CHARACTERISTIC MATRICES FOR LINEAR PERIODIC DELAY DIFFERENTIAL EQUATIONS

JAN SIEBER∗ AND ROBERT SZALAI†

Abstract. Szalai et al. (SIAM J. on Sci. Comp. 28(4), 2006) gave a general construction for characteristic matrices for systems of linear delay-differential equations with periodic coefficients. First, we show that matrices constructed in this way can have a discrete set of poles in the complex plane, which may possibly obstruct their use when determining the stability of the linear system. Then we modify and generalize the original construction such that the poles get pushed into a small neighborhood of the origin of the complex plane.

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1. Introduction. Linear delay differential equations (DDEs) with coefficients periodic in time come up naturally when one analyses nonlinear DDEs: one of the most common and simplest invariant sets encountered in nonlinear systems are periodic orbits [5]. If one wants to find out if a periodic orbit is dynamically stable or unstable, or how its stability changes as one changes system parameters one has to linearize the nonlinear DDE system in the periodic orbit, which results in a linear DDE with time-periodic coefficients [2, 6]. This linear DDE is typically written in the form

\[ \dot{x}(t) = L(t)x_t \]  (1.1)

where the coefficients (hidden in the operator \( L(t) \)) are periodic in \( t \). Without loss of generality we can assume that the time dependence of \( L \) is of period 1 (this corresponds to rescaling time and \( L \)). Then the exponential asymptotic stability of the periodic orbit in the full nonlinear system is determined by the eigenvalues of the time-1 map \( T \) generated by the linear DDE (1.1). See the textbooks [2, 6] for a summary of the fundamental results on Floquet theory and [11] for robust numerical methods to calculate these eigenvalues in practical problems. For analytical purposes (and, possibly, for numerical purposes) it is useful to reduce the eigenvalue problem for the infinite-dimensional time-1 map \( T \) to an algebraic problem via a characteristic matrix \( \Delta(\lambda) \in \mathbb{C}^{n \times n} \). One would expect that this matrix \( \Delta \) should satisfy, for example, that

(i) \( \Delta(\lambda) \) is regular if (and only if) \( \lambda \in \mathbb{C} \) is in the resolvent set of \( T \) (that is, \( \lambda I - T \) is an isomorphism),

(ii) \( \lambda \) is a root of \( \det \Delta(\cdot) \) of order \( q \) if and only if \( \lambda \) is an eigenvalue of \( T \) of algebraic multiplicity \( q \) (see [8] or Lemma 4.2 in §4 how one can construct the Jordan chains of \( T \) from \( \Delta \)), and

(iii) \( \dim \ker \Delta(\lambda) \) is the geometric multiplicity of \( \lambda \) as an eigenvalue of \( T \).

An immediate consequence of the existence of such a matrix \( \Delta \) would be the upper limit \( n \) (the dimension of \( \Delta(\lambda) \)) on the geometric multiplicity of eigenvalues of \( T \).

For time-invariant linear DDEs (where \( L \) does not depend on \( t \) in (1.1)) the existence of a characteristic matrix has been known and used for a long time. The general theoretical background and construction (extended to more general evolutionary systems such as, for example, neutral equations and some classes of hyperbolic partial differential equations) is given by Kaashoek and Verduyn-Lunel in [8].

For time-periodic linear DDEs the textbook of Hale and Verduyn-Lunel [6] has developed characteristic matrices only for the case where all time delays are multiples of the period. This has been generalized to delays depending rationally on the period in order to derive analytical stability results for periodic orbits of the classical scalar delayed
positive feedback problem [12]. A characteristic function in the form of a continued fraction expansion is given for a scalar DDE in [7]. In general, precise analytical statements about the location of Floquet multipliers (or the number of unstable Floquet multipliers) can be made only for DDEs with very special structure, typically scalar DDEs with single delays and particular requirements on the right-hand side, see [1, 3, 14, 15] for classical works. An extension of these results to cyclic DDEs using a discrete Lyapunov function is given in [10].

Yanchuk and Perlikowski consider periodic orbits of nonlinear DDEs with a single fixed delay, see Appendix B. They consider periodic coefficient matrices with single delays and particular requirements on the right-hand side, see [1, 3, 14, 15] for classical works. An extension of these results to cyclic DDEs using a discrete Lyapunov function is given in [10].

Szalai et al. give a general construction for a characteristic matrix Δ for time-periodic DDEs in [13] that lends itself easily to robust numerical computation. Since we expect that Δ(λ) must have an essential singularity at λ = 0 (the eigenvalues of the compact time-1 map T accumulate at 0) the constructed matrix is rather of the form Δ(μ) where μ = λ−1. We show (in §3) that the characteristic matrix Δ(μ) as constructed by Szalai et al. typically has poles in the complex plane, making it unusable whenever eigenvalues of T of interest coincide with these poles. However, the set of these poles is discrete and accumulates only at ∞. Moreover, in §4 we provide a modification Δk(μ) of the original matrix Δ(μ), which allows us to push the poles of Δk to the outside of any given ball of radius R ≥ 1 in the complex plane by increasing the dimension of Δk. This modification reduces the eigenvalue problem for T to a root-finding problem for det Δk(μ) for all eigenvalues of T with modulus larger than 1/R, which is useful because one is typically interested in the largest eigenvalues of T.

To keep the notational overhead limited we develop our arguments for the case of a DDE with a single delayed argument and a constant delay τ < 1. In Appendix B we show that the generalization to arbitrary delays (distributed and reaching arbitrarily far into the past) is straightforward.

One immediate application for the characteristic matrix came up in [16] where Yanchuk and Perlikowski consider periodic orbits of nonlinear DDEs with a single fixed delay τ but change the delay to τ + kP (where k is an integer and P is the period of the periodic orbit), and then study the stability of the periodic orbit (which does not change its shape) in the limit k → ∞. The limit is, of course, a singular limit if one considers the time-P map T but the characteristic matrix Δk has a well-defined regular limit. This permitted the authors to draw conclusions about the linear stability properties of periodic orbits for sufficiently large delays from the properties of the so-called pseudo-continuous spectrum [16]. As [16] relies on the original construction from [13], which may have poles near the unit circle, our note closes a gap in the argument of [16].

2. Construction of characteristic matrix for a single fixed delay. Consider a periodic linear differential equation of dimension n with a single delay τ and continuous periodic coefficient matrices A(t) and B(t) (both of size n × n):

\[
\dot{x}(t) = A(t)x(t) + B(t)x(t - \tau).
\]

where we assume that the period of the time dependence of A and B equals 1 without loss of generality (this corresponds to rescaling time, τ, A and B). We also assume for simplicity of notation that τ < 1. Let us denote the monodromy operator (also called the time-1 map) for (2.1) by \( T : C([-1, 0); \mathbb{C}^n) \rightarrow C([-1, 0); \mathbb{C}^n) \). \( C([-1, 0]; \mathbb{C}^n) \) is the space of continuous functions on the interval \([-1, 0]\) with values in \( \mathbb{C}^n \). That is, \( T \) maps an initial condition \( x \in C([-1, 0]; \mathbb{C}^n) \) to the solution of (2.1) after time 1 starting from the initial value \( x(0) \) (the head point) and using the history segment \( x(\cdot) \). For \( x \in C([-1, 0]; \mathbb{C}^n) \) we define \( Tx \) precisely as the solution \( y \in C([-1, 0]; \mathbb{C}^n) \) of

\[
\dot{y}(t) = A(t)y(t) + B(t)\begin{cases} y(t - \tau) & \text{if } t \in [\tau - 1, 0] \\ x(1 + t - \tau) & \text{if } t \in [-1, \tau - 1] \end{cases}
\]

\( y(-1) = x(0) \).
The complex number $\lambda$ is an eigenvalue of $T$ (a Floquet multiplier for (2.1)) if one can find a non-zero $x$ such that $Tx = \lambda x$. We know that $T$ is compact [6] for $\tau < 1$ (and that, for arbitrary $\tau$, we can find a power $T^m$ that is compact). Hence, any non-zero $\lambda$ is either in the resolvent set of $T$ (that is, $\lambda I - T$ is regular) or it is an eigenvalue of $T$ with a finite algebraic multiplicity.

If we introduce $\mu = 1/\lambda$ we may write the eigenvalue problem for $T$ as follows:

$$\begin{align*}
x(t) &= A(t)x(t) + B(t)\begin{cases} x(t - \tau) & \text{if } t \in [\tau - 1, 0], \\
\mu x(1 + t - \tau) & \text{if } t \in [-1, \tau - 1], \end{cases} \\
x(-1) &= \mu x(0) \quad (2.3) \\
x(-1) &= \nu \quad (2.4)
\end{align*}$$

where $x \in C([-1, 0] ; \mathbb{C}^n)$ is continuous. Reference [13] constructs a characteristic matrix $\Delta(\mu)$ for $T$ based on the following hypothesis.

**Hypothesis 2.1 (Szalai'06)**

The initial-value problem on the interval $[-1, 0]$

$$\begin{align*}
\dot{x}(t) &= A(t)x(t) + B(t)\begin{cases} x(t - \tau) & \text{if } t \in [\tau - 1, 0], \\
\mu x(1 + t - \tau) & \text{if } t \in [-1, \tau - 1], \end{cases} \\
x(-1) &= \nu \quad (2.5) \\
x(-1) &= \mu x(0) \quad (2.6)
\end{align*}$$

has a unique solution $x \in C([-1, 0] ; \mathbb{C}^n)$ for all $\mu \in \mathbb{C}$ and $\nu \in \mathbb{C}^n$.

Note that the problem (2.5)–(2.6) differs from the boundary-value problem (2.3)–(2.4) because we have replaced the boundary condition $x(-1) = \mu x(0)$ by an initial condition $x(-1) = \nu$. The formulation of the hypothesis in [13] is (only apparently) weaker: it claims the existence only for $\mu \in \mathbb{C}$ for which $1/\mu$ is either in the resolvent set of $T$ or an eigenvalue of $T$ of algebraic multiplicity one. If Hypothesis 2.1 is true then one can define the characteristic matrix $\Delta(\mu)$ via

$$\Delta(\mu) \nu = \nu - \mu x(0) \quad (2.7)$$

where $x(0)$ is the value of the unique solution of (2.5)–(2.6) at the end of the time interval $[-1, 0]$.

**Integral equation for (2.5)–(2.6).** For system (2.5)–(2.6) one has to clarify what it means for $x \in C([-1, 0] ; \mathbb{C}^n)$ to be a solution. We call $x \in C([-1, 0] ; \mathbb{C}^n)$ a solution of (2.5)–(2.6) if $x$ satisfies the integral equation

$$\begin{align*}
x(t) &= x + M_1(\mu)x(t) \quad (2.8)
\end{align*}$$

where

$$\begin{align*}
[M_1(\mu)x](t) &= \int_{-1}^{t} A(s)x(s) + B(s)\begin{cases} x(s - \tau) & \text{if } s \in [\tau - 1, 0], \\
\mu x(1 + s - \tau) & \text{if } s \in [-1, \tau - 1], \end{cases} \quad (2.9)
\end{align*}$$

pointwise for all times $t$ on the interval $[-1, 0]$. The linear operator $M_1(\mu)$ maps $C([-1, 0] ; \mathbb{C}^n)$ back into $C([-1, 0] ; \mathbb{C}^n)$ and is compact. ($M_1$ will be generalized later to $M_k$ for $k > 1$.)

**Lemma 2.2**

For $\mu$ sufficiently close to 0 the initial-value problem (2.5)–(2.6) has a unique solution for all $\nu$.

**Proof:** For (2.8) to have a unique solution it is enough to show that $I - M_1(\mu)$ is invertible. Since the time-1 map $T$ is well defined and linear in $x = 0$ (that is, $Tx = 0$ for $x = 0$) we have that (2.2) has only the trivial solution for $x = 0$. This in turn implies that the corresponding integral equation

$$y = M_1(0)y \quad (2.10)$$
has only the trivial solution in $C([-1, 0]; \mathbb{C}^n)$. Thus, the kernel of $I - M_1(0)$ is trivial, and, since $I - M_1(0)$ is a compact perturbation of the identity, $I - M_1(0)$ is invertible. Moreover, the real-valued function $\mu \mapsto \|M_1(\mu) - M_1(0)\|$ (using the operator norm induced by the maximum norm of $C([-1, 0]; \mathbb{C}^n)$) is continuous in $\mu$ because $M_1$ depends continuously on $\mu$ (see definition of $M_1$ in (2.9)). This implies that $I - M_1(\mu)$ is also invertible for small $\mu$, which in turn means that (2.8)–(2.9) has a unique solution for small $\mu$.

The following lemma states that the function on the complex domain $\mu \mapsto [I - M_1(\mu)]^{-1}$, which has operators in $\mathcal{L}(C([-1, 0]; \mathbb{C}^n))$ as its values, is well defined and analytic everywhere in the complex plane except, possibly, in a discrete set of poles.

**Lemma 2.3**

The operator-valued complex function $\mu \mapsto [I - M_1(\mu)]^{-1}$ can have at most a finite number of poles in any bounded subset of $\mathbb{C}$. All poles of $[I - M_1(\mu)]^{-1}$ (if there are any) are of finite multiplicity. If $\mu$ is not a pole then $\mu \mapsto [I - M_1(\mu)]^{-1}$ is analytic in $\mu$.

**Proof.** We can split the operator $M_1(\mu)$ into a sum of two operators:

$$M_1(\mu) = M_1(0) + \mu L \quad \text{where}$$

$$[Lx](t) = \int_{-1}^{t} B(s) \begin{cases} 0 & \text{if } s \in [\tau - 1, 0), \\ x(1 + s - \tau) & \text{if } s \in [-1, \tau - 1) \end{cases} ds.\quad (2.12)$$

The operator $I - M_1(0)$ is invertible (and a compact perturbation of the identity) and $L : C([-1, 0]; \mathbb{C}^n) \mapsto C([-1, 0]; \mathbb{C}^n)$ is compact. Thus, we can rewrite $I - M_1(\mu)$ as

$$I - M_1(\mu) = (I - M_1(0)) \left[ I - \mu(I - M_1(0))^{-1}L \right]$$

and, thus,

$$[I - M_1(\mu)]^{-1} = \lambda \cdot \left[ \lambda I - (I - M_1(0))^{-1}L \right]^{-1} (I - M_1(0))^{-1}$$

(keeping in mind that $\mu = 1/\lambda$). The function $\lambda \mapsto [\lambda I - (I - M_1(0))^{-1}L]^{-1}$ on the right-hand side is the standard resolvent of the compact operator $(I - M_1(0))^{-1}$, which has only finitely many poles of finite multiplicity outside of any neighborhood of the complex origin according to [9], and is analytic everywhere else. □

**Lemma 2.3** carries over to the characteristic matrix $\Delta(\mu)$ defined in (2.7): since $x = [I - M_1(\mu)]^{-1}v$ (if we extend $v \in \mathbb{C}^n$ to a constant function), $\Delta(\mu)v = v - \mu x(0)$ is an analytic function with at most a discrete set of poles of finite multiplicity (possibly accumulating at $\infty$).

The characteristic matrix $\Delta(\mu)$ (where it is well defined) has the properties one expects: for a complex number $\mu$ in the domain of definition of $\Delta$, $1/\mu$ is a Floquet multiplier of the DDE (2.1) if and only if $\Delta(\mu)$ is singular ($\det \Delta(\mu) = 0$). Any vector $v$ in its kernel is the value of an eigenfunction for $1/\mu$ at time $t = -1$. The full eigenfunction can be obtained as the solution of the initial-value problem (2.5)–(2.6). If $\Delta(\mu)$ is regular then $1/\mu$ is in the resolvent set of the eigenvalue problem (2.3)–(2.4). The Jordan chains associated to eigenvalues $1/\mu$ can be obtained by following the general theory described in [8]. We give a precise statement (see Lemma 4.2 in Section 4) about the Jordan chain structure after discussing Hypothesis 2.1.

The construction can be extended in a straightforward manner to multiple fixed discrete delays and distributed delays. The only modification is that higher powers of $\mu$ have to be included for delays larger than 1. For example, if $\tau \in (1, 2]$ then the initial-value problem (2.5)–(2.6) is modified to

$$\dot{x}(t) = A(t)x(t) + B(t) \begin{cases} \mu x(1 + t - \tau) & \text{if } t \in [\tau - 2, 0), \\ \mu^2 x(2 + t - \tau) & \text{if } t \in [-1, \tau - 2), \end{cases}$$

$x(-1) = v$. 
Thus, also for arbitrary delays $\Delta(\mu)$ always has at worst finitely many poles of finite multiplicity in any bounded subset of $\mathbb{C}$. Appendix B discusses the case where the functional accessing the history of $x$ is a Lebesgue-Stieltjes integral.

3. The poles of the characteristic matrix. Hypothesis 2.1 is not true in its stated form (and neither in the form stated in [13]). Let us analyse the simple scalar example

$$\dot{x}(t) = x \left( t - \frac{1}{2} \right)$$

(3.1)

on the interval $[-1, 0]$. Problem (2.5)--(2.6) restated for this example is equivalent to

$$\begin{align*}
\dot{x}_1(t) &= \mu x_2(t), \quad x_1(0) = v \\
\dot{x}_2(t) &= x_1(t), \quad x_2(0) = x_1(1/2)
\end{align*}$$

(3.2)

where $x_1$ and $x_2$ are in $C([-1, 1/2]; \mathbb{C})$, and $x_1(t) = x(t - 1)$ and $x_2(t) = x(t - 1/2)$ for $t$ in the interval $[0, 1/2]$. The time-$1/2$ map for the IVP of the ODE (3.2) with initial conditions $x_1(0) = v$, $x_2(0) = w$ is

$$\begin{bmatrix}
x_1(1/2) \\
x_2(1/2)
\end{bmatrix} = 
\begin{bmatrix}
\cosh \left( \frac{1}{2} \sqrt{\mu} \right) & \sqrt{\mu} \sinh \left( \frac{1}{2} \sqrt{\mu} \right) \\
\sinh \left( \frac{1}{2} \sqrt{\mu} \right) / \sqrt{\mu} & \cosh \left( \frac{1}{2} \sqrt{\mu} \right)
\end{bmatrix} 
\begin{bmatrix}
v \\
w
\end{bmatrix}
$$

where $\sqrt{\mu}$ refers to the principal branch of the complex square root (the set of solutions is the same for both branches as all entries in the matrix are even). The continuity condition $x_2(0) = x_1(1/2)$ in (3.2) is satisfied if and only if

$$\left[ 1 - \sqrt{\mu} \sinh \left( \frac{1}{2} \sqrt{\mu} \right) \right] w + \cosh \left( \frac{1}{2} \sqrt{\mu} \right) v = 0,$$

which has a unique solution for arbitrary $v$ if and only if

$$0 \neq 1 - \sqrt{\mu} \sinh \left( \frac{1}{2} \sqrt{\mu} \right).$$

Thus, we see that for example (3.1) the initial-value problem (2.5)--(2.6) cannot be solved for non-zero $v$ whenever $\mu$ is a root of the function

$$f(\mu) = 1 - \sqrt{\mu} \sinh \left( \frac{1}{2} \sqrt{\mu} \right).$$

This function $f$ is a globally defined real analytic function of $\mu$ (the expression on the right-hand side is even in $\sqrt{\mu}$). It has an infinite number of complex roots accumulating at infinity. One of the roots is real: $f(0) = 1$, and $f(\mu) \to -\infty$ for $\mu \to +\infty$ and $\mu \in \mathbb{R}$ (let’s call it $\mu_0$; $\mu_0 \approx 1.8535$). Consequently, example (3.1) provides a counterexample to Hypothesis 2.1. If one extends example (3.1) by a decoupled equation that has its Floquet multiplier at $1/\mu_0$ (for example, $y = -[\log \mu_0] y$) then it also contradicts the hypothesis as stated in [13].

As we have discussed in Section 2, one finds that $\mu = 1/\lambda$ is a pole of $\Delta(\mu)$ if $\lambda$ is a non-zero eigenvalue of the operator $[I - M_1(0)]^{-1} L$ where $M_1$ and $L$ are defined in (2.9) and (2.12). Thus, whenever $[I - M_1(0)]^{-1} L$ has a non-zero spectral radius we must expect that $\Delta(\mu)$ has poles.

4. The extended characteristic matrix. The construction of $\Delta(\mu)$ suffers from the problem that the poles of $\Delta(\mu)$ may coincide with the inverse of Floquet multipliers of (2.1) of interest. A simple extension of the construction of $\Delta(\mu)$ permits one to push the poles to the outside of a circle of any desired radius $R$. Then the characteristic matrix can be used to find all Floquet multipliers outside of the ball of radius $1/R$. We explain
the extension in detail for a single delay $\tau < 1$ (the straightforward generalization to the general case is relegated to Appendix B).

The idea is based on the observation that the linear part of the right-hand side in the integral equation formulation (2.8)-(2.9) of initial-value problem (2.5)-(2.6) has a norm less than 1 if the interval is sufficiently short (thus, making the fixed point problem (2.8)-(2.9) uniquely solvable).

Similar to a multiple shooting approach, we partition the interval $\left[-1, 0\right]$ into $k$ intervals of size $1/k$:

$$J_i = \left[t_i, t_{i+1}\right] = \left[-1 + \frac{i}{k}, -1 + \frac{i+1}{k}\right] \text{ for } i = 0, \ldots, k - 1. \quad (4.1)$$

Then we write a system of $k$ coupled initial value problems for a vector of $k$ initial values $\left(v_0, \ldots, v_{k-1}\right)^T \in \mathbb{C}^n$:

$$\dot{x}(t) = A(t)x(t) + B(t) \begin{cases} x(t - \tau) & \text{if } t \in [\tau - 1, 0], \\ \mu x(1 + t - \tau) & \text{if } t \in [-1, \tau - 1), \end{cases} \quad (4.2)$$

$$x(t_i) = v_i \quad \text{for } i = 0, \ldots, k - 1 \quad (4.3)$$

where $t \in [-1, 0]$. System (4.2)-(4.3) corresponds to a coupled system of $k$ differential equations for the $k$ functions $x|_{J_i}$ in the time intervals $J_i$. More precisely, $x$ is defined as a solution of the fixed point problem

$$x(t) = v_s(t) + \int_{a_i(t)}^{t} A(s)x(s) + B(s) \begin{cases} x(s - \tau) & \text{if } s \in [\tau - 1, 0], \\ \mu x(1 + s - \tau) & \text{if } s \in [-1, \tau - 1) \end{cases} ds \quad (4.4)$$

$$v_s(t) = v_i \quad \text{if } t \in J_i, \quad (4.5)$$

$$a_i(t) = t_i \quad \text{if } t \in J_i. \quad (4.6)$$

We point out that a solution $x$ of (4.2)-(4.3) is not necessarily in $C([-1, 0]; \mathbb{C}^n)$ because it will typically have discontinuities at the restarting times $t_i$ for $i = 1 \ldots k - 1$.}

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**Fig. 4.1:** Illustration of solution $x \in C_k$ for multiple-shooting problem (4.2)-(4.3) with $k = 3$ sub-intervals. The right-sided limits $x(t_i)_+$ are equal to the variables $v_i$. The gaps between the left-sided limits $x(t_i)_-$ and $x(t_i)_+$ are zero if $[v_0, \ldots, v_k]$ is in the kernel of the characteristic matrix $\Delta_k(\mu)$. The equations for $x|_{J_i}$ on the sub-intervals are coupled due to the terms $t - \tau$ or $1 + t - \tau$ in (4.2).
as illustrated in Fig. 4.1. Thus, we should define the space in which we look for solutions as the space of piecewise continuous functions:

\[ C_k = \{ x : [-1, 0] \rightarrow \mathbb{C}^n : x \text{ continuous on each (half-open) subinterval } J_i \} \tag{4.7} \]

\[ C_{k,0} = \{ x \in C_k : x(t_i) = 0 \text{ for all } i = 0 \ldots k - 1 \}. \tag{4.8} \]

We call \( x \in C_k \) a solution of (4.4)–(4.6) if it satisfies (4.4) for every \( t \in [-1, 0] \). Both spaces, \( C_k \) and \( C_{k,0} \), are equipped with the usual maximum norm. For example, the piecewise constant function \( v_i \) as defined in (4.5) is an element of \( C_k \). Several operations are useful when dealing with functions in \( C_k \) to keep the notation compact. Any element \( x \in C_k \) has well-defined one-sided limits at the boundaries \( t_i \), which we denote by the subscripts + and −:

\[ x(t_i)_+ = \lim_{t \downarrow t_i} x(t) \quad \text{for } i = 1 \ldots k, \]

\[ x(t_i)_- = \lim_{t \uparrow t_i} x(t) = x(t_i) \quad \text{for } i = 0 \ldots k - 1. \]

We also define the four maps

\[ S : \mathbb{C}^{nk} \rightarrow C_k \quad S[v_0 \ldots v_{k-1}]^T(t) = v_i \text{ if } t \in [t_i, t_{i+1}) \text{ for } i = 0 \ldots k - 1, \]

\[ \Gamma_+ : C_k \rightarrow \mathbb{C}^{nk} \quad \Gamma_+[x(\cdot)] = [x(-1)_+, x(t_1)_+, \ldots, x(t_{k-1})_+]^T, \]

\[ \Gamma_-\mu : C_k \rightarrow \mathbb{C}^{nk} \quad \Gamma_-\mu[x(\cdot)] = [\mu x(0)_-, x(t_1)_-, \ldots, x(t_{k-1})_-]^T, \]

\[ M_k(\mu) : C_k \rightarrow C_{k,0} \text{ where } \]

\[ [M_k(\mu)x](t) = \int_{a_k(t)}^t A(s)x(s) + B(s) \begin{cases} x(s - \tau) & \text{if } s \in [\tau - 1, 0], \\ \mu x(1 + s - \tau) & \text{if } s \in [-1, \tau - 1] \end{cases} \, ds \tag{4.9} \]

(the piecewise constant function \( a_k \) was defined in (4.6) as \( a_k(t) = t_i \) if \( t \in [t_i, t_{i+1}) \)). The map \( S \) takes a tuple of vectors \( v_0, \ldots, v_{k-1} \) and maps it to a piecewise constant function, assigning \( x(t_i) = v_i \) and then extending with a constant to the subinterval \( J_i \). For example, the function \( v_i \) in (4.4)–(4.5) is equal to \( Sv_i \). The map \( \Gamma_+ \) takes the right-side limits of a piecewise continuous function (thus, \( \Gamma_+ S \) is the identity in \( \mathbb{C}^{nk} \)). The map \( \Gamma_-\mu \) takes the left-side limits at the interior boundaries \( t_i \) (\( i \geq 1 \)), and in its first component it takes the left-side limit at \( t_k = 0 \) (the end of the interval) multiplied by \( \mu \). The map \( M_k(\mu) \) is the generalization of \( M_1 \) defined in (2.9): in the definition of \( M_k \) the lower boundary in the integral is not −1 but \( a_k(t) \) such that we can estimate its norm:

\[ \|M_k(\mu)\|_\infty \leq \frac{1}{k} \max_{t \in [-1,0]} \|A(t)\|_\infty + |\mu| \max_{t \in [0,1]} \|B(t)\|_\infty =: \frac{C_k(|\mu|)}{k}. \tag{4.10} \]

According to the Theorem of Arzelá-Ascoli (see [4], page 266) the operator \( M_k(\mu) \) is also compact because it maps any bounded set into a set with uniformly bounded Lipschitz constants. Using the above notation we can write the differential equation (4.2)–(4.3) (or, rather, the corresponding integral equation (4.4)) as a fixed point problem in \( C_k \). For a given \( v = [v_0, \ldots, v_{k-1}] \in \mathbb{C}^{nk} \) and \( \mu \in \mathbb{C} \) we find a function \( x \in C_k \) such that

\[ x = Sv + M_k(\mu)x. \tag{4.11} \]

Due to the norm estimate (4.10) on \( M_k(\mu) \) we can state the following fact about the existence of a fixed point \( x \) of (4.11) for a given \( v \) and \( \mu \).
Lemma 4.1

Let \( R \geq 1 \) be given. If we choose \( k > C_k(R) \) then the system (4.2)–(4.3) has a unique solution \( x \in C_k \) for all initial values \((v_0 \ldots v_{k-1})^T\) and all \( \mu \) satisfying \( |\mu| < R \).

Proof.: The differential equation (4.2)–(4.3) is equivalent to fixed point problem (4.11). The norm estimate (4.10) implies that \( I - M_k(\mu) \) is invertible such that the solution \( x \) of (4.11) is given by \( (I - M_k(\mu))^{-1}Sv \).

Lemma 4.1 implies that for \( \mu \) satisfying \( C_k(|\mu| < k \) the operator \( I - M_k(\mu) \) is an isomorphism on \( C_k \).

4.1. Extended characteristic matrix. We can use Lemma 4.1 to define the extended characteristic matrix \( \Delta_k(\mu) \) for \( \mu \) satisfying \( C_k(|\mu| < k \). This is a matrix in \( \mathbb{C}^{nk \times nk} \) and is defined as

\[
\Delta_k(\mu)v = \Delta(\mu)\begin{bmatrix} v_0 \\ \vdots \\ v_{k-1} \end{bmatrix} = \begin{bmatrix} v_0 - \mu x(0) \\ v_1 - x(t_1) \\ \vdots \\ v_{k-1} - x(t_{k-1}) \end{bmatrix}
\]

(4.12)

where \( x = (I - M_k(\mu))^{-1}Sv \in C_k \) is the unique solution of the coupled system of initial-value problems, (4.2)–(4.3), with initial condition \( v = (v_0, \ldots, v_{k-1})^T \in \mathbb{C}^{nk} \). The first row in the right side of definition (4.12) corresponds to the boundary condition (2.4). The other \( k - 1 \) rows, \( v_i - x(t_i) \), guarantee for \( v \in \ker(\Delta_k(\mu)) \) that the right and the left limit of \( x \) at the restarting times \( t_i \) agree, making the gaps shown in Fig. 4.1 zero. Thus, \( x = (I - M_k(\mu))Sv \) is not only in \( C_k \) but in \( C([-1, 0]; \mathbb{C}^n) \) if \( \Delta_k(\mu)v = 0 \). What is more, if \( \Delta_k(\mu)v = 0 \) then the right-hand side of (4.2) for the solution \( x = (I - M_k(\mu))^{-1}Sv \) is continuous such that \( x \) is continuously differentiable and satisfies for every \( t \) the differential equation (2.3)–(2.4), which defines the eigenvalue problem for \( T \).

Let us list several useful facts about \( C_k, C_{k,0}, \Delta_k(\mu), T \) and the quantities defined in (4.9).

1. We can express \( \Delta_k(\mu) \) using the maps defined in (4.9):

\[
\Delta_k(\mu) = I - \Gamma_-(\mu) \left[ I - M_k(\mu) \right]^{-1} S.
\]

(4.13)

2. \( C_k \) is isomorphic to \( \mathbb{C}^{nk} \times C_{k,0} \):

\[
(v, \varphi) \in \mathbb{C}^{nk} \times C_{k,0} \quad \rightarrow \quad Sv + \varphi \in C_k \\
\psi \in C_k \quad \rightarrow \quad (\Gamma_+\psi, \psi - S\Gamma_+\psi) \in \mathbb{C}^{nk} \times C_{k,0}.
\]

(4.14)

3. The differential equation

\[
\begin{align*}
\dot{x}(t) &= A(t)x + B(t) \begin{cases} x(t - \tau) & \text{if } t \in [\tau - 1, 0], \\
0 & \text{if } t \in [-1, \tau - 1], \end{cases} \\
x(-1)_+ &= 0 \\
x(t_i)_+ &= x(t_i)_- & \text{for } i = 1, \ldots, k - 1
\end{align*}
\]

(4.15)

has only the trivial solution (it is identical to (2.5)–(2.6) for \( \mu = 0 \)). Its integral formulation is \( x = \Gamma_-(0)x + M_k(0)x \). Thus, the operator \( I - \Gamma_-(0) - M_k(0) : C_k \rightarrow C([-1, 0]; \mathbb{C}^n) \) (which is of Fredholm index \( 0 \)) is invertible.

4. The time-1 map \( T \) of the linear DDE \( \dot{x}(t) = A(t)x(t) + B(t)x(t - \tau) \), defined in (2.2) for initial history segments in \( x \in C([-1, 0]; \mathbb{C}^n) \), can be easily extended to initial history segments in \( C_k \). In fact, the extension of \( T \) to \( C_k \) can be defined by

\[
T = \left[ I - \Gamma_-(0) - M_k(0) \right]^{-1} \left[ S(\Gamma_-(1) - \Gamma_-(0)) + M_k(1) - M_k(0) \right],
\]

where \( x = (I - M_k(\mu))Sv \).
which maps $C_k$ into $C([-1,0];\mathbb{C}^n)$, such that

$$I - \mu T = \left[ I - S \Gamma_0(0) - M_1(0) \right]^{-1} \left[ I - S \Gamma_1(\mu) - M_k(\mu) \right],$$

which maps $C_k$ into $C_k$ (see Appendix A for a detailed decomposition of expression (4.16)). Since the extension of $T$ maps into $C([-1,0];\mathbb{C}^n)$ (the original domain of the non-extended $T$) its spectrum is identical to the spectrum of the original $T$.

Lemma 4.1 and the above facts permit us to apply the general theory developed in [8], relating the extended characteristic matrix $\Delta_k(\mu) \in \mathbb{C}^{nk \times nk}$ to spectral properties of the time-1 map $T$; eigenvalues, eigenvectors, and the length of their Jordan chains.

**Lemma 4.2 (Jordan chains)**

Assume that $0 < |\mu_s| < R$ and $k > C_s(R)$ where $C_s(R)$ is defined by (4.10). If $\Delta_k(\mu_s)$ is regular then $1/\mu_s$ is in the resolvent set of the monodromy operator $T$. If $\Delta_k(\mu_s)$ is not regular then $\lambda_s = 1/\mu_s$ is an eigenvalue of $T$. The Jordan chain structure for $\lambda_s$ as an eigenvalue of $T$ is also determined by $\Delta_k$:

1. the dimension $l_0$ of ker $\Delta_k(\mu_s)$ is the geometric multiplicity of $\lambda_s$,
2. the order $p$ of the pole $\mu_s$ of $\Delta_k(\mu)^{-1}$ is the length of the longest Jordan chain associated to $\lambda_s$,
3. the order of $\mu_s$ as a root of det $\Delta_k(\mu)$ is the algebraic multiplicity of $\lambda_s$,
4. the Jordan chains of length less than or equal to $p$ can be found as solutions $(y_0, \ldots, y_{p-1})$ (where $y_{p-1}$ is non-zero) of the linear system

$$0 = \Delta_k^{(0)} y_0$$
$$0 = \Delta_k^{(1)} y_0 + \Delta_k^{(0)} y_1$$
$$0 = \frac{\Delta_k^{(2)}}{2} y_0 + \Delta_k^{(1)} y_1 + \Delta_k^{(0)} y_2$$
$$\vdots$$
$$0 = \frac{\Delta_k^{(p-1)}}{(p-1)!} y_0 + \ldots + \Delta_k^{(1)} y_{p-2} + \Delta_k^{(0)} y_{p-1}$$

(4.17)

where $\Delta_k^{(j)}$ is the $j$th derivative of $\Delta_k(\mu)$ in $\mu_s$ and $\Delta_k^{(0)} = \Delta_k(\mu_s)$.

System (4.17) is equivalent to the requirement

$$\Delta_k(\mu) \left[ y_0 + (\mu - \mu_s)y_1 + \ldots + (\mu - \mu_s)^{p-1}y_{p-1} \right] = O((\mu - \mu_s)^p)$$

for all $\mu \approx \mu_s$ (this is, in fact, the general definition of a Jordan chain for a nonlinear eigenvalue problem). The vectors $y_j$ in the Jordan chains found as solutions of (4.17) are still vectors in $\mathbb{C}^{nk}$. We give the precise relation between the vectors $y_j$ and the generalized eigenvectors of $I - \mu T$ in Equation (4.21) in the proof below.

**Proof.** The proof extends and modifies the construction used by [13], and makes the general theory of [8] applicable to time-periodic delay equations.

First, we give a brief summary of the relevant statement of [8]. Let $G : \Omega \rightarrow \mathcal{L}(X_1; Y_1)$ and $H : \Omega \rightarrow \mathcal{L}(X_2; Y_2)$ be two functions which are holomorphic on the open subset $\Omega$ of the complex plane $\mathbb{C}$, and map into the space of bounded linear operators from one Banach space $X_{1,2}$ into another Banach space $Y_{1,2}$. The functions $G$ and $H$ are called equivalent if there exist two functions $E : \Omega \rightarrow \mathcal{L}(X_2; X_1)$ and $F : \Omega \rightarrow \mathcal{L}(Y_1; Y_2)$ (also holomorphic on $\Omega$) whose values are isomorphisms such that

$$H(\mu) = F(\mu)G(\mu)E(\mu)$$

for all $\mu \in \Omega$. 

If two operators \( G \) and \( H \) are equivalent then every Jordan chain \( (u_0, \ldots, u_{p-1}) \) of the eigenvalue problem \( H(\mu)u = 0 \) for \( \mu \), corresponds to exactly one Jordan chain \( (v_0, \ldots, v_{p-1}) \) of the eigenvalue problem \( G(\mu)v = 0 \) at \( \mu_* \) via the isomorphism

\[
\begin{align*}
    v_0 &= E_0 u_0 \\
    v_1 &= E_0 u_1 + E_1 u_0 \\
    &\vdots \\
    v_{p-1} &= E_0 u_{p-1} + \ldots + E_{p-1} u_0,
\end{align*}
\]

where the operators \( E_i \) are the expansion coefficients of the isomorphism \( E \) in \( \mu = \mu_* \):

\[
E(\mu) = E_0 + (\mu - \mu_1) E_1 + \ldots + (\mu - \mu_p)^{p-1} E_{p-1} + O((\mu - \mu_p)^p).
\]

We construct the equivalence initially only for the case of a single delay \( \tau < 1 \). The necessary modifications for arbitrary delays can be found in Appendix B (the underlying idea is identical but there is more notational overhead). Also, we restrict ourselves here to merely stating what the operators \( H, F, G \) and \( E \) are, relegating the detailed checks and calculations to Appendix A.

The subset of permissible \( \mu, \Omega \subset \mathbb{C} \), is the ball of radius \( R \) around \( 0 \), where the initial-value problem (4.2)–(4.3) has a unique solution. In our equivalence the operator functions \( G \) and \( H \) are

\[
G(\mu) = I - \mu T : \quad X_1 = C_k \quad \mapsto \quad Y_1 = X_1,
\]

\[
H(\mu) = \begin{pmatrix} \Delta_0(\mu) & 0 \\ 0 & I \end{pmatrix} : \quad X_2 = \mathbb{C}^{nk} \times C_{k,0} \quad \mapsto \quad Y_2 = X_2.
\]

Note that the spaces \( X_1 \) and \( X_2 \) are isomorphic via relation (4.14). The eigenvalue problem \( G(\mu)v = 0 \) is the eigenvalue problem for \( T \) (reformulated for \( \mu = \lambda^{-1} \)). The eigenvalue problem \( H(\mu)[v, \varphi] = 0 \) is equivalent to \( \Delta_0(\mu)v = 0, \varphi = 0 \). Thus, establishing equivalence between \( G \) and \( H \) proves Lemma 4.2. We define the isomorphisms \( E \) and \( F \) as

\[
E(\mu)[v, \varphi] = [I - M_k(\mu)]^{-1} [Sv + \varphi]
\]

\[
F(\mu)\psi = \begin{bmatrix} F_1(\mu)\psi \\ F_2(\mu)\psi \end{bmatrix} = \begin{bmatrix} [\Gamma_+ - \Gamma_-(0)] \psi + \Gamma_-(\mu) [I - M_k(\mu)]^{-1} [I - ST_+ - M_k(0)] \psi \\ [I - ST_+ - M_k(0)] \psi \end{bmatrix}
\]

(4.20)

See (4.13) and (4.16) for expressions giving \( \Delta_k(\mu) \) and \( I - \mu T \), defining \( G \) and \( H \) in terms of \( S, \Gamma_\pm \) and \( M_k \). The inverse of \( I - M_k(\mu) \) is guaranteed to exist by Lemma 4.1. The only choice we make is the definition of \( E \), which is clearly an isomorphism from \( X_2 \) to \( X_1 \) (because \( (v, \varphi) \rightarrow Sv + \varphi \) is an isomorphism from \( X_2 \) to \( X_1 \)). The definition of \( F \) is then uniquely determined if we want to achieve the equivalence \( F(\mu)G(\mu)E(\mu) = H(\mu) \), and follows from a straightforward but technical calculation, given in Appendix A. The concrete form of \( E \) and \( F \) allows us to specify the precise relation between Jordan chains of \( \Delta_k(\mu)y = 0 \) and \( (I - \mu T)x = 0 \): \((x_0, \ldots, x_{p-1}) \in C([-1, 0]; \mathbb{C}^n)^p \) is a Jordan chain of \( (I - \mu T)x = 0 \) in \( \mu = \mu_* \) if and only if

\[
\begin{align*}
    x_0 &= E_0[y_0, 0], & y_0 &= \Gamma_+ x_0 \\
    x_1 &= E_0[y_1, 0] + E_1[y_0, 0], & y_1 &= \Gamma_+ x_1 - \Gamma_+ E_1[y_0, 0] \\
    &\vdots & &\vdots \\
    x_{p-1} &= E_0[y_{p-1}, 0] + \ldots + E_{p-1}[y_0, 0], & y_{p-1} &= \Gamma_+ x_{p-1} - \Gamma_+ E_1[y_{p-2}, 0] - \cdots - \Gamma_+ E_{p-1}[y_0, 0].
\end{align*}
\]
which can be rearranged for $\tilde{y}$ because the factor $I - S\Gamma_-(0) - M_k(0)$ (which is in front of $y$) is invertible as explained below (4.15) in the list of useful facts in §4:

$$\tilde{y} = [I - S\Gamma_-(0) - M_k(0)]^{-1} [S\Gamma_-(\mu) - S\Gamma_-(0) + M_k(\mu) - M_k(0)] x.$$  \hspace{1cm} (A.6)
We notice that the operator in front of $x$ is also defined for all $x \in C$, which allows us to extend the definition of $x \rightarrow y = \mu T x$ to $x \in C$. Moreover, when applying $\Gamma_+$ to (A.5) we observe that $\hat{y}$ is continuous in $[-1, 0]$, that is, the map $x \rightarrow \mu T x$ maps $C$ into $C([-1, 0]; \mathbb{C}^n)$. Inserting $\hat{y} = x - y$ into (A.6) gives expression (4.16), which is equivalent to the fixed point problem
\[
y = S\Gamma_-(0)y + M_k(0)y + x - S\Gamma_-(\mu)x - M_k(\mu)x.
\] (A.7)

**Determination of $F(\mu)$**. From the fixed point equation (A.7) it becomes clear how to choose the other isomorphism $F(\mu)$ such that $F(\mu)G(\mu)E(\mu)[v, \varphi] = H(\mu)[v, \varphi]$. We observe that (A.7) implies (by applying $\Gamma_+$ to both sides)
\[
\Gamma_+ y = \Gamma_-(0)y + \Gamma_+ x - \Gamma_-(\mu)x.
\] (A.8)

We apply $S$ to this identity and subtract it from the fixed point equation (A.7) defining $y$:
\[
y - S\Gamma_+ y = M_k(0)y + x - S\Gamma_+ x - M_k(\mu)x.
\] If $x = E(\mu)[v, \varphi] = [I - M_k(\mu)]^{-1}[Sv + \varphi]$ then (A.2) implies $\Gamma_+ x = v$, such that we obtain for $y = G(\mu)x$
\[
y - S\Gamma_+ y - M_k(0)y = \varphi.
\] (A.9)

Thus, $[I - S\Gamma_+ - M_k(0)]G(\mu)E(\mu)[v, \varphi] = \varphi$ for all $(v, \varphi) \in X_1$, which justifies our choice of the second component of $F(\mu)$ in (4.20). Inserting the expression $E(\mu)[v, \varphi] = [I - M_k(\mu)]^{-1}[Sv + \varphi]$ for $x$ and $\Gamma_+ x = v$ into (A.8) we obtain that
\[
\Gamma_+ y - \Gamma_-(0)y = v - \Gamma_-(\mu)[I - M_k(\mu)]^{-1}[Sv + \varphi] = [I - \Gamma_-(\mu)[I - M_k(\mu)]^{-1}S]v - \Gamma_-(\mu)[I - M_k(\mu)]^{-1}\varphi = \Delta_k(\mu)v - \Gamma_-(\mu)[I - M_k(\mu)]^{-1}\varphi,
\] (A.10)

where we used the definition (4.13) of the characteristic matrix $\Delta_k(\mu)$. Since we know that $\varphi$ can be recovered from $y$ via (A.9) we obtain from (A.11) the first component of the definition of $F(\mu)$ in (4.20):
\[
\Gamma_+ y - \Gamma_-(0)y + \Gamma_-(\mu)[I - M_k(\mu)]^{-1} [I - S\Gamma_+ - M_k(0)] y = \Delta_k(\mu)v.
\]

**Inverse of $F(\mu)$**. It remains to be checked that $F$ is invertible if $|\mu| < R$. Given $v \in C^{nk}$ and $\varphi \in C_{k,0}$ such that
\[
v = \Gamma_+ y - \Gamma_-(0)y + \Gamma_-(\mu)[I - M_k(\mu)]^{-1}\varphi,\quad (A.12)
\]
\[
\varphi = y - S\Gamma_+ y - M_k(0)y,\quad (A.13)
\]

how do we recover $y$? Applying $S$ to (A.12) and adding the result to (A.13) gives
\[
Sv + \varphi = S\Gamma_-(\mu)[I - M_k(\mu)]^{-1}\varphi + [I - S\Gamma_-(0) - M_k(0)] y,\quad (A.14)
\]
which is equivalent to (A.12)–(A.13) because the map $[v, \varphi] \in C^{nk} \times C_{k,0} \rightarrow Sv + \varphi \in C_k$ is an isomorphism. Equation (A.14) can be rearranged for $y$ because $I - S\Gamma_-(0) - M_k(0)$ is invertible. Hence, $[v, \varphi] \in C^{nk} \times C_{k,0}$ and $y \in C_k$ satisfy $[v, \varphi] = F(\mu)y$ (that is, system (A.12)–(A.13)) if and only if
\[
y = [I - S\Gamma_-(0) - M_k(0)]^{-1} [Sv + \varphi - S\Gamma_-(\mu)[I - M_k(\mu)]^{-1}\varphi].
\]
Consequently,
\[
F(\mu)^{-1}[v, \varphi] = [I - S\Gamma_-(0) - M_k(0)]^{-1} [Sv + \varphi - S\Gamma_-(\mu)[I - M_k(\mu)]^{-1}\varphi].
\] (A.15)
Appendix B. Characteristic matrices for general periodic delay equations. Consider a linear DDE with coefficients of time period 1 and with maximal delay less than or equal to an integer \( m \geq 1 \):

\[
\dot{x}(t) = \int_{0}^{m} d_{\theta} \eta(t, \theta) x(t - \theta) \tag{B.1}
\]

where \( \eta(t, \theta) \) is bounded, measurable and periodic in its first argument \( t \) (with period 1), and of bounded variation in its second argument \( \theta \). We choose \( \eta(t, \cdot) \in NBV(\mathbb{R}; \mathbb{C}^{n \times n}) \), that is, \( \eta(t, \cdot) = 0 \) on \((-\infty, 0] \), \( \eta(t, \cdot) \) is continuous from the right on \((0, m] \), and \( \eta(t, \cdot) \) is constant on \([m, \infty) \). In addition we assume that the total variations of all \( \eta(t, \cdot) \) have a common upper bound:

\[
V_{0}^{m} \eta(t, \cdot) \leq \bar{V} \tag{B.2}
\]

where \( V_{0}^{m} f \) is the total variation of \( f \) [2]. In formulation (B.1) the dependent variable is the function \( x \) on the history interval \([-m, 0] \). The time-1 map \( Tx \) of (B.1) for \( x \in C([-m, 0]; \mathbb{C}^{n}) \) is defined as the solution \( y \in C([-m, 0]; \mathbb{C}^{n}) \) of the equation

\[
\begin{align*}
& t > -1 : \quad \dot{y}(t) = \int_{0}^{m} d_{\theta} \eta(t, \theta) \begin{cases} 
y(t - \theta) & \text{if } t - \theta \geq -1 \\
x(1 + t - \theta) & \text{if } t - \theta < -1
\end{cases} \\
& t \leq -1 : \quad \lim_{t \downarrow -1} y(t) = x(1 + t).
\end{align*} \tag{B.3}
\]

Only the part of \( y \) on the interval \((-1, 0] \) is the result of a differential equation (namely (B.3)). The remainder of \( y, y|_{[-m, -1]} \), is merely a shift of \( x \), given by (B.4). Note that we have included the initial condition for (B.3) into (B.4), namely \( y(-1) = x(0) \). Correspondingly, the eigenvalue problem for \( T \) reads

\[
\begin{align*}
& t > -1 : \quad \dot{x}(t) = \int_{0}^{m} d_{\theta} \eta(t, \theta) \begin{cases} 
x(t - \theta) & \text{if } t - \theta \geq -1 \\
\mu x(1 + t - \theta) & \text{if } t - \theta < -1
\end{cases} \\
& t \leq -1 : \quad \lim_{t \downarrow -1} x(t) = \mu x(1 + t).
\end{align*} \tag{B.5}
\]

The spaces \( C_{k} \) and \( C_{k,0} \), and the maps \( S, \Gamma_{\pm} \) and \( M_{\mu}(\mu) \) defined in (4.7), (4.8) and (4.9) for a single delay can be extended in a straightforward way. We partition the interval \([-m, 0] \) into sub-intervals of length \( 1/k \), \( J_{i} = [t_{i}, t_{i+1}) = [-1 + i/k, -1 + (i + 1)/k) \) \((i = k - mk \ldots k - 1) \); note that \( i \) can become negative), and define the spaces

\[
C_{k} = \{ x : [-m, 0] \to \mathbb{C}^{n} : x \text{ continuous on all (half-open) subintervals } J_{i}, \text{ and } \lim_{t \searrow t_{i}} x(t) \text{ exists for all } i = k - mk + 1 \ldots k \} \tag{B.7}
\]

\[
C_{k,0} = \{ x \in C_{k} : x(t_{i}) = 0 \text{ for all } i = 0 \ldots k - 1 \}.
\]

Note that non-negative indices \( i \) of \( t_{i} \) correspond to times \( t_{i} \) in the interval \([-1, 0] \) whereas negative indices \( i \) correspond to times \( t_{i} \) in the interval \([-m, -1] \). These points of discontinuity are treated differently in the definition of \( C_{k,0} \): to be an element of \( C_{k,0} \) the function \( x \) is only required to be zero at \( t_{i} \) with non-negative \( i \). The space \( C_{k} \) consists of piecewise continuous functions where the points of discontinuity are the times \( t_{i} \in [-m, 0] \). For functions in \( C_{k} \) we define the maps (again analogous to the
Lemma B.1 (Unique solution for initial-value problem)

Let \( R > 1 \) be fixed. Then the fixed point problem (B.12) has a unique solution \( x \) for all \( v \in \mathbb{C}^n \) and for all complex \( \mu \) satisfying \( |\mu| < R \) if the number \( k \) of the sub-intervals is chosen such that

\[
R^{1/k} < 1 + \tilde{V}^{-1} \frac{R - |\mu|}{R^{m+1}} \log R.
\]

Note that \( \tilde{V} \) is an upper bound on the total variation of \( \eta \). The left-hand side of (B.13) approaches 1 for \( k \to \infty \) whereas the right-hand side is larger than 1 if \( R > 1 \) and \( R > |\mu| \).

Proof. We define the norm

\[
\|x\|_R = \max_{t \in [-m,0]} |x'(t)|
\]
on the space \( L^\infty([-m, 0]; \mathbb{C}^n) \) of bounded functions and the spaces \( C_k \) and \( C_{k,0} \) (which are subspaces of \( L^\infty \)). This norm is equivalent to the standard maximum norm. Thus, (B.12) has a unique solution in \( C_k \) whenever \( \|M_k(\mu)\|_R < 1 \). Two operators that appear in \( M_k(\mu) \) are shift and integration. Their respective domains and norms are:

\[
S(\theta) : \mathbb{L}^\infty([-m, 0]; \mathbb{C}^n) \rightarrow \mathbb{L}^\infty([-m, 0]; \mathbb{C}^n) \quad \text{defined by}
\]

\[
[S(\theta)x](t) = \begin{cases} 
   x(t - \theta) & \text{if } t - \theta \in [-m, 0], \\
   0 & \text{otherwise},
\end{cases}
\]

\[
J : \mathbb{L}^\infty([-m, 0]; \mathbb{C}^n) \rightarrow C_{k,0} \quad \text{defined by}
\]

\[
[Jx](t) = \int_{a_k(t)}^{t} x(s)ds,
\]

\[
\|Jx\|_R \leq \frac{R^{1/k} - 1}{\log R} \|x\|_R.
\]

In order to find the \( \|\cdot\|_R \)-norm of the integrand (which is in \( \mathbb{L}^\infty([-m, 0]; \mathbb{C}^n) \)) if we extend it with zero for \( t \in [-m, -1) \) in the definition of \( M_k(\mu) \), (B.8), we have to estimate the \( \|\cdot\|_R \)-norm for a Stieltjes sum over an arbitrary partition \( 0 \leq \theta_0 < \ldots < \theta_N \leq m \) (with arbitrary intermediate values \( \tilde{\theta} \in [\theta_j, \theta_{j+1}] \)):

\[
\left\| \sum_{j=0}^{N-1} \eta(t, \theta_{j+1}) - \eta(t, \theta_j) \right\|_R \leq \sup_{t \in [-1, 0]} R^j \sum_{j=0}^{N-1} \left| \eta(t, \theta_{j+1}) - \eta(t, \theta_j) \right| R^{j} \|S(\tilde{\theta}_j)x\|_R \|S(\tilde{\theta}_j - 1)x\|_R
\]

\[
\leq \|R^m \|x\|_R \sup_{t \in [-1, 0]} V^m \eta(t, \cdot) = R^m \|x\|_R \tilde{V}.
\]

Combined with the norm estimate for integration from \( a_k(t) \) to \( t \) we obtain that

\[
\| [M_k(\mu)x] |_{[-1, 0]} \|_R \leq \|J\|_R R^m \tilde{V} \|x\|_R \leq R^m \tilde{V} \frac{R^{1/k} - 1}{\log R} \|x\|_R.
\]

The part of \( M_k(\mu)x \) on the interval \([-m, -1)\) has the \( \|\cdot\|_R \)-norm:

\[
\| [M_k(\mu)x] |_{[-m, -1]} \|_R \leq |\mu| \|S(-1)x\|_R \leq |\mu| R^{-1} \|x\|_R,
\]

where we note that \( S(-1) \) maps \( C_k \) back into \( C_k \). Adding up the inequalities (B.14) and (B.15) we obtain an upper bound for \( \|M_k(\mu)x\|_R \):

\[
\|M_k(\mu)x\|_R \leq \left[ R^m \tilde{V} \frac{R^{1/k} - 1}{\log R} + \frac{|\mu|}{R} \right] \|x\|_R.
\]

The factor in front of \( \|x\|_R \) on the right-hand side is less than unity if \( k \) satisfies the inequality (B.13) required in the statement of the lemma. \( \square \)

The invertibility of \( t - M_k(\mu) \) permits us to choose exactly the same constructions for the characteristic matrix \( \Delta_k(\mu) \) and for the isomorphisms \( E \) and \( F \) proving Lemma 4.2.
The existence of this inverse follows from the fact that the time-1 map, \( T \), makes the infinite-dimensional eigenvalue problem of the time-1 map, \([I - \mu T]x = 0\), equivalent to the finite-dimensional eigenvalue problem \( \Delta_\mu(\mu) = 0 \).

We remark that the inverse of \( I - \sigma(0) - M_k(0) \), which is present in the definition of \( G(\mu) \), exists because the initial-value problem (B.3)–(B.4) has only the trivial solution for \( x = 0 \) (the operator is a compact perturbation of the identity). In other words, the existence of this inverse follows from the fact that the time-1 map \( Tx \) is well defined (and equal to zero) in \( x = 0 \). Moreover, the extension of \( T \) to \( C_k \), defined by

\[
T = \left[ I - \sigma(0) - M_k(0) \right]^{-1} \left[ \sigma(1) - \sigma(0) \right] + M_k(1) - M_k(0) ,
\]

which maps \( C_k \) back into \( C_k \), has the property that \( T^m x \) is continuous (that is, \( T^m x \in C([-m,0];C^n) \)) for all \( x \in C_k \). Hence, the spectrum of the extension \( T : C_k \rightarrow C_k \) is the same as the spectrum of the original \( T : C([-m,0];C^n) \rightarrow C([-m,0];C^n) \).

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