (see e.g. [BC16, Chap. 1] for applications).
The equations describing such phenomenons are

$$\frac{\partial y}{\partial t}(t,x) + \Lambda \frac{\partial y}{\partial x}(t,x) = My(t,x).$$
(1a)

Above,  $t \in (0,T)$  is the time variable, T > 0,  $x \in (0,1)$  is the space variable and the state is  $y: (0,T) \times (0,1) \to \mathbb{R}^n$   $(n \ge 2)$ . The matrix  $\Lambda \in \mathbb{R}^{n \times n}$  will always be assumed diagonal  $\Lambda = \operatorname{diag}(\lambda_1, \ldots, \lambda_n)$ , with  $m \ge 1$  negative speeds and  $p \ge 1$  positive speeds (m + p = n):

$$\lambda_1 < \dots < \lambda_m < 0 < \lambda_{m+1} < \dots < \lambda_{m+p}.$$
(1b)

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The matrix  $M \in \mathbb{R}^{n \times n}$  couples the equations of the system inside the domain and will be called 30 the internal coupling matrix. We will consider an initial condition at time t = 0: 31

$$y(0,x) = y^0(x).$$
 (1c)

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Let us now discuss the boundary conditions. The structure of  $\Lambda$  induces a natural splitting 33 of the state into components corresponding to negative and positive speeds, denoted respectively 34 by  $y_{-}$  and  $y_{+}$ . For the above system to be well-posed in  $(0,T) \times (0,1)$  with an initial condition 35 at time t = 0, we then need to add boundary conditions at x = 1 for  $y_{-}$  and at x = 0 for  $y_{+}$ . We 36 will consider the following type of boundary conditions, motivated by its numerous applications 37 (see again [BC16]): 38 y

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$$y_{-}(t,1) = u(t), \quad y_{+}(t,0) = Qy_{-}(t,0).$$
 (1d)

The function u is the so-called control, it will be at our disposal. It only acts on one part of 40 the boundary and, on the other part of the boundary, the equations are coupled by the matrix 41  $Q \in \mathbb{R}^{p \times m}$ . This matrix will be called the boundary coupling matrix. In what follows, (1a), (1c) 42 and (1d) together will be referred to as system (1). 43

We recall that system (1) is well-posed in the following functional setting: for every T > 0,  $y^0 \in L^2(0,1)^n$  and  $u \in L^2(0,T)^m$ , there exists a unique solution y to system (1) with

$$y \in C^0([0,T]; L^2(0,1)^n) \cap C^0([0,1]; L^2(0,T)^n).$$

By solution we mean "solution along the characteristics". We refer for instance to [Cor+21] for a 44

proof of this well-posedness result in such a setting (see also [BC16, Appendix A] when u = 0). 45

The regularity  $C^0([0,T]; L^2(0,1)^n)$  of the solution allows us to consider control problems in 46 the space  $L^{2}(0,1)^{n}$ : 47

**Definition 1.1.** Let T > 0 be fixed. We say that system (1) is null controllable in time T if, for every  $y^0 \in L^2(0,1)^n$ , there exists  $u \in L^2(0,T)^m$  such that the corresponding solution y to system (1) satisfies

$$y(T, \cdot) = 0$$

Since controllability in time  $T_1$  implies controllability in any time  $T_2 \ge T_1$ , it is natural to 48 try to find the smallest possible control time, the so-called "minimal control time". 49

**Definition 1.2.** For any  $\Lambda, M$  and Q as above, we denote by  $T_{inf}(\Lambda, M, Q) \in [0, +\infty]$  the 50 minimal null control time of system (1), that is 51

$$T_{\inf}(\Lambda, M, Q) = \inf \{T > 0 \mid \text{System (1) is null controllable in time } T \}.$$
(2)

The time  $T_{inf}(\Lambda, M, Q)$  is named "minimal" null control time according to the current litera-53

ture, despite it is not always a minimal element of the set. We keep this naming here, but we use 54

the notation with the "inf" to avoid eventual confusions. The goal of this article is to characterize 55

 $T_{inf}(\Lambda, M, Q)$  in some new situations. 56

In order to state our results and those of the literature, we need to introduce the following times:

$$T_i = \frac{1}{-\lambda_i}$$
 if  $i \le m$ ,  $T_i = \frac{1}{\lambda_i}$  if  $i \ge m+1$ .

The time  $T_i$  is the time needed for the controllability of a single equation (the transport equation) 57 with speed  $\lambda_i$ . Note that the assumption (1b) implies in particular the following order relation: 58

 $\begin{cases} T_1 \leq \cdots \leq T_m, \\ T_n \leq \cdots \leq T_{m+1}. \end{cases}$ 

(3)

#### 60 1.2 Literature

Here, we briefly describe the results of the literature about the null controllability of system (1).
All the results below are also valid for space-dependent versions of this system.

• It was first proved in the celebrated survey [Rus78] that system (1) is null controllable in any time  $T \ge T_{m+1} + T_m$ . A strength of this result is that it is valid for any M and Q. However, it was also observed in that paper that the minimal control time can be smaller than  $T_{m+1} + T_m$ . Finding the minimal control time even in the simpler case M = 0 was then left as an open problem.

- For M = 0, the minimal null control time was eventually found in [Wec82]. The author gave an explicit expression of this time in terms of some indices related to Q.
- Finding the minimal null control time for arbitrary M and Q is still an open challenging problem. There has been a recent resurgence on the characterization of this time. A first result in this direction was obtained in [CN19] and then completed in [CN21a]. Therein, the authors introduced a class of boundary coupling matrices Q for which they showed that the minimal null control time is smaller than  $T_{m+1} + T_m$ , whatever the internal coupling matrix M is.
- For full row rank boundary coupling matrices (rank Q = p), the minimal null control time was found in [HO21a]. In this case, it has been shown that this time is the same as for the system without internal coupling (M = 0).
- For systems of n = 2 equations, the minimal null control time was found in [Cor+13] and [HO21b]. In particular, it has been shown in the second reference that this time depends on the internal coupling matrix M when the boundary coupling matrix is zero. This is a feature that was not observed nor highlighted in all the other works and that the results of the present paper will also share.
- Finally, the smallest and largest values that the minimal null control time can take with respect to the internal coupling matrix M were found in [HO22].

Other related works include for instance [Cor+21; CN21b; MA22] about time-dependent versions of system (1) and [Li10; LR10; Hu15; CN20; CN22] for quasilinear versions of this system (in a  $C^1$  framework).

#### <sup>89</sup> 1.3 Notations and important definitions

<sup>90</sup> To state the main results of this article we first need to introduce some notations.

All along this article, we denote by  $A^{\mathsf{T}}$  the transpose of a matrix A. For any integer  $N \geq 1$ ,  $\mathbb{R}_{0}^{N \times N}$  denotes the set of matrices of size  $N \times N$  with diagonal entries all equal to zero. The matrix  $\mathrm{Id}_{N}$  denotes the  $N \times N$  identity matrix. A matrix (or matrix-valued function) of size  $N_{1} \times N_{2}$ will simply be denoted using the corresponding lowercase letter when  $N_{2} = 1$  (e.g.  $Q \in \mathbb{R}^{p \times m}$ will be denoted by  $q \in \mathbb{R}^{p}$  when m = 1). The inner product of two vectors  $v_{1}, v_{2} \in \mathbb{R}^{n-1}$  will be denoted by  $\langle v_{1}, v_{2} \rangle$ .

Let us now introduce a sequence that will play a key role throughout this paper. For any  $i \in \{1, ..., n\}$ , we first define the following quantities.

• For every  $r, j \in \{1, \ldots, n\}$ , we denote by

$$\alpha_{rj} = \frac{m_{rj}}{\lambda_i - \lambda_j}$$
 if  $j \neq i$ ,  $\alpha_{ri} = \frac{m_{ri}}{\lambda_i}$ .

• Let then  $A, D \in \mathbb{R}^{(n-1) \times (n-1)}$  be the matrices defined by

$$A = (\alpha_{rj})_{r,j \neq i}, \quad D = \text{diag} \left(\frac{\lambda_i + \lambda_j}{\lambda_i - \lambda_j}\right)_{j \neq i}$$

and let  $\psi^0, w \in \mathbb{R}^{n-1}$  be the vectors defined by

$$\psi^0 = (\alpha_{ij})_{j \neq i}, \quad v = \left(-\frac{1}{2}\alpha_{ji}\right)_{j \neq i}.$$

• Let  $(\psi^r)_{r\in\mathbb{N}}\subset\mathbb{R}^{n-1}$  be the sequence defined by  $\psi^0$  and

$$\psi^{1} = A^{\mathsf{T}}\psi^{0}, \quad \psi^{r} = A^{\mathsf{T}}\psi^{r-1} + D\sum_{\ell=0}^{r-2} \langle v, \psi^{r-2-\ell} \rangle \psi^{\ell}, \quad \forall r \ge 2.$$
(4)

• Finally, for  $q \in \mathbb{R}^{n-1}$ , let  $b \in \mathbb{R}^{n-1}$  be the nonzero vector defined by

$$b = (\beta_j)_{j \neq i}, \quad \beta = -\Lambda \begin{pmatrix} 1\\ q \end{pmatrix}.$$
 (5)

All the above quantities depend on the index i but we omit it for clarity.

With the previous notations, we introduce the following sets. For any  $k \in \{2, \ldots, n+1\}$ , we denote by  $\mathcal{C}_k$  the set of  $(M,q) \in \mathbb{R}^{n \times n}_0 \times \mathbb{R}^{n-1}$  such that, for every  $2 \leq i < k$ , we have  $q_{i-1} = 0$ and

$$\langle b, \psi^r \rangle = 0, \quad \forall r \in \mathbb{N}.$$
 (6)

Here, we use the convention that  $C_2 = \mathbb{R}_0^{n \times n} \times \mathbb{R}^{n-1}$ . Additionally, we will denote by  $C_{n+2} = \emptyset$ . Note that we then have  $C_2 \supset C_3 \supset \cdots \supset C_{n+1} \supset C_{n+2}$ .

#### 110 1.4 Main results and comments

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The first result of this article is the following characterization of the minimal null control time in the case of one negative speed.

**Theorem 1.3.** Assume that m = 1. Let us denote by

$$\tau_i = \max\{T_1 + T_i, T_2\} \quad if \ 2 \le i \le n, \quad \tau_{n+1} = \max\{T_1, T_2\}, \tag{7}$$

(we have  $\tau_2 \ge \tau_3 \ge \cdots \ge \tau_{n+1}$  from (3)). Then, for any  $M \in \mathbb{R}^{n \times n}_0$  and  $q \in \mathbb{R}^{n-1}$ , we have:

116 1.  $T_{inf}(\Lambda, M, q) \in \{\tau_2, \ldots, \tau_{n+1}\}$ . Moreover, the infimum is always reached (in (2)).

2. For any  $k \in \{2, ..., n+1\}$ , we have

$$T_{\inf}(\Lambda, M, q) = \tau_k \quad \Longleftrightarrow \quad (M, q) \in \mathcal{C}_k \setminus \mathcal{C}_{k+1}$$

We recall that the set  $C_k$  is defined at the end of Section 1.3.

**Remark 1.4.** Theorem 1.3 solves the open problem raised at the end of [HO21b, Section 5] for systems with constant coefficients.

**Remark 1.5.** Theorem 1.3 remains valid if we replace everywhere in this article the null controllability property by the finite-time stabilization property by  $L^2$  bounded feedbacks (that is when the control u is looked under the more particular form  $u(t) = \int_0^1 r(\xi)^{\mathsf{T}} y(t,\xi) d\xi$  with  $r \in L^2(0,1)^n$ ). This easily follows from the proof below.

Even though the set  $C_k$  is explicit in function of the parameters of the system, the orthogonality conditions (6) that define this set can be difficult to study in general because the sequence  $(\psi^r)_{r\in\mathbb{N}}$  is defined by a nonlinear recurrence relation. Note however that there always exists an integer  $N \ge 1$  such that

$$\psi^r \in \text{span} \{\psi^s \mid s \leq N-1\}, \quad \forall r \geq N,$$

124 so that

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 $\langle b, \psi^r \rangle = 0, \quad \forall r \in \mathbb{N} \quad \iff \quad \langle b, \psi^r \rangle = 0, \quad \forall r \in \{0, \dots, N-1\}.$  (8)

This means that (6) only needs to be checked for the first N values of r. However, such a Ndepends on the sequence and it is a priori unknown, so that, in practice, we do not know when we have to stop checking the orthogonality conditions. Our second result provides information on this issue in two particular cases:

**Theorem 1.6.** Let  $i \in \{1, \ldots, n\}$  be fixed. Define

$$N_{\psi} = \min\{N \ge 1 \mid (8) \ holds\}.$$

130 We have  $N_{\psi} \leq 3$  for n = 3 and  $N_{\psi} \leq 6$  for n = 4.

**Remark 1.7.** It would be interesting to find a bound of  $N_{\psi}$  with respect to *n* for arbitrary *n*.

Our main results can for instance be combined to deduce a very explicit characterization of the minimal null control time in the following particular case:

**Corollary 1.8.** Assume that m = 1 and p = 2. Then, for any  $M \in \mathbb{R}_0^{3 \times 3}$  and  $q \in \mathbb{R}^2$ , we have:

1.  $T_{inf}(\Lambda, M, q) = \max\{T_1, T_2\}$  if, and only if, (M, q) satisfies

$$q = 0, \quad m_{21} = m_{31} = 0.$$

2.  $T_{inf}(\Lambda, M, q) = \max\{T_1 + T_3, T_2\}$  if, and only if, (M, q) satisfies

$$q = 0, \quad m_{21} = m_{23} = 0, \quad m_{31} \neq 0,$$

or

$$q_1 = 0, \quad q_2 \neq 0 \quad and \quad \left( m_{21} = m_{23} = 0 \quad or \quad \begin{cases} m_{21} = rsm_{23}, \\ m_{31} = r^2sm_{13}, \\ m_{32} = -rm_{12}, \end{cases} \right),$$

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36 3. In all the other situations, 
$$T_{inf}(\Lambda, M, q) = T_1 + T_2$$
.

where  $r = -\frac{\lambda_3 q_2}{\lambda_1}$  and  $s = \frac{\lambda_2 - \lambda_1}{\lambda_2 - \lambda_3}$ .

For p = 3, there is no simple presentation as for p = 2, even though the orthogonality conditions are explicit (see also Remark 1.10 below) and we know that we only have to check the first six conditions. Therefore, we only give a nontrivial example: **Example 1.9.** Let  $\sigma \in \mathbb{R} \setminus \{0\}$  be arbitrary and consider system (1) with

$$\Lambda = \operatorname{diag} (-1, 1, 2, 3), \quad M = \begin{pmatrix} 0 & -3 & 1/(2\sigma) & 0 \\ 2 & 0 & 0 & -2 \\ 3\sigma & 0 & 0 & -\sigma \\ 0 & 2 & 1/(2\sigma) & 0 \end{pmatrix}, \quad q = \begin{pmatrix} 0 \\ 0 \\ 1/3 \end{pmatrix}.$$

• For i = 2, we have  $b = \begin{pmatrix} 1 & 0 & -1 \end{pmatrix}^{\mathsf{T}}$  and

$$\psi^{0} = \begin{pmatrix} 1\\0\\1 \end{pmatrix}, \quad \psi^{1} = -\frac{1}{\sigma} \begin{pmatrix} 0\\1\\0 \end{pmatrix}, \quad \psi^{2} = -\frac{3}{2}\psi^{0}, \quad \psi^{3} = -3\psi^{1}, \quad \psi^{4} = \frac{9}{2}\psi^{0}, \quad \psi^{5} = \frac{45}{4}\psi^{1}.$$

• For i = 3, we have the same b and

$$\psi^0 = \sigma \begin{pmatrix} 1\\0\\1 \end{pmatrix}, \quad \psi^1 = -\sigma \begin{pmatrix} 0\\1\\0 \end{pmatrix}, \quad \psi^2 = -\frac{3}{4}\psi^0, \quad \psi^3 = -\frac{3}{2}\psi^1, \quad \psi^4 = \frac{9}{8}\psi^0, \quad \psi^5 = \frac{45}{16}\psi^1.$$

From Theorems 1.3 and 1.6, we deduce that  $T_{inf}(\Lambda, M, q) = \tau_4 = 4/3$ .

**Remark 1.10.** For arbitrary n, we will see that the orthogonality conditions (6) are satisfied if one of the following three conditions holds:

(C1) Kal 
$$(A, v)^{\mathsf{T}} \psi^0 = \text{Kal} (A, b)^{\mathsf{T}} \psi^0 = 0$$
, where Kal  $(A, h) = (h|Ah| \cdots |A^{n-2}h) \in \mathbb{R}^{(n-1) \times (n-1)}$   
denotes the Kalman matrix of  $(A, h)$ , for any  $h \in \mathbb{R}^{n-1}$ .

(C2) There exists  $\emptyset \neq J \subsetneq \{1, \ldots, n-1\}$  such that  $\psi_j^0 = a_{rj} = b_r = 0$  for every  $j \notin J$  and  $r \in J$ .

(C3)  $\langle b, \psi^0 \rangle = 0$  and there exists  $j_0 \in \{1, \ldots, n-1\}$  such that  $b_{j_0} = 0$  and rank  $\Delta_{j_0} = 1$ , where

$$\Delta_{j_0} = \begin{pmatrix} (D - d_{j_0})\psi^0 & A^{\mathsf{T}}e_{j_0} \\ A^{\mathsf{T}}\psi^0 & v_{j_0}\psi^0 - \langle v, \psi^0 \rangle e_{j_0} \end{pmatrix} \in \mathbb{R}^{2(n-1)\times 2},$$

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where  $d_{j_0}$  is the  $j_0$ -th diagonal entry of D and  $e_{j_0}$  is the  $j_0$ -th canonical vector of  $\mathbb{R}^{n-1}$ .

We will also see that, for n = 3 (resp. n = 4), it is necessary that one of the conditions (C1), (C2) (resp. (C1), (C2), (C3)) holds (it is however preferable to use Theorem 1.6 in these cases).

The rest of this article is organized as follows. In Section 2, we use the equivalence between 149 the controllability of system (1) and that of a simpler system to obtain a characterization of 150 this property in terms of some orthogonality conditions for the derivatives at the origin of any 151 solution to the so-called kernel equations. In Section 3, we compute these derivatives for a 152 particular solution and we obtain a general formula for this solution. In Section 4, we study the 153 orthogonality conditions associated with the previous solution and we deduce our main results. In 1 54 Section 5, we supplement ours results by studying the structure of the solution associated with the 155 orthogonality conditions. Finally, in Appendix A, we give a simple proof of the characterization 156 of the controllability properties for the equivalent system and, in Appendix B, we prove the 157 existence of a solution to the kernel equations by a new approach. 158

#### <sup>159</sup> 2 An equivalent system and the kernel equations

The first step in the proof of our results is to consider a system which is equivalent to our initialsystem from a control point of view.

**Lemma 2.1.** For any T > 0, system (1) is null controllable in time T if, and only if, so is the system

$$\begin{cases} \frac{\partial \tilde{y}}{\partial t}(t,x) + \Lambda \frac{\partial \tilde{y}}{\partial x}(t,x) = F(x)\tilde{y}_{-}(t,0), \\ \tilde{y}_{-}(t,1) = \tilde{u}(t), \quad \tilde{y}_{+}(t,0) = Q\tilde{y}_{-}(t,0), \\ \tilde{y}(0,x) = \tilde{y}^{0}(x), \end{cases}$$
(9)

165 where  $F \in C^0([0,1])^{n \times m}$  is defined by

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$$F(x) = -K(x,0)\Lambda \begin{pmatrix} \mathrm{Id}_m \\ Q \end{pmatrix}, \tag{10}$$

167 and  $K \in C^0(\overline{\mathcal{T}})^{n \times n}$  is any solution to

$$\begin{cases} \Lambda \frac{\partial K}{\partial x}(x,\xi) + \frac{\partial K}{\partial \xi}(x,\xi)\Lambda + K(x,\xi)M = 0, \\ \Lambda K(x,x) - K(x,x)\Lambda = M, \end{cases}$$
(11)

in the closure of the triangle  $\mathcal{T} = \{(x,\xi) \in \mathbb{R}^2 \mid 0 < \xi < x < 1\}.$ 

By solution to (11) we mean solution along the characteristics. This result is by now wellknown: it consists in using the invertible spatial transformation

$$\tilde{y}(t,x) = y(t,x) - \int_0^x K(x,\xi)y(t,\xi)\,d\xi,$$

in order to transform a solution of system (1) into a solution of system (9) (see e.g. [Hu+19, Section 2.2]). This idea is the starting point of the so-called backstepping method for partial differential equations and introduced more specifically for hyperbolic systems of two equations in [Cor+13]. Equations (11) are thus called the kernel equations. The difficult point is not so much the result of the previous lemma but rather to prove that (11) actually has at least a solution. It follows from the results of [Hu+16] that there are many solutions to the kernel equations (11) in  $\overline{T}$ .

Remark 2.2. The choice of solution to the kernel equations (11) does not affect the controllability properties of system (1) because all the corresponding systems (9)-(10) are equivalent from a control point of view.

- 180 Now, two problems naturally arise:
- 1. Can we characterize the null controllability of the equivalent system (9) in function of  $\Lambda, Q$ and F?
- 2. If so, can this characterization be presented explicitly in terms of  $\Lambda$ , Q and M?

These problems are still open in general. One particular case where the first problem has been completely solved is the case m = 1 (one negative speed). This was done in [HO21b, Section 5]. **Theorem 2.3.** Assume that m = 1. Then, system (9) is null controllable in time T if, and only if,

$$T \ge \max_{2 \le i \le n} \{T_1 + T_i^*, T_2\},\$$

where

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$$T_i^* = \begin{cases} T_i & \text{if } q_{i-1} \neq 0, \\ T_i(1 - \ell(f_i)) & \text{if } q_{i-1} = 0, \end{cases}$$

where  $\ell(f_i) = \sup I(f_i)$  with  $I(f_i) = \{\ell \in (0,1) \mid f_i = 0 \text{ in } (0,\ell)\}$ , if  $I(f_i) \neq \emptyset$ , and  $\ell(f_i) = 0$ otherwise.

The second problem could not be solved though because, even if the conditions for f are explicit, the "map"  $M \mapsto f$  ("defined" by (11)-(10), with m = 1) is quite complicated. It was left as an open problem in the same paper. This is precisely where our main results step in.

From the above result of [HO21b] we see that the values at x = 0 of f and its derivatives (assuming it is smooth) can affect the minimal null control time  $T_{inf}(\Lambda, -, q, f)$  of the system  $(T_i^* = T_i \text{ if } f_i^{(N)}(0) \neq 0 \text{ for some } N \geq 0)$ . Our idea is to show that these values in fact completely characterize  $T_{inf}(\Lambda, -, q, f)$  because M is constant and that we can explicitly relate them to Mthanks to the kernel equations.

It is clear that  $T_{inf}(\Lambda, -, q, f)$  is solely characterized by f(0), f'(0), f''(0), etc. if we have

$$f$$
 is analytic in a neighborhood of  $[0, 1)$ . (12)

<sup>198</sup> Under such an assumption, Theorem 2.3 takes a simpler form:

**Corollary 2.4.** Assume that m = 1, let  $q \in \mathbb{R}^{n-1}$  be given and assume (12). Then, we have:

200 1.  $T_{inf}(\Lambda, -, q, f) \in \{\tau_2, \dots, \tau_{n+1}\}$  (recall (7)). Moreover, the infimum is always reached.

2. For any  $k \in \{2, ..., n+1\}$ , we have

$$T_{\inf}(\Lambda, -, q, f) = \tau_k \quad \Longleftrightarrow \quad (q, f) \in \mathcal{S}_k \setminus \mathcal{S}_{k+1}$$

where, for every  $k \in \{2, ..., n+1\}$ ,  $S_k$  is the set of  $(q, f) \in \mathbb{R}^{n-1} \times C^0([0,1])^n$  such that  $q_{i-1} = f_i = 0$  for every  $2 \le i < k$  (we use the convention that  $S_2 = \mathbb{R}^{n-1} \times C^0([0,1])^n$ ), and  $S_{n+2} = \emptyset$ .

This result is immediate from the previous theorem but we give a simple and direct proof in Appendix A. Note that it is the complete analogue of Theorem 1.3 for system (9). By Lemma 206 2.1, the minimal null control time for the initial system (1) is thus also completely determined 207 by the sets  $S_k$ . However, apart from  $S_{n+1}$ , these sets are not explicit in terms of M, which is 208 unsatisfactory.

Assumption (12) is indeed satisfied in our framework because we can always find an analytic solution to the kernel equations since M is constant. More precisely, we have the following result:

**Theorem 2.5.** Let  $m, p \ge 1$  be arbitrary. Assume that  $M \in \mathbb{R}_0^{n \times n}$ . For any  $\delta \in \mathbb{R}$  with  $\delta \ne 1$ , there exists a unique  $K \in C^{\infty}(\mathbb{R}^2)^{n \times n}$  that satisfies (11) for every  $(x, \xi) \in \mathbb{R}^2$  and the condition

$$\operatorname{diag} K(x, \delta x) = 0, \quad \forall x \in \mathbb{R}.$$
(13)

214 Moreover, it satisfies the estimate

$$\forall \text{ bounded } V \subset \mathbb{R}^2, \exists C > 0, \quad \|K\|_{C^s(\overline{V})^{n \times n}} \le C^s, \quad \forall s \in \mathbb{N}.$$

$$(14)$$

The kernel equations (11) have been extensively studied in the literature (see e.g. [Cor+13; 216 DVK13; HD15; Hu+16; Hu+19; CN19) but Theorem 2.5 does not follow from the results 217 contained therein. The most important difference is that, in Theorem 2.5, the kernel exists on 218 a larger set than the triangle  $\mathcal{T}$ . This is crucial since we want  $x \mapsto f(x) = -K(x,0)\Lambda \begin{pmatrix} 1 & q \end{pmatrix}^{\dagger}$ 219 to be analytic in an interval of the form  $(-\varepsilon, 1), \varepsilon > 0$ , which does not lie entirely in  $\mathcal{T}$ . This 220 yields nontrivial issues in the standard fixed point approach, notably because we now have to 221 consider points that are "on the other side" of the diagonal (x, x), that is the condition imposed 222 in the kernel equations at (x, x) cannot be consider as a boundary condition anymore. We have 223 developed in Appendix B a new approach to solve the kernel equations that encompasses in 224 particular the proof of Theorem 2.5. 225

**Remark 2.6.** Estimate (14) and Taylor's theorem show that the solution K to (11)-(13) is in fact a power series.

As a consequence of Theorem 2.5, we see that, if  $q_{i-1} = 0$ , then  $f = -K(\cdot, 0)\Lambda \begin{pmatrix} 1 & q \end{pmatrix}^{\dagger}$ satisfies  $f_i = 0$  in (0, 1) if, and only if,

$$\left\langle b, \frac{\partial^r k^{\mathfrak{c}}}{\partial x^r}(0,0) \right\rangle = 0, \quad \forall r \in \mathbb{N},$$
(15)

where  $b \in \mathbb{R}^{n-1}$  is defined in (5) and  $k^{\mathfrak{c}} = (k_{ij})_{j \neq i}$ . It remains to relate the derivatives of the kernel at the origin with M. This is the purpose of the next section. This will be done only for a very well chosen particular solution to the kernel equations (i.e. for one  $\delta \neq 1$ ) but this will be enough for our purposes as already underlined in Remark 2.2.

**Remark 2.7.** We emphasize that, in all the sections below and unless specifically mentioned, the number of negative speeds m is arbitrary (the orthogonality conditions (15) are studied for any nonzero  $b \in \mathbb{R}^{n-1}$ ).

#### <sup>238</sup> 3 The derivatives of the kernel at the origin

#### 239 3.1 Normalization of the equations

Let us first observe that a feature of the kernel equations (11)-(13) is that it does not couple different rows of K:

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$$\begin{cases} \lambda_i \frac{\partial k_{ij}}{\partial x}(x,\xi) + \frac{\partial k_{ij}}{\partial \xi}(x,\xi)\lambda_j + \sum_{r=1}^n k_{ir}(x,\xi)m_{rj} = 0,\\ \lambda_i k_{ij}(x,x) - k_{ij}(x,x)\lambda_j = m_{ij} \quad (j \neq i), \quad k_{ii}(x,\delta x) = 0. \end{cases}$$
(16)

Therefore, all along Section 3,  $i \in \{1, ..., n\}$  is fixed and we will drop the dependence on i for clarity.

Let us now introduce some important notations. Some of them have already been introduced in Section 1.3 but they are recalled here for the sake of the presentation.

• It is convenient to normalize the kernel equations by  $\lambda_i - \lambda_j$  for  $j \neq i$  and by  $\lambda_i$  otherwise. The kernel equations (16) become

$$\begin{cases} \mu_j \frac{\partial k_j}{\partial x}(x,\xi) + \frac{\partial k_j}{\partial \xi}(x,\xi)\nu_j + \sum_{r=1}^n k_r(x,\xi)\alpha_{rj} = 0, \\ k_j(x,x) = \alpha_{ij} \quad (j \neq i), \quad k_i(x,\delta x) = 0, \end{cases}$$
(17)

where  $k = \begin{pmatrix} k_{i1} & \cdots & k_{in} \end{pmatrix}^{\mathsf{T}}$  and

$$\mu_j = \frac{\lambda_i}{\lambda_i - \lambda_j}, \quad \nu_j = \frac{\lambda_j}{\lambda_i - \lambda_j}, \quad (j \neq i), \quad \mu_i = \nu_i = 1,$$
$$\alpha_{rj} = \frac{m_{rj}}{\lambda_i - \lambda_j}, \quad (j \neq i), \quad \alpha_{ri} = \frac{m_{ri}}{\lambda_i}.$$

Note that, with this normalization, we have in particular  $\mu_j - \nu_j = 1$  for  $j \neq i$ .

• Since the component  $k_i$  plays a different role than all the other components  $k_j$  with  $j \neq i$ , we rewrite (17) in a matrix form separating both quantities. Let us denote by  $n^{\mathfrak{c}} = n - 1$ and introduce  $k^{\mathfrak{c}} = (k_j)_{j\neq i}$ . Then, system (17) can be written as

$$\begin{cases} D_{\mu} \frac{\partial k^{\mathfrak{c}}}{\partial x}(x,\xi) + D_{\nu} \frac{\partial k^{\mathfrak{c}}}{\partial \xi}(x,\xi) + A^{\mathsf{T}} k^{\mathfrak{c}}(x,\xi) + k_{i}(x,\xi)\psi^{0} = 0, \\ \frac{\partial k_{i}}{\partial x}(x,\xi) + \frac{\partial k_{i}}{\partial \xi}(x,\xi) + \langle w, k^{\mathfrak{c}}(x,\xi) \rangle = 0, \\ k^{\mathfrak{c}}(x,x) = \psi^{0}, \quad k_{i}(x,\delta x) = 0, \end{cases}$$
(18)

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where  $D_{\mu}, D_{\nu}, A \in \mathbb{R}^{n^{\mathfrak{c}} \times n^{\mathfrak{c}}}$  are the matrices defined by

$$D_{\mu} = \operatorname{diag}(\mu_j)_{j \neq i}, \quad D_{\nu} = \operatorname{diag}(\nu_j)_{j \neq i}, \quad A = (\alpha_{rj})_{r, j \neq i},$$

and  $\psi^0, w \in \mathbb{R}^{n^{\mathfrak{c}}}$  are the vectors defined by

$$\psi^0 = (\alpha_{ij})_{j \neq i}, \quad w = (\alpha_{ji})_{j \neq i}.$$

Note that we used that  $\alpha_{ii} = 0$  (since  $M \in \mathbb{R}^{n \times n}_0$ ). Finally, it will also be convenient to use the matrix  $D \in \mathbb{R}^{n^c \times n^c}$  and the vector  $v \in \mathbb{R}^{n^c}$  defined by

$$D = D_{\mu} + D_{\nu}, \quad v = -\frac{1}{2}w.$$
 (19)

#### 258 3.2 Computation of the derivatives

<sup>259</sup> The main result of this section is the following.

**Theorem 3.1.** For the solution to (18) with  $\delta = -1$ , we have

$$\frac{\partial^{\gamma+\sigma}k^{\mathfrak{c}}}{\partial x^{\gamma}\partial\xi^{\sigma}}(0,0) = \sum_{r=0}^{\gamma}\sum_{s=0}^{\sigma}(-1)^{r}\binom{\gamma}{r}\binom{\sigma}{s}\psi_{\gamma+\sigma-(r+s),r+s}, \quad \forall \gamma, \sigma \in \mathbb{N},$$
(20)

262 where  $(\psi_{r,s})_{r,s\in\mathbb{N}}$  is the sequence defined by

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$$\psi_{r,0} = \psi^r, \quad \psi_{r,s} = 0 \quad \text{if } r < s, \quad \psi_{r,s} = \sum_{\ell=0}^{r-s} \langle v, \psi_{r-1-\ell,s-1} \rangle \psi_{\ell,0} \quad \text{if } r \ge s \ge 1,$$
 (21)

where  $(\psi^r)_{r\in\mathbb{N}}$  is the sequence defined in (4).

Combining this result with the estimates (14) and Taylor's theorem, we obtain an explicit formula for the solution to (18) when  $\delta = -1$ : **Corollary 3.2.** For  $\delta = -1$ , the solution to (18) is given by

$$k^{\mathfrak{c}}(x,\xi) = \sum_{r=0}^{+\infty} \sum_{s=0}^{+\infty} \frac{(-1)^r}{r!s!} \psi_{r,s}(x-\xi)^r (x+\xi)^s, \quad k_i(x,\xi) = -\int_{\frac{x-\xi}{2}}^{x} \langle w, k^{\mathfrak{c}}(\sigma,\sigma-x+\xi) \rangle \, d\sigma,$$

for every  $(x,\xi) \in \mathbb{R}^2$ , where  $(\psi_{r,s})_{r,s\in\mathbb{N}}$  is the sequence defined by (21) and the series is normally convergent on any compact set of  $\mathbb{R}^2$ .

**Remark 3.3.** Explicit solutions to the kernel equations were also obtained in [VK14, Section 3.4] for systems of n = 2 equations.

Proof of Theorem 3.1. 1. To explain the special role played by  $\delta = -1$ , we start the computations with an arbitrary  $\delta \neq 1$ . The first idea is to form a system involving only  $k^{\mathfrak{c}}$  by expressing  $k_i$  as a function of  $k^{\mathfrak{c}}$ :

$$k_i(x,\xi) = -\int_{\frac{x-\xi}{1-\delta}}^x \langle w, k^{\mathfrak{c}}(\sigma, \sigma - x + \xi) \rangle \, d\sigma.$$

As a result,  $k^{\mathfrak{c}}$  solves

$$\begin{cases} D_{\mu} \frac{\partial k^{\mathfrak{c}}}{\partial x}(x,\xi) + D_{\nu} \frac{\partial k^{\mathfrak{c}}}{\partial \xi}(x,\xi) + A^{\mathsf{T}} k^{\mathfrak{c}}(x,\xi) - \left( \int_{\frac{x-\xi}{1-\delta}}^{x} \left\langle w, k^{\mathfrak{c}}(\sigma,\sigma-x+\xi) \right\rangle d\sigma \right) \psi^{0} = 0, \\ k^{\mathfrak{c}}(x,x) = \psi^{0}. \end{cases}$$

We now transform this system into a Cauchy problem by introducing the transformation

$$h(t,\theta) = k^{\mathfrak{c}}\left(\frac{-t+\theta}{2}, \frac{t+\theta}{2}\right).$$

Using that  $D_{\mu} - D_{\nu} = \mathrm{Id}_{n^{\mathfrak{c}}}$ , we can check that h satisfies the system

$$\begin{cases} \frac{\partial h}{\partial t}(t,\theta) = D\frac{\partial h}{\partial \theta}(t,\theta) + A^{\mathsf{T}}h(t,\theta) + \left(\int_{-\frac{1+\delta}{1-\delta}t}^{\theta} \langle v, h(t,\eta) \rangle \, d\eta\right) \psi^{0}, \\ h(0,\theta) = \psi^{0}, \end{cases}$$
(22)

where we recall that D and v are defined in (19). Note as well that

$$\frac{\partial^{\gamma+\sigma}k^{\mathfrak{c}}}{\partial x^{\gamma}\partial\xi^{\sigma}}(x,\xi) = \sum_{r=0}^{\gamma}\sum_{s=0}^{\sigma}(-1)^{r}\binom{\gamma}{r}\binom{\sigma}{s}\frac{\partial^{\gamma+\sigma}h}{\partial t^{\gamma+\sigma-(r+s)}\partial\theta^{r+s}}(-x+\xi,x+\xi), \quad \forall \gamma,\sigma\in\mathbb{N},$$

so that the derivatives of  $k^{\mathfrak{c}}$  at (0,0) can be deduced from those of h. They will be computed from (22) and we see that the computations considerably simplify if the lower bound of the integral vanishes, that is if we choose  $\delta = -1$ . For this choice, we define

$$\psi_{r,s} = \frac{\partial^{r+s}h}{\partial t^r \partial \theta^s}(0,0).$$

We are going to show that it satisfies (21).

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274 2. All along the rest of the proof, we will use the notation  $c_{r,s} = \langle v, \psi_{r,s} \rangle$ . First observe that 275 system (22) (with  $\delta = -1$ ) yields the following identities:

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$$\begin{cases} \psi_{r+1,0} = D\psi_{r,1} + A^{\mathsf{T}}\psi_{r,0}, \\ \psi_{r+1,s} = D\psi_{r,s+1} + A^{\mathsf{T}}\psi_{r,s} + c_{r,s-1}\psi_{0,0}, \\ \psi_{0,s} = 0, \end{cases}$$
(23)

for every  $r \ge 0$  and  $s \ge 1$ . The second property in (21) is easily proved by induction on  $r \ge 1$ . To establish the two other identities, it is sufficient to prove the following statement for any  $N \ge 1$ :

$$\begin{cases} \psi_{s+q,s} = \sum_{\ell=0}^{q} c_{s-1+q-\ell,s-1} \psi_{\ell,0}, & \forall s \ge 1, \,\forall 0 \le q \le N, \\ \psi_{r,0} = A^{\mathsf{T}} \psi_{r-1,0} + D \sum_{\ell=0}^{r-2} c_{r-2-\ell,0} \psi_{\ell,0}, & \forall 2 \le r \le N+1. \end{cases}$$
(24)

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We prove it by induction. For N = 1, this is clear. Indeed, for any  $s \ge 1$ , we have

$$\psi_{s,s} = D\psi_{s-1,s+1} + A^{\mathsf{T}}\psi_{s-1,s} + c_{s-1,s-1}\psi_{0,0} \quad (\text{by (23)}),$$
  
=  $c_{s-1,s-1}\psi_{0,0}$  (by the second property in (21)), (25)

$$\begin{split} \psi_{s+1,s} &= D\psi_{s,s+1} + A^{\mathsf{T}}\psi_{s,s} + c_{s,s-1}\psi_{0,0} \quad \text{(by (23))}, \\ &= A^{\mathsf{T}}\psi_{s,s} + c_{s,s-1}\psi_{0,0} \quad \text{(by the second property in (21))}, \\ &= c_{s-1,s-1}\psi_{1,0} + c_{s,s-1}\psi_{0,0} \quad \text{(by (25))}, \end{split}$$

 $\operatorname{and}$ 

$$\psi_{2,0} = D\psi_{1,1} + A^{\mathsf{T}}\psi_{1,0} \quad \text{(by (23))},$$
$$= c_{0,0}D\psi_{0,0} + A^{\mathsf{T}}\psi_{1,0} \quad \text{(by (25))}.$$

Assume now that (24) holds for  $N \ge 1$  and let us prove it for N + 1. We first show that

$$\psi_{s+N+1,s} = \sum_{\ell=0}^{N+1} c_{s+N-\ell,s-1} \psi_{\ell,0}, \quad \forall s \ge 1.$$
(26)

For any  $s \ge 1$ , we have

$$\begin{split} \psi_{s+N+1,s} &= D\psi_{s+N,s+1} + A^{\mathsf{T}}\psi_{s+N,s} + c_{s+N,s-1}\psi_{0,0} \quad \text{(by (23))}, \\ &= D\sum_{\ell=0}^{N-1} c_{s+N-1-\ell,s}\psi_{\ell,0} + A^{\mathsf{T}}\sum_{\ell=0}^{N} c_{s-1+N-\ell,s-1}\psi_{\ell,0} \\ &+ c_{s+N,s-1}\psi_{0,0} \quad \text{(by assumption (24))}, \\ &= D\sum_{\ell=0}^{N-1} c_{s+N-1-\ell,s}\psi_{\ell,0} + \sum_{r=2}^{N+1} c_{s+N-r,s-1}A^{\mathsf{T}}\psi_{r-1,0} \\ &+ c_{s-1+N,s-1}\psi_{1,0} + c_{s+N,s-1}\psi_{0,0}. \end{split}$$

283 Therefore, if we show the identity

$$\sum_{\ell=0}^{N-1} c_{s+N-1-\ell,s} \psi_{\ell,0} = \sum_{r=2}^{N+1} c_{s+N-r,s-1} \left( \sum_{\ell=0}^{r-2} c_{r-2-\ell,0} \psi_{\ell,0} \right),$$
(27)

then we can use the second condition in (24) to obtain the desired identity (26). To establish (27), we use the first condition in (24) to deduce that, for every  $0 \le \ell \le N - 1$ ,

$$\langle v, \psi_{s+N-1-\ell,s} \rangle = \sum_{j=0}^{N-1-\ell} c_{s-1+N-1-\ell-j,s-1} c_{j,0}$$
$$= \sum_{r=\ell+2}^{N+1} c_{s+N-r,s-1} c_{r-2-\ell,0}.$$

Finally, a simple change of order of summation leads to (27). It remains to show the second identity in (24) for r = N + 2, namely

$$\psi_{N+2,0} = A^{\mathsf{T}} \psi_{N+1,0} + D \sum_{\ell=0}^{N} c_{N-\ell,0} \psi_{\ell,0}$$

We have

$$\psi_{N+2,0} = D\psi_{N+1,1} + A^{\mathsf{T}}\psi_{N+1,0} \quad \text{(by (23))},$$
$$= D\sum_{\ell=0}^{N} c_{N-\ell,0}\psi_{\ell,0} + A^{\mathsf{T}}\psi_{N+1,0} \quad \text{(by (24))}.$$

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**Remark 3.4.** Theoretically, we can also compute all the derivatives at (0,0) of the solution to 286 (18) for arbitrary  $\delta \neq 1$ . This can be done by taking derivatives and inverting some matrix. 287 However, the size of this matrix grows with the order of derivatives and computations rapidly 288 become more and more complicated. It seems difficult with such a strategy to obtain a suitable 289 formula for arbitrary  $\delta$ . At the same time, we see from (22) that a different choice of  $\delta$  means 290 more derivatives to be computed, as for instance with  $\delta = 0$  which leads to an integral of the 291 form  $\int_{-t}^{\theta}$ . In addition to that, we recall that one choice of  $\delta$  is actually sufficient for the purposes 292 of this paper (Remark 2.2). 293

### <sup>294</sup> 4 Study of the orthogonality conditions

In this section, we use the computations obtained in the previous section to study the orthogonality conditions (15). We start with the conclusion of the proof of our first main result.

Proof of Theorem 1.3. We recall that, from the results of the previous sections, we only have to show the equivalence between the orthogonality conditions (15) and (6). First observe that, from the definition (21) of the sequence  $(\psi_{r,s})_{r,s\in\mathbb{N}}$ , it is clear that (6) is equivalent to

$$\langle b, \psi_{r,s} \rangle = 0, \quad \forall r, s \in \mathbb{N}.$$

<sup>297</sup> We can check that this condition is equivalent to (15) using (20) and (21).

**Remark 4.1.** The proof above and the analyticity of  $k^{\mathfrak{c}}$  in  $\mathbb{R}^2$  shows that the following four properties are in fact equivalent:

300 1. 
$$\langle b, k^{\mathfrak{c}}(x,\xi) \rangle = 0$$
 for every  $(x,\xi) \in \mathbb{R}^2$ 

301 2. 
$$\langle b, k^{\mathfrak{c}}(x, 0) \rangle = 0$$
 for every  $x \in \mathbb{R}$ .

302 3. 
$$\langle b, \psi_{r,0} \rangle = 0$$
 for every  $r \in \mathbb{N}$ .

303 4.  $\langle b, \psi_{r,s} \rangle = 0$  for every  $r, s \in \mathbb{N}$ .

We are now going to study the orthogonality conditions (6) for the sequence  $(\psi^r)_{r\in\mathbb{N}}$  and prove our second main result. For the rest of Section 4, the matrices  $A, D \in \mathbb{R}^{n^c \times n^c}$  and the vectors  $\psi^0, v, b \in \mathbb{R}^{n^c}$  can in fact be arbitrary. We emphasize that n is also arbitrary, it is only during the proof of Theorem 1.6 that we will assume that n = 3 or n = 4.

#### <sup>308</sup> 4.1 Some invariant subspaces of the sequence

We start the general study of the orthogonality conditions (6) with the description of two simple invariant subspaces of the sequence  $(\psi^r)_{r\in\mathbb{N}}$ .

**Proposition 4.2.** Assume that  $\psi^0 \in E$  for some  $E \subset \mathbb{R}^{n^c}$  satisfying one of the following two conditions:

$$A^{\mathsf{T}}(E) \subset E, \quad E \subset \ker v^{\mathsf{T}}.$$
 (28)

$$A^{\mathsf{T}}(E) \subset E, \quad D(E) \subset E.$$
<sup>(29)</sup>

311 Then,  $\psi^r \in E$  for every  $r \in \mathbb{N}$ .

*Proof.* We prove the result by induction on r. For r = 0 this is trivial and for r = 1 this follows from the definition  $\psi^1 = A^{\mathsf{T}}\psi^0$  and the property  $A^{\mathsf{T}}(E) \subset E$ . Assume then that  $\psi^{\ell} \in E$  for every  $0 \leq \ell \leq r$  for some  $r \geq 1$  and let us show that  $\psi^{r+1} \in E$ . Since  $r+1 \geq 2$ , we have

$$\psi^{r+1} = A^{\mathsf{T}}\psi^r + \sum_{\ell=0}^{r-1} \langle v, \psi^{r-1-\ell} \rangle D\psi^\ell.$$

Clearly, the first part  $A^{\mathsf{T}}\psi^r$  belongs to E since  $\psi^r \in E$  and  $A^{\mathsf{T}}(E) \subset E$ . The remaining part also belongs to E since either  $\langle v, \psi^\ell \rangle = 0$  for every  $0 \leq \ell \leq r$  (if  $E \subset \ker v^{\mathsf{T}}$ ) or  $D\psi^\ell \in E$  for every  $0 \leq \ell \leq r$  (if  $D(E) \subset E$ ).

If we can find a subspace E such as in the previous proposition and which is in addition included in ker  $b^{\mathsf{T}}$ , then we see that the whole sequence will be guaranteed to stay in ker  $b^{\mathsf{T}}$ .

#### 317 4.2 Characterization of rank one sequences

In this section, we characterize when the rank of  $(\psi^r)_{r\in\mathbb{N}}$  is equal to one, and we use it to prove Theorem 1.6 in the case n = 3. We recall that, by definition, rank  $(\psi^r)_{r\in S} = \dim \operatorname{span} \{\psi^r \mid r \in S\}$ for any  $S \subset \mathbb{N}$ .

From now on, it will be convenient to use the following notation:

$$E_s = \operatorname{span} \{ \psi^r \mid r \leq s \}, \quad \forall s \in \mathbb{N}.$$

First of all, it is clear that

$$\operatorname{rank} (\psi^r)_{r \in \mathbb{N}} = 1 \quad \Longleftrightarrow \quad \left(\psi^0 \neq 0, \quad \psi^r \in E_0, \quad \forall r \ge 1\right).$$

321 We have the following characterization:

- **Proposition 4.3.** The following three conditions are equivalent:
- 323 1.  $\psi^r \in E_0$  for every  $r \ge 1$ .
- 324 2.  $\psi^r \in E_0$  for every  $r \in \{1, 2\}$ .
- 325 3.  $E_0$  satisfies (28) or (29).

*Proof.* The implication  $1 \Longrightarrow 2$  is trivial. The implication  $3 \Longrightarrow 1$  follows from Proposition 4.2. Let us show that  $2 \Longrightarrow 3$ . We write

$$\psi^r = \alpha_r \psi^0, \quad r = 1, 2,$$

for some  $\alpha_r \in \mathbb{R}$ . The condition for r = 1 gives  $A^{\mathsf{T}}\psi^0 = \alpha_1\psi^0$ , which is equivalent to  $A^{\mathsf{T}}(E_0) \subset E_0$ . The condition for r = 2 gives  $A^{\mathsf{T}}\psi^1 + \langle v, \psi^0 \rangle D\psi^0 = \alpha_2\psi^0$ , which implies  $\langle v, \psi^0 \rangle D\psi^0 \in E_0$ , that is either  $E_0 \subset \ker v^{\mathsf{T}}$  or  $D(E_0) \subset E_0$ . This establishes the desired equivalences.

Proof of Theorem 1.6 (case n = 3). Assume that

$$\langle b, \psi^r \rangle = 0, \quad \forall r \in \{0, 1, 2\}$$

Since  $n^{\mathfrak{c}} = 2$ , we have dim ker  $b^{\mathsf{T}} = 1$  and thus

$$\operatorname{rank}(\psi^r)_{r \in \{0,1,2\}} \le 1.$$

As a result,  $\psi^r \in E_0$  for  $r \in \{1, 2\}$  and it follows that  $\psi^r \in E_0$  for every  $r \ge 1$  by Proposition 331 4.3.

#### <sup>332</sup> 4.3 Characterization of rank two sequences

In this section, we characterize when the rank of  $(\psi^r)_{r\in\mathbb{N}}$  is equal to two, and we use it to prove Theorem 1.6 in the case n = 4. The study of rank two sequences is considerably more difficult than for rank one.

We start with the following simple observation.

**Proposition 4.4.** Let  $\{0, 1, 2\} \subset S \subset \mathbb{N}$ . We have rank  $(\psi^r)_{r \in S} = 2$  if, and only if, we have one of the following two conditions:

339 1. rank 
$$(\psi^0 | \psi^1) = 1$$
, rank  $(\psi^0 | \psi^2) = 2$  and  $\psi^r \in E_2$  for every  $r \in S \setminus \{0, 1, 2\}$ .

340 2. rank 
$$(\psi^0|\psi^1) = 2$$
 and  $\psi^r \in E_1$  for every  $r \in S \setminus \{0,1\}$ .

Proof. We only need to observe that the situation rank  $(\psi^0|\psi^1) = \operatorname{rank}(\psi^0|\psi^2) = 1$  does not happen. Indeed, in that situation we have  $\psi^1, \psi^2 \in E_0$  and thus  $\psi^r \in E_0$  for every  $r \ge 1$  by Proposition 4.3, which shows that the sequence cannot be of rank two.

We now characterize both conditions of Proposition 4.4.

**Proposition 4.5.** Assume that rank  $(\psi^0|\psi^1) = 1$  and rank  $(\psi^0|\psi^2) = 2$ . Then, the following three conditions are equivalent:

347 1. 
$$\psi^r \in E_2$$
 for every  $r \geq 3$ 

348 2.  $\psi^r \in E_2$  for every  $r \in \{3, 4\}$ .

349 3.  $E_2$  satisfies (29).

Proof. The implication  $1 \implies 2$  is trivial. The implication  $3 \implies 1$  follows from Proposition 4.2. Let us show that  $2 \implies 3$ . From the definition of  $\psi^2$  and the rank assumptions, we have  $\langle v, \psi^0 \rangle \neq 0$  and  $D\psi^0 \in \text{span } \{\psi^0, \psi^2\}$ . Using these facts, we can first check that the condition  $\psi^3 \in E_2$  gives  $A^{\mathsf{T}}(E_2) \subset E_2$  and then that the condition  $\psi^4 \in E_2$  yields  $D(E_2) \subset E_2$ .

The second condition in Proposition 4.4 is more difficult to characterize.

**Proposition 4.6.** Assume that rank  $(\psi^0|\psi^1) = 2$ . Then, the following two conditions are equivalent:

357 1.  $\psi^r \in E_1$  for every  $r \geq 2$ .

358 2.  $\psi^r \in E_1$  for every  $r \in \{2, \dots, 5\}$ .

Before proving Proposition 4.6, we prove our second main result.

Proof of Theorem 1.6 (case n = 4). Assume that

$$\langle b, \psi^r \rangle = 0, \quad \forall r \in \{0, \dots, 5\}.$$

Since  $n^{\mathfrak{c}} = 3$ , we have dim ker  $b^{\mathsf{T}} = 2$  and thus

$$\operatorname{rank}(\psi^r)_{r \in \{0,...,5\}} \le 2.$$

If the rank is in fact less than or equal to 1, then we conclude as in the proof of Theorem 1.6 in the case n = 3. If the rank is exactly equal to 2, then the conclusion follows from Propositions 4.4, 4.5 and 4.6.

We now turn to the proof of the key proposition.

Proof of Proposition 4.6. If  $E_1$  satisfies (28) or (29), then the result follows from Proposition 4.2. Therefore, from now on, we assume that  $E_1$  does not meet any of these conditions.

1. Let  $N \ge 4$  be arbitrary and consider the property: for every  $r \in \{2, ..., N\}$ , there exist  $\alpha_r, \beta_r \in \mathbb{R}$  such that

$$\psi^r = \alpha_r \psi^0 + \beta_r \psi^1. \tag{30}$$

Let us also set

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$$\alpha_0 = 1, \quad \beta_0 = 0, 
\alpha_1 = 0, \quad \beta_1 = 1,$$
(31)

so that the previous identity is always true for r = 0, 1. Using the definition (4) of the sequence we see that, for  $r \ge 2$ , identity (30) is equivalent to

$$\alpha_{r}\psi^{0} + \beta_{r}\psi^{1} = \alpha_{r-1}A^{\mathsf{T}}\psi^{0} + \beta_{r-1}A^{\mathsf{T}}\psi^{1} + \left(\sum_{\ell=0}^{r-2} c_{r-2-\ell}\alpha_{\ell}\right)D\psi^{0} + \left(\sum_{\ell=0}^{r-2} c_{r-2-\ell}\beta_{\ell}\right)D\psi^{1}$$

where we introduced, for every  $s \in \{0, \ldots, N\}$ ,

$$c_s = \alpha_s \langle v, \psi^0 \rangle + \beta_s \langle v, \psi^1 \rangle.$$

Let us eliminate the terms  $A^{\mathsf{T}}\psi^0$ ,  $A^{\mathsf{T}}\psi^1$ . By definition of the sequence, we have  $A^{\mathsf{T}}\psi^0 = \psi^1$ . On the other hand, condition (30) for r = 2 yields

$$A^{\mathsf{T}}\psi^{1} = \alpha_{2}\psi^{0} + \beta_{2}\psi^{1} - c_{0}D\psi^{0}.$$
(32)

As a result, for  $r \ge 3$ , identity (30) is equivalent to

$$(\psi^0|\psi^1)u_r = (D\psi^0|D\psi^1)f_{r-1}, \tag{33}$$

where  $u_r, f_{r-1} \in \mathbb{R}^2$  are the vectors defined by

$$u_{r} = \begin{pmatrix} \alpha_{r} - \beta_{r-1}\alpha_{2} \\ \beta_{r} - (\alpha_{r-1} + \beta_{r-1}\beta_{2}) \end{pmatrix}, \quad f_{r-1} = \begin{pmatrix} \beta_{r-1}(-c_{0}) + \sum_{\ell=0}^{r-2} c_{r-2-\ell}\alpha_{\ell} \\ \sum_{\ell=0}^{r-2} c_{r-2-\ell}\beta_{\ell} \end{pmatrix}.$$
 (34)

- 378 Let us make some observations.
  - For any  $r_1, r_2$ , if  $\lambda f_{r_1-1} + \mu f_{r_2-1} = 0$  for some  $\lambda, \mu \in \mathbb{R}$ , then  $\lambda u_{r_1} + \mu u_{r_2} = 0$ . Indeed, denoting by  $u = \lambda u_{r_1} + \mu u_{r_2}$ , we have  $(\psi^0 | \psi^1) u = 0$  and thus u = 0 since we assumed that  $\psi^0, \psi^1$  are linearly independent.
  - $f_{r_1-1}$  and  $f_{r_2-1}$  are linearly dependent for any  $r_1, r_2$ . Indeed, otherwise we obtain  $(\psi^0|\psi^1)U = (D\psi^0|D\psi^1)$  for some matrix  $U \in \mathbb{R}^{2\times 2}$ . This means that  $D(E_1) \subset E_1$ . It then follows from (32) that  $A^{\mathsf{T}}(E_1) \subset E_1$  as well. Therefore,  $E_1$  satisfies (29), but this situation has been excluded from the beginning of the proof.
- $f_2$  and  $f_3$  are linearly dependent only if  $c_0 \neq 0$ . Indeed, if  $c_0 = 0$ , then  $A^{\mathsf{T}}\psi^1 \in E_1$ by (32), so that  $A^{\mathsf{T}}(E_1) \subset E_1$ . If  $c_0 = 0$ , we also have  $\det(f_2|f_3) = c_1^2$ , which cannot be zero since otherwise  $c_0 = c_1 = 0$ , that is  $E_1 \subset \ker v^{\mathsf{T}}$ , and thus  $E_1$  satisfies (28) (excluded).

The necessary condition  $det(f_2|f_{r-1}) = 0$  is equivalent to the identity

$$\left(\sum_{\ell=0}^{r-2} c_{r-2-\ell}\beta_{\ell}\right) f_2 - c_0 f_{r-1} = 0,$$

which, by the first observation above, in turn implies that

$$\left(\sum_{\ell=0}^{r-2} c_{r-2-\ell} \beta_{\ell}\right) u_3 - c_0 u_r = 0$$

Since  $c_0 \neq 0$ , this gives the following formulas:

$$\begin{cases} f_{r-1} = \left(\sum_{\ell=0}^{r-2} \bar{c}_{r-2-\ell} \beta_\ell\right) f_2, \\ u_r = \left(\sum_{\ell=0}^{r-2} \bar{c}_{r-2-\ell} \beta_\ell\right) u_3, \end{cases}$$
(35)

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where we introduced  $\bar{c}_s = \frac{c_s}{c_0}$ . Conversely, it is clear that if we have (35), then (33) will also hold for  $r \ge 3$ , provided that it holds for r = 3. Finally, observe that the second formula in (35), combined with the definition (34) of  $u_r$ , uniquely determines all the  $\alpha_r$ ,  $\beta_r$  for  $r \ge 4$  as a function of  $\alpha_2$ ,  $\beta_2$  and  $\alpha_3$ ,  $\beta_3$ :

$$\begin{cases} \alpha_{r} = \beta_{r-1}\alpha_{2} + \left(\sum_{\ell=0}^{r-2} \bar{c}_{r-2-\ell}\beta_{\ell}\right) (\alpha_{3} - \beta_{2}\alpha_{2}), \\ \beta_{r} = \alpha_{r-1} + \beta_{r-1}\beta_{2} + \left(\sum_{\ell=0}^{r-2} \bar{c}_{r-2-\ell}\beta_{\ell}\right) (\beta_{3} - (\alpha_{2} + \beta_{2}^{2})), \end{cases}$$
(36)

and that the first formula in (35) is equivalent to

$$\beta_{r-1} = \sum_{\ell=0}^{r-2} \bar{c}_{r-2-\ell} \left( \alpha_{\ell} + (\beta_2 - \bar{c}_1) \beta_{\ell} \right).$$
(37)

In summary, we have shown that the property considered is equivalent to:  $c_0 \neq 0$  and there exist  $\alpha_2, \beta_2$  and  $\alpha_3, \beta_3$  such that (30) holds for r = 2, 3 and such that the sequence defined by (36) (with (31)) satisfies (37) for every  $r \in \{4, \ldots, N\}$ .

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  2. Let us now study the sequence (36). The proof of the result will be complete after we show
  403 that the following three conditions are equivalent:
- (a) Condition (37) holds for every  $r \ge 4$ .
- (b) Condition (37) holds for r = 4, 5.
- 406 (c)  $\alpha_3$  and  $\beta_3$  are given by

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$$\begin{cases} \beta_3 = (\alpha_2 + \beta_2^2) - (\beta_2 - \bar{c}_1)^2 + \alpha_2 + (\beta_2 - \bar{c}_1)\beta_2, \\ \alpha_3 = \beta_2\alpha_2 - (\beta_2 - \bar{c}_1)(\beta_3 - (\alpha_2 + \beta_2^2)). \end{cases}$$
(38)

We start with the implication 2b  $\implies$  2c. Condition (37) for r = 4 immediately gives  $\beta_3$  as a function of  $\alpha_2, \beta_2$ :

$$\beta_3 = \sum_{\ell=0}^2 \bar{c}_{2-\ell} \left( \alpha_\ell + (\beta_2 - \bar{c}_1) \beta_\ell \right).$$

On the other hand, condition (37) for r = 5 gives

$$\beta_4 = \sum_{\ell=1}^2 \bar{c}_{3-\ell} \left( \alpha_\ell + (\beta_2 - \bar{c}_1)\beta_\ell \right) + \bar{c}_0 \left( \alpha_3 + (\beta_2 - \bar{c}_1)\beta_3 \right) + \bar{c}_3 \left( \alpha_0 + (\beta_2 - \bar{c}_1)\beta_0 \right)$$
$$= 2\alpha_3 + \beta_2\beta_3 + \sum_{\ell=1}^2 \bar{c}_{3-\ell} \left( \alpha_\ell + (\beta_2 - \bar{c}_1)\beta_\ell \right),$$

whereas, by definition (36),

$$\beta_4 = \alpha_3 + \beta_3 \beta_2 + \left(\sum_{\ell=0}^2 \bar{c}_{2-\ell} \beta_\ell\right) \left(\beta_3 - (\alpha_2 + \beta_2^2)\right).$$

Identifying both expressions gives  $\alpha_3$  as a function of  $\alpha_2, \beta_2$ :

$$\alpha_3 = -\sum_{\ell=1}^2 \bar{c}_{3-\ell} \left( \alpha_\ell + (\beta_2 - \bar{c}_1)\beta_\ell \right) + \left( \sum_{\ell=0}^2 \bar{c}_{2-\ell}\beta_\ell \right) \left( \beta_3 - (\alpha_2 + \beta_2^2) \right).$$

We can check that the previous formulas are equivalent to (38) (we prefer the expressions in (38) because they make appear some coefficients involved in (36)).

Let us now prove the implication  $2c \implies 2a$ . We prove it by induction on r. For r = 4, this holds by very definition of  $\beta_3$  as we have seen above. Assume now that (37) holds for some arbitrary  $r \ge 4$ , and let us prove it for r + 1, that is to prove that we have

$$\beta_r = \sum_{\ell=0}^{r-1} \bar{c}_{r-1-\ell} \left( \alpha_\ell + (\beta_2 - \bar{c}_1) \beta_\ell \right).$$
(39)

By definition (36) of  $\beta_r$ , we have

$$\beta_r = \alpha_{r-1} + \beta_{r-1}\bar{c}_1 + (\beta_2 - \bar{c}_1) \left( \beta_{r-1} - (\beta_2 - \bar{c}_1) \sum_{\ell=0}^{r-2} \bar{c}_{r-2-\ell}\beta_\ell \right) + (\beta_3 - (\alpha_2 + \beta_2^2) + (\beta_2 - \bar{c}_1)^2) \sum_{\ell=0}^{r-2} \bar{c}_{r-2-\ell}\beta_\ell.$$

Using the induction hypothesis (37) and the definition (38) of  $\beta_3$ , we obtain

$$\beta_r = \bar{c}_{r-1} + \sum_{\ell=0}^{r-2} \bar{c}_{r-2-\ell} \left( (\beta_2 - \bar{c}_1)\alpha_\ell + (\alpha_2 + (\beta_2 - \bar{c}_1)\beta_2)\beta_\ell \right).$$

Now observe that, using the definition of  $\alpha_3$ , we have

$$\alpha_{\ell+1} + (\beta_2 - \bar{c}_1)\beta_{\ell+1} = \beta_\ell \alpha_2 + (\beta_2 - \bar{c}_1)(\alpha_\ell + \beta_\ell \beta_2) = (\beta_2 - \bar{c}_1)\alpha_\ell + (\alpha_2 + (\beta_2 - \bar{c}_1)\beta_2)\beta_\ell.$$

Formula (39) easily follows this identity and the previous one.

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#### 5 Kernel associated with the orthogonality conditions

<sup>417</sup> In this section, we supplement our results by giving a more explicit characterization of the <sup>418</sup> conditions found in the previous section and that guaranteed the orthogonality conditions. Then, <sup>419</sup> we discuss the structure of the associated kernel.

#### <sup>420</sup> 5.1 Kernel associated with the invariant subspaces

Here we discuss properties related to the invariant subspaces of Section 4.1. We recall that, for any vector  $h \in \mathbb{R}^{n^c}$ , we denote the Kalman matrix of (A, h) by

$$\operatorname{Kal}(A,h) = (h|Ah|A^2h|\cdots|A^{n^{\mathfrak{c}}-1}h) \in \mathbb{R}^{n^{\mathfrak{c}} \times n^{\mathfrak{c}}}.$$

#### **Proposition 5.1.** 1. There exists E satisfying (28), $\psi^0 \in E$ and $E \subset \ker b^{\mathsf{T}}$ , if, and only if,

$$\operatorname{Kal}(A, v)^{\mathsf{T}} \psi^{0} = \operatorname{Kal}(A, b)^{\mathsf{T}} \psi^{0} = 0.$$
 (40)

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2. Assume that  $\operatorname{Kal}(A, v)^{\mathsf{T}} \psi^0 = 0$ . Then, for any  $\delta \neq 1$ , the solution to the kernel equations (18) is

$$k_i(x,\xi) = 0, \quad k^{\mathfrak{c}}(x,\xi) = e^{-A^{\mathsf{T}}(x-\xi)}\psi^0, \quad \forall (x,\xi) \in \mathbb{R}^2.$$

If, moreover,  $\operatorname{Kal}(A, b)^{\mathsf{T}} \psi^0 = 0$ , then the orthogonality conditions (15) are satisfied.

*Proof.* 1. Assume that (40) holds. Let us define

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$$E = \ker \operatorname{Kal}(A, v)^{\mathsf{T}} \cap \ker \operatorname{Kal}(A, b)^{\mathsf{T}}.$$

By assumption,  $\psi^0 \in E$  and it is clear that  $E \subset \ker v^{\mathsf{T}}$  and  $E \subset \ker b^{\mathsf{T}}$ . Finally, E is stable by  $A^{\mathsf{T}}$  thanks to Cayley-Hamilton theorem. Conversely, assume that (28) holds for some  $E \subset \ker b^{\mathsf{T}}$  with  $\psi^0 \in E$ . Since  $\psi^0 \in E$  and E is stable by  $A^{\mathsf{T}}$ , we have  $(A^{\mathsf{T}})^k \psi^0 \in E$  for every  $k \in \mathbb{N}$ . Since  $E \subset \ker v^{\mathsf{T}}$  and  $E \subset \ker b^{\mathsf{T}}$ , we obtain (40).

2. We see from the kernel equations (18) that  $k_i = 0$  if, and only if, we have

$$\begin{cases} D_{\mu} \frac{\partial k^{\mathfrak{c}}}{\partial x}(x,\xi) + D_{\nu} \frac{\partial k^{\mathfrak{c}}}{\partial \xi}(x,\xi) + A^{\mathsf{T}} k^{\mathfrak{c}}(x,\xi) = 0, \\ \langle v, k^{\mathfrak{c}}(x,\xi) \rangle = 0, \\ k^{\mathfrak{c}}(x,x) = \psi^{0}. \end{cases}$$

Using that  $D_{\mu} - D_{\nu} = \mathrm{Id}_{n^{c}}$ , it is clear that  $k^{c}(x,\xi) = e^{-A^{\mathsf{T}}(x-\xi)}\psi^{0}$  satisfies the first equation. The second condition follows from the assumption  $\mathrm{Kal}(A, v)^{\mathsf{T}}\psi^{0} = 0$  and Cayley-Hamilton theorem. The third condition is trivial. Finally, the orthogonality conditions are clearly satisfied under the additional assumption  $\mathrm{Kal}(A, b)^{\mathsf{T}}\psi^{0} = 0$ .

Let us now address the second type of invariant subspaces introduced in Section 4.1.

#### **Proposition 5.2.** Assume that $\psi^0, b \neq 0$ .

There exists E satisfying (29), ψ<sup>0</sup> ∈ E and E ⊂ ker b<sup>T</sup> if, and only if, there exists a nonempty J ⊊ {1,...,n<sup>c</sup>} such that

$$\psi_j^0 = a_{rj} = b_r = 0, \quad \forall j \notin J, \, \forall r \in J.$$

2. Assume that there exists a nonempty  $J \subsetneq \{1, \ldots, n^{\mathfrak{c}}\}$  such that  $\psi_j^0 = a_{rj} = 0$  for every  $j \notin J$  and  $r \in J$ . Then, for any  $\delta \neq 1$ , the solution to the kernel equations (18) satisfies

$$k_j^{\mathfrak{c}} = 0, \quad \forall j \notin J. \tag{41}$$

If, moreover, 
$$b_r = 0$$
 for every  $r \in J$ , then the orthogonality conditions (15) are satisfied.

*Proof.* 1. Since D is a diagonal matrix with distinct entries, its invariant subspaces are of the form

$$E = \operatorname{span} \{ e_r \mid r \in J \}$$

for some  $J \subset \{1, \ldots, n^{\mathfrak{c}}\}$ , where  $e_1, \ldots, e_{n^{\mathfrak{c}}}$  are the canonical vectors of  $\mathbb{R}^{n^{\mathfrak{c}}}$ . Since  $\psi^0 \neq 0$ (resp.  $b \neq 0$ ), we have  $J \neq \emptyset$  (resp.  $J \neq \{1, \ldots, n^{\mathfrak{c}}\}$ ). Then, we easily check that such a subspace is invariant by  $A^{\mathsf{T}}$  if, and only if,  $a_{rj} = 0$  for every  $r \in J$  and  $j \notin J$  and that it is included in ker  $b^{\mathsf{T}}$  if, and only if,  $b_r = 0$  for every  $j \in J$ .

443 444 445 2. Property (41) is a consequence of the uniqueness of the solution to the kernel equations. The orthogonality conditions are clearly satisfied under the additional assumption that  $b_r = 0$  for every  $r \in J$ .

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**Remark 5.3.** The first item in the above propositions gives explicit conditions that guarantee that the orthogonality conditions (15) hold (when combined with the results of the previous sections). We found these conditions with an algebraic approach. On the other hand, once these conditions are known, the second item of the above propositions show how to use them to obtain an analytic proof of the orthogonality conditions. Observe in addition that these different proofs are valid for arbitrary  $\delta \neq 1$ .

#### 453 5.2 Kernel associated with nontrivial rank two sequences

In the same spirit as in the previous section, we now we discuss the following property, related to Proposition 4.6:

$$E_1 \subset \ker b^{\mathsf{T}}, \quad \psi^r \in E_1, \quad \forall r \in \{2, \dots, 5\}.$$
 (42)

457 Below, we denote by  $c_0 = \langle v, \psi^0 \rangle$ .

**Proposition 5.4.** 1. Assume that rank  $(\psi^0|\psi^1) = 2$ ,  $E_1$  satisfies neither (28) nor (29),  $c_0 \neq 0$  and  $\langle b, \psi^0 \rangle = 0$ . Then, condition (42) holds if, and only if, there exists  $j_0 \in \{1, \ldots, n^c\}$  such that

$$b_{j_0} = 0, \quad \operatorname{rank} \Delta_{j_0} = 1,$$

where  $\Delta_{j_0} \in \mathbb{R}^{2n^c \times 2}$  is given by

$$\Delta_{j_0} = \begin{pmatrix} (D - d_{j_0})\psi^0 & A^{\mathsf{T}} e_{j_0} \\ A^{\mathsf{T}}\psi^0 & v_{j_0}\psi^0 - c_0 e_{j_0} \end{pmatrix},$$

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where  $d_{j_0}$  is the  $j_0$ -th diagonal entry of D and  $e_{j_0}$  is the  $j_0$ -th canonical vector of  $\mathbb{R}^{n^c}$ .

2. Assume that n = 4, i = 2,

$$b = \begin{pmatrix} 1\\ 0\\ -\rho \end{pmatrix}, \quad \rho \neq 0,$$

459  $c_0 \neq 0, \ \alpha_{24} \neq 0, \ \langle b, \psi^0 \rangle = 0 \ and \ rank \ \Delta_2 = 1.$  Then, the solution to the kernel equations 460 (18) with  $\delta = -1$  is given, for some  $\sigma \in \mathbb{R} \setminus \{0\}$ , by

$$\begin{cases} k_{21} = k_{24}\rho, \\ k_{22} = \frac{1}{-\sigma\alpha_{24}} \left( \sigma\mu_3 \frac{\partial k_{24}}{\partial x} + \sigma\nu_3 \frac{\partial k_{24}}{\partial \xi} - \alpha_{32}k_{24} \right), \\ k_{23} = \frac{1}{-\sigma\alpha_{24}} \left( \frac{\partial k_{24}}{\partial x} + \frac{\partial k_{24}}{\partial \xi} - \sigma\alpha_{23}k_{24} \right), \end{cases}$$
(43)

where  $k_{24} \in C^{\infty}(\mathbb{R}^2)$  is the solution to

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$$\begin{cases} \mu_{3} \frac{\partial^{2} k_{24}}{\partial x^{2}}(x,\xi) + (\mu_{3} + \nu_{3}) \frac{\partial^{2} k_{24}}{\partial x \partial \xi}(x,\xi) + \nu_{3} \frac{\partial^{2} k_{24}}{\partial \xi^{2}}(x,\xi) = -2c_{0}k_{24}(x,\xi), \\ \mu_{3} \frac{\partial k_{24}}{\partial x}(x,-x) + \nu_{3} \frac{\partial k_{24}}{\partial \xi}(x,-x) = \frac{\alpha_{32}}{\sigma}k_{24}(x,-x), \\ k_{24}(x,x) = \alpha_{24}. \end{cases}$$
(44)

To see that (44) indeed has a solution, we can introduce  $h = \mu_3 \frac{\partial k_{24}}{\partial x} + \nu_3 \frac{\partial k_{24}}{\partial \xi}$  and observe that ( $k_{24}, h$ ) satisfies a first-order hyperbolic system similar to the kernel equations (16) and whose well-posedness can be established as in the proof of Theorem 2.5.

Proof of Proposition 5.4. 1. Under the assumptions of the proposition and from the proof of Proposition 4.6, we have  $\psi^r \in E_1$  for every  $r \in \{2, \ldots, 5\}$  if, and only if, there exist  $\alpha_2, \beta_2 \in \mathbb{R}$  such that

$$\begin{cases} \psi^2 = \alpha_2 \psi^0 + \beta_2 \psi^1, \\ \psi^3 = \alpha_3 \psi^0 + \beta_3 \psi^1, \end{cases}$$

where  $\alpha_3, \beta_3$  are given by (38). We can check that this is equivalent to the existence of  $\rho, \theta \in \mathbb{R}$  such that

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$$\begin{cases} (A^{\mathsf{T}} - \rho)\psi^{1} = -c_{0}(D - \theta)\psi^{0} + \bar{c}_{1}(\psi^{1} - \rho\psi^{0}), \\ (D - \theta)(\psi^{1} - \rho\psi^{0}) = 0, \end{cases}$$
(45)

where we recall that  $\bar{c}_1 = \frac{\langle v, \psi^1 \rangle}{c_0}$ . Since *D* is diagonal with distinct entries and  $\psi^0, \psi^1$  are linearly independent, the second condition in (45) is equivalent to the existence of some  $j_0 \in \{1, \ldots, n^c\}$  and  $r \in \mathbb{R} \setminus \{0\}$  such that

$$\theta = d_{j_0}, \quad \psi^1 = \rho \psi^0 + r e_{j_0}.$$
 (46)

Plugging the second identity in the first condition in (45) and recalling that  $\psi^1 = A^{\mathsf{T}} \psi^0$ , we see that this condition simply becomes

$$rA^{\mathsf{T}}e_{j_0} = -c_0(D - d_{j_0})\psi^0 + r\bar{c}_1e_{j_0}$$

474 Comparing the  $j_0$ -th components of both quantities, using that the diagonal of A is zero 475 and  $r \neq 0$ , we see that  $\bar{c}_1 = 0$ . Recalling (46), the condition  $\bar{c}_1 = 0$  is equivalent to 476  $\rho = -\bar{r}v_{j_0}$  with  $\bar{r} = \frac{r}{c_0}$ . In summary, there exist  $\rho, \theta \in \mathbb{R}$  such that (45) holds if, and only 477 if, there exist  $j_0 \in \{1, \ldots, n^c\}$  such that

$$\exists \bar{r} \in \mathbb{R} \setminus \{0\}, \quad \begin{cases} (D - d_{j_0})\psi^0 + \bar{r}A^{\mathsf{T}}e_{j_0} = 0, \\ \psi^1 + \bar{r}(v_{j_0}\psi^0 - c_0e_{j_0}) = 0. \end{cases}$$
(47)

We can check that this condition is equivalent to rank  $\Delta_{j_0} = 1$ . Finally, it is clear that  $E_1 \subset \ker b^{\mathsf{T}}$ , i.e.  $\langle b, \psi^1 \rangle = 0$ , if, and only if,  $b_{j_0} = 0$ .

2. Since i = 2, we have (for the notations, see Section 3.1)

$$A^{\mathsf{T}} = \begin{pmatrix} 0 & \alpha_{31} & \alpha_{41} \\ \alpha_{13} & 0 & \alpha_{43} \\ \alpha_{14} & \alpha_{34} & 0 \end{pmatrix}, \quad D = 2\text{diag}\,(\mu_1, \mu_3, \mu_4) - \text{Id}_3, \quad \psi^0 = \begin{pmatrix} \alpha_{21} \\ \alpha_{23} \\ \alpha_{24} \end{pmatrix}, \quad v = -\frac{1}{2} \begin{pmatrix} \alpha_{12} \\ \alpha_{32} \\ \alpha_{42} \end{pmatrix}$$

481 Clearly,  $\langle b, \psi^0 \rangle = 0$  is equivalent to

$$\alpha_{21} = \alpha_{24}\rho. \tag{48}$$

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On the other hand, using the characterization (47), we see that rank  $\Delta_2 = 1$  if, and only if,

$$\alpha_{31} = \sigma(\mu_1 - \mu_3)\rho\alpha_{24},\tag{49}$$

$$\alpha_{34} = \sigma(\mu_4 - \mu_3)\alpha_{24},\tag{50}$$

$$\alpha_{41} = -\frac{\rho}{\sigma} \left( \alpha_{32} + \sigma^2 (\mu_1 - \mu_3) \alpha_{23} \right), \tag{51}$$

$$\alpha_{42} + \alpha_{12}\rho = \sigma(\alpha_{43} + \alpha_{13}\rho), \tag{52}$$

$$\alpha_{14} = -\frac{1}{\rho\sigma} \left( \alpha_{32} + \sigma^2 (\mu_4 - \mu_3) \alpha_{23} \right), \tag{53}$$

where  $\sigma = -\frac{2}{\bar{r}}$ . Using these conditions we easily check that  $((k^{\mathfrak{c}})^{\mathsf{T}}, k_2) = (k_{21}, k_{23}, k_{24}, k_{22})$ defined by (43)-(44) satisfies the kernel equations (18) with  $\delta = -1$ .

To conclude this section we will present a method which shows how conditions (48)-(53) can also be found from an analytic point of view.

# Another proof of Proposition 5.4, item 2. 1. For every j, let us denote by $\mathcal{P}_j$ the first-order linear partial differential operator

$$\mathcal{P}_j = \mu_j \frac{\partial}{\partial x} + \frac{\partial}{\partial \xi} \nu_j.$$

Since we want the orthogonality condition  $\langle b, k^{\mathfrak{c}}(\cdot, 0) \rangle = 0$ , we look for a solution satisfying

$$k_{21} = k_{24}\rho,$$

(recall also Remark 4.1). In particular, we assume (48). Then, the problem is to find a solution to  $(1 + 1)^{1/2}$ 

$$\begin{cases} k_{22}\alpha_{21} + k_{23}\alpha_{31} = (-\rho P_1 - \alpha_{41})k_{24}, \\ \mathcal{P}_2 k_{22} + k_{23}\alpha_{32} + k_{24}(\alpha_{42} + \alpha_{12}\rho) = 0, \\ \mathcal{P}_3 k_{23} + k_{22}\alpha_{23} + k_{24}(\alpha_{43} + \alpha_{13}\rho) = 0, \\ k_{22}\alpha_{24} + k_{23}\alpha_{34} = (-\mathcal{P}_4 - \alpha_{14}\rho)k_{24}. \end{cases}$$

$$(54)$$

Let us denote by

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$$\omega = \det \begin{pmatrix} \alpha_{21} & \alpha_{31} \\ \alpha_{24} & \alpha_{34} \end{pmatrix}.$$

Assume that  $\omega \neq 0$  (this will follow a posteriori from (49), (50), using also that  $\sigma, \rho, \alpha_{24} \neq 0$ ). Then, the first and fourth equations in (54) give

$$\begin{cases} k_{22} = \frac{1}{\omega} \left( -\alpha_{34}\rho \mathcal{P}_1 + \alpha_{31}\mathcal{P}_4 - \alpha_{34}\alpha_{41} + \alpha_{31}\alpha_{14}\rho \right) k_{24}, \\ k_{23} = \frac{1}{\omega} \left( \alpha_{24}\rho \mathcal{P}_1 - \alpha_{21}\mathcal{P}_4 + \alpha_{24}\alpha_{41} - \alpha_{21}\alpha_{14}\rho \right) k_{24}. \end{cases}$$
(55)

Plugging these relations in the second and third equations in (54) give the following two second-order partial differential equations for  $k_{24}$ :

$$\mathcal{Q}k_{24} = 0, \quad \tilde{\mathcal{Q}}k_{24} = 0,$$

where  $\mathcal{Q} = \mathcal{Q}^{(2)} + \mathcal{Q}^{(1)} + \mathcal{Q}^{(0)}$ , with

$$Q^{(2)} = \mathcal{P}_2 \left( -\alpha_{34}\rho \mathcal{P}_1 + \alpha_{31}\mathcal{P}_4 \right), Q^{(1)} = \left( -\alpha_{34}\alpha_{41} + \alpha_{31}\alpha_{14}\rho \right) \mathcal{P}_2 + \left( \alpha_{24}\rho \mathcal{P}_1 - \alpha_{21}\mathcal{P}_4 \right) \alpha_{32}, Q^{(0)} = \left( \alpha_{24}\alpha_{41} - \alpha_{21}\alpha_{14}\rho \right) \alpha_{32} + \omega (\alpha_{42} + \alpha_{12}\rho),$$

and  $\tilde{\mathcal{Q}} = \tilde{\mathcal{Q}}^{(2)} + \tilde{\mathcal{Q}}^{(1)} + \tilde{\mathcal{Q}}^{(0)}$ , with  $\tilde{\mathcal{Q}}^{(2)} = \mathcal{P}_3 \left( \alpha_{24} \rho \mathcal{P}_1 - \alpha_{21} \mathcal{P}_4 \right),$ 

$$\tilde{\mathcal{Q}}^{(1)} = (\alpha_{24}\alpha_{41} - \alpha_{21}\alpha_{14}\rho)\mathcal{P}_3 + (-\alpha_{34}\rho\mathcal{P}_1 + \alpha_{31}\mathcal{P}_4)\alpha_{23}$$
$$\tilde{\mathcal{Q}}^{(0)} = (-\alpha_{34}\alpha_{41} + \alpha_{31}\alpha_{14}\rho)\alpha_{23} + \omega(\alpha_{43} + \alpha_{13}\rho).$$

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 2. We are going to find conditions to guarantee that these two equations are compatible. To
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$$\mathcal{Q}^{(r)} = \sigma \tilde{\mathcal{Q}}^{(r)}, \quad r = 0, 1, 2, \tag{56}$$

for some  $\sigma \in \mathbb{R}$ . We first look at the operators of highest order. Using the identities

$$\alpha_{21} = \alpha_{24}\rho, \quad \mu_j - \nu_j = 1 \quad (j \neq 2), \quad \mu_2 = \nu_2 = 1,$$
(57)

we have

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$$\alpha_{24}\rho \mathcal{P}_1 - \alpha_{21}\mathcal{P}_4 = \alpha_{24}\rho \left(\mathcal{P}_1 - \mathcal{P}_4\right)$$
$$= \alpha_{24}\rho(\mu_1 - \mu_4)\mathcal{P}_2. \tag{58}$$

#### 499 It follows that

$$\tilde{\mathcal{Q}}^{(2)} = \alpha_{24} \rho(\mu_1 - \mu_4) \mathcal{P}_3 \mathcal{P}_2.$$
(59)

(60)

<sup>501</sup> Consequently, we see that (56) holds for r = 2 if we have

This identity holds if  $(\alpha_{34}, \alpha_{31})$  satisfies

$$\begin{pmatrix} -\rho\mu_1 & \mu_4 \\ -\rho\nu_1 & \nu_4 \end{pmatrix} \begin{pmatrix} \alpha_{34} \\ \alpha_{31} \end{pmatrix} = \sigma\alpha_{24}\rho(\mu_1 - \mu_4) \begin{pmatrix} \mu_3 \\ \nu_3 \end{pmatrix},$$

 $-\alpha_{34}\rho\mathcal{P}_1 + \alpha_{31}\mathcal{P}_4 = \sigma\alpha_{24}\rho(\mu_1 - \mu_4)\mathcal{P}_3.$ 

which is equivalent to (50)-(49) (using  $\rho \neq 0$  and (57)).

3. Let us now compute the first-order differential operators. We have

$$\mathcal{Q}^{(1)} = (-\alpha_{34}\alpha_{41} + \alpha_{31}\alpha_{14}\rho + \alpha_{24}\rho(\mu_1 - \mu_4)\alpha_{32})\mathcal{P}_2 \quad (by (58))$$
$$\tilde{\mathcal{Q}}^{(1)} = (\alpha_{24}\alpha_{41} - \alpha_{21}\alpha_{14}\rho + \sigma\alpha_{24}\rho(\mu_1 - \mu_4)\alpha_{23})\mathcal{P}_3 \quad (by (60)).$$

As a result, we have (56) for r = 1, if  $\mathcal{Q}^{(1)} = \tilde{\mathcal{Q}}^{(1)} = 0$ , that is, if

$$\begin{cases} -\alpha_{34}\alpha_{41} + \alpha_{31}\alpha_{14}\rho = -\alpha_{24}\rho(\mu_1 - \mu_4)\alpha_{32}, \\ \alpha_{24}\alpha_{41} - \alpha_{21}\alpha_{14}\rho = -\sigma\alpha_{24}\rho(\mu_1 - \mu_4)\alpha_{23}. \end{cases}$$
(61)

This holds if  $(\alpha_{41}, \alpha_{14})$  satisfies

$$\begin{pmatrix} -\alpha_{34} & \alpha_{31}\rho \\ \alpha_{24} & -\alpha_{21}\rho \end{pmatrix} \begin{pmatrix} \alpha_{41} \\ \alpha_{14} \end{pmatrix} = -\alpha_{24}\rho(\mu_1 - \mu_4) \begin{pmatrix} \alpha_{32} \\ \sigma\alpha_{23} \end{pmatrix},$$

which is equivalent to (51) and (53) (using (48), (49), (50) and  $\alpha_{24}, \rho \neq 0$ ).

4. Let us now compute the zero order terms. Using (61), we immediately see that

$$\tilde{\mathcal{Q}}^{(0)} = -\alpha_{24}\rho(\mu_1 - \mu_4)\alpha_{32}\alpha_{23} + \omega(\alpha_{43} + \alpha_{13}\rho), 
\mathcal{Q}^{(0)} = -\sigma\alpha_{24}\rho(\mu_1 - \mu_4)\alpha_{23}\alpha_{32} + \omega(\alpha_{42} + \alpha_{12}\rho)$$

As a result, we see that (56) holds for r = 0 if we have condition (52). Moreover, using 507 (48), (49) and (50), we have508  $\omega = -\sigma \alpha_{24}^2 \rho(\mu_1 - \mu_4),$ (62)

so that, using again (48) and the definition of  $c_0$ , we obtain

$$Q^{(0)} = \sigma \alpha_{24} \rho (\mu_1 - \mu_4) (2c_0).$$

It follows that  $k_{24}$  indeed satisfies the first equation in (44) (recall that  $Q^{(2)} = \sigma \tilde{Q}^{(2)}$  with 510 (59) and  $Q^{(1)} = 0$ ). 511

5. Using (60), (61), (58) and (62), we can simplify the expressions in (55) to obtain

$$\begin{cases} k_{22} = \frac{1}{-\sigma\alpha_{24}} \left(\sigma \mathcal{P}_3 - \alpha_{32}\right) k_{24}, \\ k_{23} = \frac{1}{-\sigma\alpha_{24}} \left(\mathcal{P}_2 - \sigma\alpha_{23}\right) k_{24}. \end{cases}$$

In addition, it follows from these formula that the remaining conditions are satisfied. In-512 deed, the condition  $k_{22}(x, -x) = 0$  is exactly the condition that we require for  $k_{24}$  at 513 (x, -x) in (44) and the condition  $k_{23}(x, x) = \alpha_{23}$  follows from the above expression since 514  $k_{24}(x,x) = \alpha_{24}$  and  $(\mathcal{P}_2 k_{24})(x,x) = \frac{d}{dx} k_{24}(x,x) = 0.$ 515 516

Remark 5.5. In [VK14, Section 3.3], the authors showed that we can solve a kernel system of two equations of the form

$$\begin{cases} \frac{\partial k_{21}}{\partial x} - \frac{\partial k_{21}}{\partial \xi} + k_{22}\alpha_{21} = 0, \\ \frac{\partial k_{22}}{\partial x} + \frac{\partial k_{22}}{\partial \xi} + k_{21}\alpha_{12} = 0, \\ k_{21}(x, x) = \alpha_{21}, \quad k_{22}(x, 0) = 0 \end{cases}$$

with  $\alpha_{21} \neq 0$  by first expressing  $k_{22}$  from the first equation and then showing that the resulting 517 second order equation for  $k_{21}$  indeed has a solution. The method we introduced in the second 518 proof of Proposition 5.4, item 2, can be seen as an extension of the method of [VK14] where, 519 instead of dividing by a scalar (namely,  $\alpha_{21}$ ), we invert a matrix. 520

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#### <sup>526</sup> A Controllability of the equivalent system

In this appendix, we give a simple and direct proof of Corollary 2.4. We recall that it can be deduced from Theorem 2.3 but this result is based on the Titchmarsh convolution theorem (see [HO21b]) and we show here how to directly prove the corollary without resorting to this difficult result.

Proof of Corollary 2.4. It is enough to show that, if  $(q, f) \in S_k \setminus S_{k+1}$  for some  $k \in \{2, \ldots, n+1\}$ , then system (9) (with m = 1) is null controllable in time T if, and only if,  $T \ge \tau_k$ .

1. We first observe that system (9) is equivalent to the same system with  $f_1 = 0$ . This follows from the invertible spatial transformation

$$\hat{y}_1(t,x) = \tilde{y}_1(t,x) - \int_0^x h(x-\xi)\tilde{y}_1(t,\xi)\,d\xi$$

where the kernel h is the solution to

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$$h(x)\lambda_1 + \int_0^x h(x-\xi)f_1(\xi) \,d\xi = f_1(x), \quad 0 < x < 1.$$

- Therefore, for the rest of the proof, we assume that  $f_1 = 0$ .
  - 2. Assume now that  $(q, f) \in S_k \setminus S_{k+1}$  for some  $k \in \{2, \ldots, n\}$  (the result for k = n + 1 is trivial). It will be convenient to use the notation  $\hat{q}_i = q_{i-1}$  for  $2 \leq i \leq n$ . Let us write system (9) (with  $f_1 = 0$ ) component-wise:

$$\begin{cases} \frac{\partial \tilde{y}_1}{\partial t}(t,x) + \lambda_1 \frac{\partial \tilde{y}_1}{\partial x}(t,x) = 0, \\ \tilde{y}_1(t,1) = \tilde{u}(t), \\ \tilde{y}_1(0,x) = \tilde{y}_1^0(x), \end{cases} \begin{cases} \frac{\partial \tilde{y}_i}{\partial t}(t,x) + \lambda_i \frac{\partial \tilde{y}_i}{\partial x}(t,x) = f_i(x)v(t), \\ \tilde{y}_i(t,0) = \hat{q}_iv(t), \\ \tilde{y}_i(0,x) = \tilde{y}_i^0(x), \end{cases}$$

for  $i \in \{2, ..., n\}$ , and where we introduced  $v(t) = \tilde{y}_1(t, 0)$ . It is clear that this system is null controllable in any time  $T \ge \tau_k = \max\{T_1 + T_k, T_2\}$  since in this case taking  $\tilde{u} = 0$  in  $(T - (T_1 + T_k), T)$  does the job. It is the necessary part that requires more work.

3. First of all, we recall that the condition  $T \ge \max\{T_1, T_2\}$  is always necessary (see e.g. the proof of [HO21b, Lemma 3.3]). Under this condition and by mimicking the second step in the proof of [HO21b, Theorem 3.1], we see that the null controllability condition  $\tilde{y}_k(T, x) = 0$  is equivalent to

$$\hat{q}_k \alpha(\tau) + \int_0^\tau \beta(\tau - \sigma) \alpha(\sigma) \, d\sigma = 0, \quad 0 < \tau < T_k, \tag{63}$$

where 
$$\alpha(\theta) = v(-\theta + T)$$
 and  $\beta(\theta) = f_k(\lambda_k \theta)$  for  $0 < \theta < T_k$ .

<sup>543</sup> 4. We now have two possibilities for (63).

(a) Case  $\hat{q}_k \neq 0$ . Then, by uniqueness of the solution to the Volterra equation of the second kind (63), we obtain  $\alpha = 0$  in  $(0, T_k)$ . This means that v = 0 in  $(T - T_k, T)$ . Since this is true for any  $\tilde{y}_0^1$ , it is possible only if  $T_1 \leq T - T_k$ , which is the desired condition. (b) Case  $\hat{q}_k = 0$ . Since  $(q, f) \in \mathcal{S}_k \setminus \mathcal{S}_{k+1}$ , we necessarily have  $f_k \neq 0$ . Since  $f_k$  is analytic in a neighborhood of [0, 1), this implies in particular that there exists  $N \geq 1$  such that

$$f_k^{(N-1)}(0) \neq 0, \quad f_k^{(\ell)}(0) = 0, \quad \forall \ell < N-1.$$

Then, taking N times the derivative with respect to  $\tau$  in (63) (with  $\hat{q}_k = 0$ ), we obtain the new Volterra equation

$$c\alpha(\tau) + \int_0^\tau \beta^{(N)}(\tau - \sigma)\alpha(\sigma) \, d\sigma = 0, \quad 0 < \tau < T_k$$

where  $c = \beta^{(N-1)}(0) = f_k^{(N-1)}(0)\lambda_k^{N-1}$ . Therefore,  $c \neq 0$  and the situation is now identical to the previous case.

#### В Solution to the kernel equations 551

In this appendix, we present a new approach to solve the kernel equations that encompasses 552 in particular the proof of Theorem 2.5. We recall that, when considering the kernel equations 553 in the triangle  $\mathcal{T} = \{(x,\xi) \in \mathbb{R}^2 \mid 0 < \xi < x < 1\}$ , the approach used in all current results 554 in the literature ([Cor+13; DVK13; HD15; Hu+16; Hu+19; CN19], etc.) consists in adding 555 "artificial boundary conditions" to close the system of kernel equations. In our approach, we will 556 not consider the condition at (x, x) as a boundary condition but rather as an initial condition. 557 We will simply let propagate this condition along the characteristics and find the corresponding 558 so-called domain of determinacy, much in the spirit of the reference books [LY85; Bre00]. Then, 559 another idea of our method is also to solve the equation for j = i and plug it into the other 5 60 equations of the system to obtain a new system with initial conditions at (x, x) only (as in the 561 proof of Theorem 3.1). Moreover, this gives a natural bound in  $|x - \xi|$  for the estimates needed 562 to prove the contraction of the mapping defining the integral equations corresponding to the new 563 system (rather than  $|x - (1 - \varepsilon)\xi|$  as in [Hu+16; Hu+19]). 564

All along this appendix,  $i \in \{1, \ldots, n\}$  is fixed and we continue using the notation k =565  $(k_{ij})_{1 \le j \le n}$  to denote the transpose of the *i*-th row of K. We also emphasize that  $m \ge 1$  is 566 arbitrary. 567

First of all, it will be more convenient to work with the kernel equations normalized by  $\lambda_i$ : 568

$$\begin{cases} \frac{\partial k_j}{\partial x}(x,\xi) + \bar{\lambda}_j \frac{\partial k_j}{\partial \xi}(x,\xi) + \sum_{r=1}^n k_r(x,\xi)\bar{m}_{rj} = 0, \\ k_j(x,x) = f_j \quad (j \neq i), \quad k_i(x,\delta x) = 0, \end{cases}$$
(64)

where

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$$\bar{\lambda}_j = \frac{\lambda_j}{\lambda_i}, \quad \bar{m}_{rj} = \frac{m_{rj}}{\lambda_i}, \quad f_j = \frac{m_{ij}}{\lambda_i - \lambda_j}$$

From now on, we will assume for instance that  $i \ge m+1$ , so that  $\lambda_i > 0$  and thus, from (1b), 570 571  $\bar{\lambda}$ 

$$\bar{\lambda}_1 < \dots < \bar{\lambda}_{i-1} < 1 < \bar{\lambda}_{i+1} < \dots < \bar{\lambda}_n.$$
(65)

For every  $(x,\xi) \in \mathbb{R} \times \mathbb{R}$ , we denote by  $s \mapsto \zeta_j(s;x,\xi)$  the solution to

$$\begin{cases} \frac{d}{ds}\zeta_j(s;x,\xi) = \bar{\lambda}_j, \quad \forall s \in \mathbb{R} \\ \zeta_j(x;x,\xi) = \xi. \end{cases}$$

Let us now consider the more general condition

$$k_j(x,x) = f_j(x) \quad (j \neq i),$$

where  $f_j$  is a function defined on an interval of the form [a, b] with a < 0 < b. Even if  $f_j$  is constant in (64), we will need to consider space-dependent data to deduce the existence of smooth solutions by an inductive argument. We will describe the largest domain  $D \subset \mathbb{R}^2$  where the system can then be solved along the characteristics. We first take care of the characteristics for  $j \neq i$ . Recalling the ordering (65), we introduce

$$D^{\mathfrak{c}} = \left\{ (x,\xi) \in \mathbb{R}^2 \mid \qquad \zeta_{i-1}(x;a,a) < \xi < \zeta_{i-1}(x;b,b) \\ \zeta_{i+1}(x;b,b) < \xi < \zeta_{i+1}(x;a,a) \end{array} \right\},\$$

(see Figure 1). Above, we use the usual conventions for i = 1 and i = n. We now take care of the characteristic for j = i. We can check that the line  $\{(x, \delta x) \mid a < x < b\}$  intersects the boundary of  $D^{c}$  at exactly two points  $(c, \delta c)$  and  $(d, \delta d)$ , with  $c, d \in (a, b)$  and c < 0 < d if  $\delta < 1$  or d < 0 < c if  $\delta > 1$ . Let then

$$D = \{ (x,\xi) \in D^{\mathfrak{c}} \mid \zeta_i(x;d,\delta d) < \xi < \zeta_i(x;c,\delta c) \},\$$

(see Figure 2 with  $\delta = -1$ ) and define I = (c, d). Here and in what follows, it will be convenient to use the notation  $(\alpha, \beta)$  to denote the interval  $(\min \{\alpha, \beta\}, \max \{\alpha, \beta\})$ , whatever  $\alpha, \beta \in \mathbb{R}$  are (we use a similar notation for  $[\alpha, \beta]$ ).



Figure 1: Domain  $D^{\mathfrak{c}}$ 

Figure 2: Domain D (in dark gray)

x

We will prove the following result.

Theorem B.1. Let a < 0 < b and  $s \in \mathbb{N}$  be fixed. For any  $(f_j)_{j \neq i} \in C^s([a, b])^{n-1}$  and  $f_i \in C^s(\overline{I})$ , there exists a unique solution  $k = (k_j)_{1 \leq j \leq n} \in C^s(\overline{D})^n$  to

$$\begin{cases} \frac{\partial k_j}{\partial x}(x,\xi) + \bar{\lambda}_j \frac{\partial k_j}{\partial \xi}(x,\xi) + \sum_{r=1}^n k_r(x,\xi)\bar{m}_{rj} = 0, \quad (x,\xi) \in D, \\ k_j(x,x) = f_j(x), \quad x \in (a,b) \quad (j \neq i), \\ k_i(x,\delta x) = f_i(x), \quad x \in I, \end{cases}$$
(66)

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580 Moreover, we have the estimate

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$$\|k\|_{C^{s}(\overline{D})^{n}} \leq C \max\left\{\max_{j \neq i} \|f_{j}\|_{C^{s}([a,b])}, \quad \|f_{i}\|_{C^{s}(\overline{I})}\right\},$$
(67)

for some C > 0 that does not depend on any  $f_j$ .

For s = 0, by solution we mean "solution along the characteristics", see below. The first part of Theorem 2.5 follows from the previous result and the following simple observation:

 $\forall V \subset \mathbb{R}^2, \exists a < 0 < b, \quad V \subset D.$ (68)

On the other hand, using that the coefficients of the system are constant and arguing as in the proof of [CN19, Lemma 6.2], we can show that (67) holds with  $C = R^s$  for some R > 0 that does not depend on s. This establishes the estimate in Theorem 2.5.

Let us now prove Theorem B.1. We start with a description of the key properties satisfied by the point where the j-th characteristic intersects the corresponding data line.

Lemma B.2. For every  $j \in \{1, ..., n\}$ , there exists  $\sigma_j \in C^{\infty}(\overline{D})$  such that, for every  $(x, \xi) \in \overline{D}$ , we have:

• 
$$(s, \zeta_j(s; x, \xi)) \in \overline{D}$$
 for every  $s \in [\sigma_j(x, \xi), x]$ 

• For every  $j \neq i$ , we have

$$\left|\sigma_{j}(x,\xi) - x\right| \le C \left|x - \xi\right|,\tag{69}$$

for some 
$$C > 0$$
 that does not depend on  $j, x, \xi$ 

We point out that  $\zeta_j$  and  $\sigma_j$  are explicit. In particular, this is how we prove estimate (69). Now, instead of writing (66) along all the characteristics (as it is usually done), we first replace  $k_i$  by formally solving the corresponding equation (recall that  $\bar{m}_{ii} = 0$ ):

$$k_i(x,\xi) = f_i(\sigma_i(x,\xi)) - \int_{\sigma_i(x,\xi)}^x \sum_{r \neq i} k_r(\eta, \zeta_i(\eta; x,\xi)) \bar{m}_{ri} \, d\eta.$$
(70)

Let us introduce the following notations to exclude the *i*-th components:  $k^{\mathfrak{c}} = (k_j)_{j \neq i}$ ,  $f^{\mathfrak{c}} = (f_j)_{j \neq i}$ ,  $\sigma^{\mathfrak{c}} = (\sigma_j)_{j \neq i}$ ,  $\zeta^{\mathfrak{c}} = (\zeta_j)_{j \neq i}$ ,  $M^{\mathfrak{c}} = (\bar{m}_{rj})_{r,j \neq i}$ ,  $\psi = (\bar{m}_{ij})_{j \neq i}$  and  $w = (\bar{m}_{ji})_{j \neq i}$ . Then, plugging the previous expression of  $k_i$  in (66) and integrating along the characteristics, we can transform this system into the following system of integral equations for  $k^{\mathfrak{c}}$ :

$$k_{\ell}^{\mathfrak{c}}(x,\xi) = f_{\ell}^{\mathfrak{c}}(\sigma_{\ell}^{\mathfrak{c}}(x,\xi)) - \int_{\sigma_{\ell}^{\mathfrak{c}}(x,\xi)}^{x} \sum_{r=1}^{n-1} k_{r}^{\mathfrak{c}}(s,\zeta_{\ell}^{\mathfrak{c}}(s;x,\xi)) m_{r\ell}^{\mathfrak{c}} ds$$
$$- \int_{\sigma_{\ell}^{\mathfrak{c}}(x,\xi)}^{x} f_{i}(\sigma_{i}^{\mathfrak{c}}(s,\zeta_{\ell}^{\mathfrak{c}}(s;x,\xi))) \psi_{\ell} ds$$
$$+ \int_{\sigma_{\ell}^{\mathfrak{c}}(x,\xi)}^{x} \left( \int_{\sigma_{i}^{\mathfrak{c}}(s,\zeta_{\ell}^{\mathfrak{c}}(s;x,\xi))}^{s} \sum_{r=1}^{n-1} k_{r}^{\mathfrak{c}}(\eta,\zeta_{i}(\eta;s,\zeta_{\ell}^{\mathfrak{c}}(s;x,\xi))) w_{r} d\eta \right) \psi_{\ell} ds, \qquad (71)$$

for every  $\ell \in \{1, \ldots, n-1\}$  and  $(x, \xi) \in \overline{D}$ .

All the quantities in (70) and (71) are well defined thanks to Lemma B.2. It remains to prove the existence and uniqueness of a  $C^s$  solution  $k^{\mathfrak{c}}$  to this system of integral equations. We start with s = 0. As usual, we use the Banach fixed point theorem and suitable estimates. A solution to this system is a fixed point of the map  $F(k^{\mathfrak{c}}) = u^0 + \Phi k^{\mathfrak{c}}$ , where

$$u_{\ell}^{0}(x,\xi) = f_{\ell}^{\mathfrak{c}}(\sigma_{\ell}^{\mathfrak{c}}(x,\xi)) - \int_{\sigma_{\ell}^{\mathfrak{c}}(x,\xi)}^{x} f_{i}(\sigma_{i}^{\mathfrak{c}}(s,\zeta_{\ell}^{\mathfrak{c}}(s;x,\xi))\psi_{\ell}\,ds,$$

and  $\Phi$  is the linear map  $\Phi = \Phi_1 + \Phi_2$  with

$$(\Phi_1 k^{\mathfrak{c}})_{\ell}(x,\xi) = -\int_{\sigma_{\ell}^{\mathfrak{c}}(x,\xi)}^{x} \sum_{r=1}^{n-1} k_r^{\mathfrak{c}}(s,\zeta_{\ell}^{\mathfrak{c}}(s;x,\xi)) m_{r\ell}^{\mathfrak{c}} ds,$$

and

$$(\Phi_2 k^{\mathfrak{c}})_{\ell}(x,\xi) = \int_{\sigma_{\ell}^{\mathfrak{c}}(x,\xi)}^{x} \left( \int_{\sigma_i^{\mathfrak{c}}(s,\zeta_{\ell}^{\mathfrak{c}}(s;x,\xi))}^{s} \sum_{r=1}^{n-1} k_r^{\mathfrak{c}}(\eta,\zeta_i(\eta;s,\zeta_{\ell}^{\mathfrak{c}}(s;x,\xi))) w_r \, d\eta \right) \psi_{\ell} \, ds,$$

for every  $\ell \in \{1, \ldots, n-1\}$  and  $(x, \xi) \in \overline{D}$ .

Let us now precisely set the functional framework. Let  $B = C^0(\overline{D})^{n-1}$  and consider the standard norm  $||k^{\mathfrak{c}}||_B = \max_{1 \le \ell \le n-1} \max_{(x,\xi) \in \overline{D}} |k^{\mathfrak{c}}_{\ell}(x,\xi)|$ . Clearly, B is a Banach space and  $F(B) \subset B$ . Let us now prove that  $F^N$  is a contraction for  $N \in \mathbb{N}^*$  large enough. This is equivalent to show that  $\Phi^N$  is a contraction. To this end, it is sufficient to prove the following key estimate:

**Lemma B.3.** There exists C > 0 such that, for every  $N \in \mathbb{N}^*$ , we have

$$\left| (\Phi^N k^{\mathfrak{c}})_{\ell}(x,\xi) \right| \leq \frac{C^N \left| x - \xi \right|^N}{N!} \left\| k^{\mathfrak{c}} \right\|_B$$

611 for every  $k^{\mathfrak{c}} \in B$ ,  $\ell \in \{1, \ldots, n-1\}$  and  $(x, \xi) \in \overline{D}$ .

*Proof.* We prove the property by induction on N. Let us first consider N = 1. We have

$$\left| (\Phi_1 k^{\mathfrak{c}})_{\ell}(x,\xi) \right| \leq C_1 \left| x - \sigma_{\ell}^{\mathfrak{c}}(x,\xi) \right| \left\| k^{\mathfrak{c}} \right\|_B,$$

with  $C_1 = \max_{\ell} \sum_r |m_{r\ell}^{\mathfrak{c}}| \ge 0$ . Similarly,

$$\left| (\Phi_2 k^{\mathfrak{c}})_j(x,\xi) \right| \le C_2 \left| x - \sigma_{\ell}^{\mathfrak{c}}(x,\xi) \right| \left\| k^{\mathfrak{c}} \right\|_B$$

with  $C_2 = \max_{\ell, (x,\xi), s} |s - \sigma_i^{\mathfrak{c}}(s, \zeta_{\ell}^{\mathfrak{c}}(s; x, \xi))| \sum_r |w_r| |\psi_{\ell}| \ge 0$ . Finally, we have

 $|x - \sigma_{\ell}^{\mathfrak{c}}(x,\xi)| \le C_3 |x - \xi|, \qquad (72)$ 

for some  $C_3 > 0$  that does not depend on  $\ell, x, \xi$  (see (69)). This proves the property for N = 1. Let us now assume that the property holds for N and let us prove it for N + 1. We have

$$\left| (\Phi_1 \Phi^N k^{\mathfrak{c}})_{\ell}(x,\xi) \right| \le \int_{[\sigma_{\ell}^{\mathfrak{c}}(x,\xi),x]} \sum_{r=1}^{n-1} \left| (\Phi^N k^{\mathfrak{c}})_r(s,\zeta_{\ell}^{\mathfrak{c}}(s;x,\xi)) \right| |m_{r\ell}^{\mathfrak{c}}| ds,$$

Using the induction assumption, we get

$$\left| (\Phi_1 \Phi^N k^{\mathfrak{c}})_{\ell}(x,\xi) \right| \le C_1 \frac{C^N}{N!} \left\| k^{\mathfrak{c}} \right\|_B \int_{[\sigma_{\ell}^{\mathfrak{c}}(x,\xi),x]} \left| s - \zeta_{\ell}^{\mathfrak{c}}(s;x,\xi) \right|^N \, ds.$$

Similarly, noting that  $\eta - \zeta_i(\eta; s, \zeta_\ell^{\mathfrak{c}}(s; x, \xi)) = s - \zeta_\ell^{\mathfrak{c}}(s; x, \xi)$ , we get

$$\left| (\Phi_2 \Phi^N k^{\mathfrak{c}})_{\ell}(x,\xi) \right| \le C_2 \frac{C^N}{N!} \left\| k^{\mathfrak{c}} \right\|_B \int_{[\sigma_{\ell}^{\mathfrak{c}}(x,\xi),x]} \left| s - \zeta_{\ell}^{\mathfrak{c}}(s;x,\xi) \right|^N \, ds$$

Now observe that  $|s - \zeta_{\ell}^{\mathfrak{c}}(s; x, \xi)| \leq C_4 |s - \sigma_{\ell}^{\mathfrak{c}}(x, \xi)|$  for some  $C_4 > 0$  that does not depend on  $\ell, s, x, \xi$ . It follows that

$$\int_{[\sigma_{\ell}^{\mathsf{c}}(x,\xi),x]} |s - \zeta_{\ell}^{\mathsf{c}}(s;x,\xi)|^{N} ds \le C_{4} \int_{[\sigma_{\ell}^{\mathsf{c}}(x,\xi),x]} |s - \sigma_{\ell}^{\mathsf{c}}(x,\xi)|^{N} ds = C_{4} \frac{|x - \sigma_{\ell}^{\mathsf{c}}(x,\xi)|^{N+1}}{N+1}.$$

We conclude thanks to the estimate (72).

Finally, the estimate (67) can be deduced from the identities  $k^{\mathfrak{c}} = F^N(k^{\mathfrak{c}}) - F^N(0) + F^N(0) = \Phi^N(k^{\mathfrak{c}}) - \Phi^N(0) + \sum_{r=0}^N \Phi^r u^0$ , combined with the fact that  $\Phi^N$  is a contraction and that  $u^0$  can be estimated by the right-hand side of (67) (with s = 0). This concludes the proof of Theorem B.1 for s = 0.

To prove the result for  $s \ge 1$  we can argue as in the proof of [Bre00, Theorem 3.6] and then use an induction argument.

**Remark B.4.** The proof above can be adapted to deal with space-dependent systems, i.e. when  $\lambda_j$  and  $m_{rj}$  depend on x. The additional condition for  $k_i$  has to be modified though, but we can for instance consider  $k_i(x,0) = f_i(x)$ . Note that we still have explicit formulas for the corresponding  $\zeta_j$  and  $\sigma_j$ .

**Remark B.5.** Our approach can be used to recover existence results in the triangle  $\mathcal{T}$ . To this end, we simply extend the parameters  $\lambda_j$  and  $m_{rj}$  to  $[a, b] \supset [0, 1]$  in a smooth way. Then, for *a, b* large enough, the domain D will contain the triangle  $\mathcal{T}$  (recall (68)) and we apply Theorem B.1 in this D. This approach is different from all the previous ones in the literature, which consisted in adding "artificial boundary conditions" at some parts of the boundary of  $\mathcal{T}$ . Note in addition that extending  $\lambda_j$  and  $m_{rj}$  outside [0, 1] in a smooth way is easier than building artificial boundary conditions that satisfy compatibility conditions associated with the kernel equations.

#### **References**

- 635 [BC16] G. Bastin and J.-M. Coron. Stability and boundary stabilization of 1-D hyperbolic systems.
  636 Vol. 88. Progress in Nonlinear Differential Equations and their Applications. Subseries in
  637 Control. Birkhäuser/Springer, [Cham], 2016, pp. xiv+307.
- (Bre00] A. Bressan. Hyperbolic systems of conservation laws. Vol. 20. Oxford Lecture Series in Mathematics and its Applications. The one-dimensional Cauchy problem. Oxford University Press, Oxford, 2000, pp. xii+250.
- [Cor+21] J.-M. Coron, L. Hu, G. Olive, and P. Shang. "Boundary stabilization in finite time of onedimensional linear hyperbolic balance laws with coefficients depending on time and space".
   In: J. Differential Equations 271 (2021), pp. 1109–1170. DOI: 10.1016/j.jde.2020.09.037.
- [CN19] J.-M. Coron and H.-M. Nguyen. "Optimal time for the controllability of linear hyperbolic systems in one-dimensional space". In: SIAM J. Control Optim. 57.2 (2019), pp. 1127–1156.
  DOI: 10.1137/18M1185600.
- G47 [CN20] J.-M. Coron and H.-M. Nguyen. "Finite-time stabilization in optimal time of homogeneous quasilinear hyperbolic systems in one dimensional space". In: ESAIM Control Optim. Calc.
   G49 Var. 26 (2020), Paper No. 119, 24. DOI: 10.1051/cocv/2020061.

- [CN21a] J.-M. Coron and H.-M. Nguyen. "Null-controllability of linear hyperbolic systems in one dimensional space". In: Systems Control Lett. 148 (2021), p. 104851. DOI: 10.1016/j.
   sysconle.2020.104851.
- 653 [CN21b] J.-M. Coron and H.-M. Nguyen. On the optimal controllability time for linear hyperbolic 654 systems with time-dependent coefficients. 2021. arXiv: 2103.02653.
- [CN22] J.-M. Coron and H.-M. Nguyen. "Lyapunov functions and finite-time stabilization in optimal time for homogeneous linear and quasilinear hyperbolic systems". In: Ann. Inst. H. Poincaré *C Anal. Non Linéaire* 39.5 (2022), pp. 1235–1260. DOI: 10.4171/aihpc/30.
- (DVK13)
   F. Di Meglio, R. Vazquez, and M. Krstic. "Stabilization of a system of n + 1 coupled first-order hyperbolic linear PDEs with a single boundary input". In: *IEEE Trans. Automat. Control* 58.12 (2013), pp. 3097–3111. DOI: 10.1109/TAC.2013.2274723.
- [Hu15] L. Hu. "Sharp time estimates for exact boundary controllability of quasilinear hyperbolic systems". In: SIAM J. Control Optim. 53.6 (2015), pp. 3383-3410. DOI: 10.1137/140983720.
- L. Hu and F. Di Meglio. "Finite-time backstepping boundary stabilization of 3 × 3 hyperbolic systems". In: 2015 European Control Conference (ECC). 2015, pp. 67–72. DOI: 10.1109/ECC.
   2015.7330527.
- (Hu+16)
   L. Hu, F. Di Meglio, R. Vazquez, and M. Krstic. "Control of homodirectional and general heterodirectional linear coupled hyperbolic PDEs". In: *IEEE Trans. Automat. Control* 61.11 (2016), pp. 3301–3314. DOI: 10.1109/TAC.2015.2512847.
- [HO21a] L. Hu and G. Olive. "Minimal time for the exact controllability of one-dimensional first-order linear hyperbolic systems by one-sided boundary controls". In: J. Math. Pures Appl. (9) 148 (2021), pp. 24-74. DOI: 10.1016/j.matpur.2021.02.009.
- [HO21b] L. Hu and G. Olive. "Null controllability and finite-time stabilization in minimal time of one-dimensional first-order 2 × 2 linear hyperbolic systems". In: ESAIM Control Optim. Calc. Var. 27 (2021), Paper No. 96, 18. DOI: 10.1051/cocv/2021091.
- Example 1 (HO22)
   L. Hu and G. Olive. "Equivalent one-dimensional first-order linear hyperbolic systems and range of the minimal null control time with respect to the internal coupling matrix". In: J. Differential Equations 336 (2022), pp. 654-707. DOI: 10.1016/j.jde.2022.07.023.
- [Hu+19] L. Hu, R. Vazquez, F. Di Meglio, and M. Krstic. "Boundary exponential stabilization of 1-dimensional inhomogeneous quasi-linear hyperbolic systems". In: SIAM J. Control Optim. 57.2 (2019), pp. 963–998. DOI: 10.1137/15M1012712.
- [LY85] T. T. Li and W. C. Yu. Boundary value problems for quasilinear hyperbolic systems. Duke
   University Mathematics Series, V. Duke University, Mathematics Department, Durham, NC, 1985, pp. viii+325.
- <sup>667</sup> [Li10] T. Li. Controllability and observability for quasilinear hyperbolic systems. Vol. 3. AIMS Series
   <sup>688</sup> on Applied Mathematics. American Institute of Mathematical Sciences (AIMS), Springfield,
   <sup>689</sup> MO; Higher Education Press, Beijing, 2010, pp. x+222.
- LR10] T. Li and B. Rao. "Strong (weak) exact controllability and strong (weak) exact observability
   for quasilinear hyperbolic systems". In: Chin. Ann. Math. Ser. B 31.5 (2010), pp. 723-742.
   DOI: 10.1007/s11401-010-0600-9.
- [MA22] Y. Mokhtari and F. Ammar-Khodja. "Boundary controllability of two coupled wave equations with space-time first-order coupling in 1-D". In: J. Evol. Equ. 22.2 (2022), Paper No. 31, 52. DOI: 10.1007/s00028-022-00790-x.
- [Rus78] D. L. Russell. "Controllability and stabilizability theory for linear partial differential equations: recent progress and open questions". In: SIAM Rev. 20.4 (1978), pp. 639-739. DOI: 10.1137/1020095.

699 [VK14]	R. Vazquez and M. Krstic. "Marcum Q-functions and explicit kernels for stabilization of $2 \times 2$
700	linear hyperbolic systems with constant coefficients". In: Systems Control Lett. 68 (2014),
701	pp. 33-42. DOI: 10.1016/j.sysconle.2014.02.008.

702 [Wec82]	N. Weck. "A remark on controllability for symmetric hyperbolic systems in one space di-
703	mension". In: SIAM J. Control Optim. 20.1 (1982), pp. 1–8. DOI: 10.1137/0320001.