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# Over-order Multiplicities and their Application in Controlling Delay Dynamics. On Zeros' Distribution of Linear Combinations of Kummer Hypergeometric Functions 

Islam Boussaada* ${ }^{* \dagger}$<br>Guilherme Mazanti*<br>Silviu-Iulian Niculescu*


#### Abstract

A series of recent works have shown that, for a system of linear functional differential equations, a spectral value having a multiplicity exceeding the order of the system tends to correspond to the spectral abscissa of the system, a property called MID for multiplicity-induced-dominancy. In particular, when this multiplicity coincides with the degree of the quasipolynomial this property is called generic MID (GMID). The GMID has been fully characterized for single-delay retarded as well as neutral delay-differential equations thanks to the representation of the corresponding quasipolynomial in terms of a Kummer hypergeometric function. However, apart from partial results, in full generality, no result of the literature enables the characterization of the dominance of a spectral value having an intermediate multiplicity, which is essentially due to the lack of existing results among the open literature pertaining to linear combinations of Kummer functions' zeros distribution. In this work, we overcome this difficulty and we further investigate the MID to cover the socalled over-order MID, that is, the cases where the multiplicity is larger than the order of the corresponding differential equation.


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Keywords. Confluent hypergeometric functions, Whittaker function, Kummer function, zeros location, stability and stabilization, exponential decay, pole placement, prescribed stabilization

## 1 Introduction

The common feature of propagation, transport phenomena/processes and population dynamics is their time heterogeneity, and there exists a wide variety of mathematical models describing these dynamic behaviors in biology, physics, economics and engineering. One of the simplest ways of capturing the said time heterogeneity of the dynamics is to use time-delay systems represented by delay-differential equations (DDEs) under appropriate initial conditions. For the basic theory of DDEs ${ }^{1}$, we refer to [11, 25, 28, 30-32, 36, 37, 39, 41, 43, 44, 70, 74]. The classification of the DDEs ${ }^{2}$ and fundamental properties of the solutions can be found in [39]. Although delay systems are infinite-dimensional, a first idea to properly understand and analyze their dynamics was to extend methods and techniques from ordinary differential equations (ODEs) to DDEs, leading to some unitary viewpoint of the qualitative properties for both classes of differential equations. Such an angle was adopted since the 1960s, and we refer, in particular, to $[31,38]$ for further insights on the underlying methods.

In this spirit, and in the same fashion as in [31] for the presentation the properties of the solutions of DDEs, we adopt some terminology coming from ODEs. In particular, the order of a DDE simply means the order

[^0]of the highest derivative involved in the equation ${ }^{3}$. In the linear DDE case, the characteristic function is a quasipolynomial (see, e.g., [11]) and, likewise the ODE case, we can introduce the notion of degree. However, since a quasipolynomial has an infinite number of (characteristic) roots, the meaning of its degree is slightly different. Indeed, contrarily to the polynomial case, the degree does not determine the number of roots, it is rather related to the number of parameters of the quasipolynomial ${ }^{4}$. By exploiting the Pólya-Szegó bound [64] pertaining to the number of roots of exponential polynomials in horizontal strips, [23] (see also [50]) showed that the maximal admissible multiplicity of a characteristic root is given by the degree of the corresponding quasipolynomial, and, in addition to the first and second-order DDE cases discussed in [63], we can mention a classical control example - stabilizing a second-order inverted pendulum by a delayed position feedback -, where one may have characteristic roots at the origin with a multiplicity larger than the order of closed-loop system $^{5}$ (see, for instance, $[22,65,68]$ and the references therein). Such multiple zeros are called over-order multiple (characteristic) roots, and to the best of the authors' knowledge, excepting some simple cases, there does not exist a systematic analysis of such (multiple) roots, and, in particular, a better comprehension of the way they may affect the dynamics of the corresponding dynamical system is missing.

One of the objectives of this paper is to further investigate the over-order multiplicity and, in particular, to explicitly determine conditions when such a root is dominant, in the sense that it defines the spectral abscissa of the corresponding dynamical system. The latter property is called multiplicity-induced-dominancy (MID). Although the existence of multiple characteristic roots for the scalar DDE was emphasized since the end of the 1940s, the first systematic study in the parameter-space for the scalar and second-order DDEs (both retarded and neutral cases) including a single delay (delay equal to one) can be found in [63]. It should be mentioned that Pinney observed that the characteristic roots with the highest (possible) multiplicity can also exhibit the greatest real part ${ }^{6}$, which, in our terminology, corresponds to the MID property. Namely, in order to avoid oscillations with the ambition of getting the maximum damping of all transient solutions, the author emphasized the difficulty of the problem and studied the triple (dominant) roots in the scalar neutral and quintuple (dominant) in the second-order neutral DDE cases, respectively. It should be pointed out that the proposed approach, which is quite close to the so-called D-decomposition method introduced by Neimark [58], cannot be simply extended to more general classes of DDEs. For a better understanding of the complexity of the analysis, we refer to the study of the over-order MID property for scalar neutral DDEs with a single delay in [12], where the authors proved the dominance by exploiting the explicit (frequency) bound for the imaginary part of the unstable (characteristic) roots ${ }^{7}$.

For DDEs of arbitrary order, the MID property was proven in the generic case (i.e., the over-order multiplicity is equal to the degree of the quasipolynomial) in [50] (retarded case) and in [15] (unified treatment of retarded and neutral cases). In fact, the authors of these papers have shown that the generic MID problem can be reduced to (i) solving an appropriate linear system of equations, and (ii) an appropriate factorization of the corresponding quasipolynomial. Finally, the dominance is proved by exploiting the properties of Kummer hypergeometric functions. For a good introduction to hypergeometric functions, we refer to [26, 34, 75] (see also [42] for an historical perspective). As underscored in [16], it should be mentioned that new interpretations of the Padé approximation of the exponential function based on the location of the zeros of Kummer functions with real parameters have been enabled [17].

For an overview of some of the methods for the characterization of the behavior of the (characteristic) multiple roots of linear DDEs in the parameter-space ${ }^{8}$, we refer to [60]. In particular, if the frequency-sweeping curves introduced in [46] capture the multiplicity of the roots located on the imaginary axis, multiplicity induced by the structure of the delay parameters, but, unfortunately, they cannot detect over-order multiplicities in the

[^1]single delay case. Next, the perturbation-based approach using the Weirstrass-Malgrange preparation theorem and proposed in [49] permits the characterization of the asymptotic behavior of the (characteristic) roots of any multiplicity, albeit without any guarantee about the dominancy of the (corresponding) root. It is worth mentioning that this last approach combined with the MID yields an appropriate local analysis of the behavior of the multiple roots with respect to some of the parameters of the corresponding system ${ }^{9}$. For further discussions on such topics, we refer to [52].

The interest for a deeper understanding and an explicit characterization of the MID property for more general classes of dynamical systems represented by DDEs emanates from the control area. As discussed in [59,69], starting with the 1950s, in engineering applications, delay was commonly associated to instability, oscillations and bad behaviors in dynamical systems, and there is an abundant literature on the elaboration of criteria allowing to guarantee the exponential stability of the systems independently of the size of the delay. However, some results showed that the delay, as a control parameter can be useful for improving the stability and/or the behavior of linear dynamical systems, see, for instance, the so-called proportional minus delay controller introduced by [71,72] or the output delay feedback that stabilizes some oscillatory systems discussed in [1]. Furthermore, as highlighted by $[45]^{10}$, increasing the delay is not necessarily associated with instability and there are cases where a larger delay may induce stability even though the property does not necessarily hold for "small" delay values.

As a consequence of the remarks/examples above, understanding the dependence of the characteristic roots with respect to the system parameters appears to be essential. Several methods/techniques grouped together under the title eigenvalue-based approaches have been the subject of several contributions, see, for instance, [54], and the references therein. These methods have the particularity of using and exploiting the duality between two types of methods for solving eigenvalue problems: one nonlinear in finite dimension and the other linear in infinite dimension and, as a consequence, to better characterize the properties of the spectral abscissa with respect to system parameters ${ }^{11}$. The derived results reinforce the interest in using the delay as a control parameter, however they also show some limitations of the approach.

Furthermore, the continuous dependence of the characteristic roots on the controller parameters represents an interesting property to be further exploited. Indeed, for instance, the so-called continuous pole placement proposed in [53] in the retarded case uses such a continuity property and explicitly uses the fact that, in the retarded case, the number of unstable roots is finite together with an appropriate monitoring of the characteristic roots with a "large" real part (see, e.g., [55] for an extension to neutral systems). This pole placement method represents one of the potential extensions of the stabilization procedure related to the classical pole placement method for ODEs [2]. Another extension of the pole placement method is the so-called partial pole placement method which simply employs the degree of the quasipolynomial and the MID property mentioned above. First, the degree of the quasipolynomial gives an upper bound on the number of functional equations to be taken into account in the appropriate tuning of the controller gains. Second, the MID property gives a guarantee for the decay rate of the solutions of the closed-loop system. This control method was discussed in [50] (retarded case) and [15] (retarded and neutral cases) in the case of maximal admissible multiplicity. A few recent results showed that this property holds for other over-order multiplicities and, in particular, in the case of the lowest over-order multiplicity, see, for instance, [9] (in the case of real-rooted plants), [18] (by exploiting the frequency bound). To the best of the authors' knowledge, there does not exist an explicit characterization of the partial pole placement in the case of over-order of multiplicity. The method proposed in this paper enables the handling of such a problem in its generality.

To summarize, the aim of this paper is to address such MID spectral problems and to propose a new method that could also encompass over-order (algebraic) multiplicities with the guarantee of dominance of the corresponding spectrum. More precisely, the contribution of the paper is threefold: first, to further investigate over-order MID cases. Roughly speaking, if the problem of multiplicity can be reduced to solving an appropriate system of functional equations, the dominance requires a deeper analysis and an appropriate understanding of the way the changes in the parameters may affect the spectrum distribution. As shown below, a central role is played by the so-called elimination-produced function (see Section 3), and its interest clearly appears

[^2]in a control setting in terms of degrees of freedom induced by the system structure and the (controller) gains selection. Next, the appropriate tool for proving the dominancy is the Green-Hille (integral) transformation introduced by Hille one century ago [40] for characterizing the location of the non-asymptotic zeros of (degenerate) Whittaker hypergeometric functions. For an introduction to confluent hypergeometric functions, we refer to [26]. A motivating example (scalar neutral DDEs of degree three representing the closed-loop dynamics of a transport equation subject to a stabilizing boundary proportional-integral (PI) action, see Section 2) helps to a better understanding of the existing links between over-order MID (double characteristic root in this case) and Hille's method.

It should be noted that the elimination-produced function represents a novelty and generalizes some of the ideas proposed earlier by the authors of this paper. In particular, we recover some of the cases when this function reduces to a polynomial, called elimination-produced polynomial (see Section 3). Next, we further exploit the properties of Kummer hypergeometric functions to handle the dominance property for over-order multiplicities leading to a more unified treatment.

Second, as a byproduct of the analysis proposed in this paper, the location of zeros of linear combination of Kummer (hypergeometric) functions can be derived, and these results allow exploiting MID ideas into a different frame and more precisely in the case of contiguous (degenerate hypergeometric) functions (see, for instance, [62] for further details and related definition ${ }^{12}$ ). To the best of the authors' knowledge, such ideas represent a novelty in the open literature.

Finally, we use the over-order MID to the control of dynamical systems by using the so-called partial pole placement method that simply consists in assigning a given multiplicity of a spectral value for the closed-loop system by an appropriate choice of the controller gains with the guarantee of the exponential stability of the closed-loop system solution with a prescribed rate given by this spectral value. To the best of the authors' knowledge, such ideas represent a novelty and propose a different angle to control dynamical systems represented by DDEs.

Some illustrative case studies show the effectiveness of the method, see for instance [8, 27, 56]. In fact we consider two control problems of two classes of infinite-dimensional systems whose dynamical equations can be represented by DDEs: the control of a transonic flow in a wind tunnel ${ }^{13}$ and the control of a first-order unstable plant including a communication delay in the input/output channel by using a "standard" PID control law ${ }^{14}$. It should be mentioned that our method offers an interesting alternative to the existing tuning rules with a prescribed decay rate for the corresponding closed-loop system.

The remaining of the paper is organized as follows: Prerequisites on degenerate hypergeometric functions, a motivating example (controlling transport equation by a PI controller), the problem formulation, and definitions of over-order multiplicities are presented in Section 2. The methodology for the study of the over-order multiplicities as well as some control perspectives are proposed in Section 3. Next, Section 4 includes two illustrative examples and briefly presents the software $P 3 \delta$ covering the over-order multiplicity. Finally, some concluding remarks end the paper.

Notation. Throughout the paper, the following notations are used: $\mathbb{N}^{*}, \mathbb{R}, \mathbb{C}$ denote the sets of positive integers, real numbers, and complex numbers, respectively, and we set $\mathbb{N}=\mathbb{N}^{*} \cup\{0\}$. For a complex number $\lambda$, $\Re(\lambda)$ and $\Im(\lambda)$ denote its real and imaginary parts, respectively. The open left and right complex half-planes are the sets $\mathbb{C}_{-}$and $\mathbb{C}_{+}$, respectively, defined by $\mathbb{C}_{-}=\{\lambda \in \mathbb{C} \mid \Re(\lambda)<0\}$ and $\mathbb{C}_{+}=\{\lambda \in \mathbb{C} \mid \Re(\lambda)>0\}$. Given $k, n \in \mathbb{N}$ with $k \leq n$, the binomial coefficient $\binom{n}{k}$ is defined as $\binom{n}{k}=\frac{n!}{k!(n-k)!}$ and this notation is extended to $k, n \in \mathbb{Z}$ by setting $\binom{n}{k}=0$ when $n<0, k<0$, or $k>n$. For $\alpha \in \mathbb{C}$ and $k \in \mathbb{N},(\alpha)_{k}$ is the Pochhammer symbol for the ascending factorial, defined inductively as $(\alpha)_{0}=1$ and $(\alpha)_{k+1}=(\alpha+k)(\alpha)_{k}$.

[^3]
## 2 Prerequisites and Problem Formulation

This section provides a brief presentation of the definitions and results that shall be of use in the sequel.

### 2.1 Dynamical systems with delay

Consider the linear time-invariant (LTI) dynamical system described by the DDE including a single delay:

$$
\begin{equation*}
y^{(n)}(t)+\sum_{k=0}^{n-1} a_{k} y^{(k)}(t)+\sum_{k=0}^{m} \alpha_{k} y^{(k)}(t-\tau)=0 \tag{1}
\end{equation*}
$$

under appropriate initial conditions, where $y(\cdot)$ is the real-valued unknown function, $\tau>0$ is the delay, and $a_{0}, \ldots, a_{n-1}, \alpha_{0}, \ldots, \alpha_{m}$ are real coefficients. The DDE (1) is said to be of retarded type ${ }^{15}$ if $m<n$, or of neutral type if $m=n$. We refer to [11], [39], [54] for a deeper discussions on DDEs and related results and properties. Notice that (1) is a particular case of the following matrix form:

$$
\begin{equation*}
\dot{\xi}(t)+B_{\tau} \dot{\xi}(t-\tau)=A_{0} \xi(t)+A_{\tau} \xi(t-\tau) \tag{2}
\end{equation*}
$$

where $\xi=\left(y(t), y^{\prime}(t), \ldots, y^{(n-1)}(t)\right)^{T} \in \mathbb{R}^{n}$ is the state-vector and $A_{0}, A_{\tau}, B_{\tau} \in \mathcal{M}_{n}(\mathbb{R})$ are real-valued matrices which can be easily constructed from (1).

The characteristic function associated with (1) is the quasipolynomial $\Delta: \mathbb{C} \rightarrow \mathbb{C}$ defined by

$$
\begin{equation*}
\Delta(\lambda)=P_{0}(\lambda)+\mathrm{e}^{-\lambda \tau} P_{\tau}(\lambda) \tag{3}
\end{equation*}
$$

where $P_{0}, P_{\tau}$ are polynomials with real coefficients given by:

$$
\left\{\begin{array}{l}
P_{0}(\lambda)=\lambda^{n}+\sum_{k=0}^{n-1} a_{k} \lambda^{k}  \tag{4}\\
P_{\tau}(\lambda)=\sum_{k=0}^{m} b_{k} \lambda_{k}
\end{array}\right.
$$

It is well-known that the exponential stability of the trivial solution of (1) is given by the location of the characteristic roots of $\Delta$, see, e.g., $[11,54]$.

As discussed in [23], the degree of the quasipolynomial $\Delta$, denoted by $\operatorname{deg}(\Delta)$, is nothing else than the Pólya-Szegő bound [64] and, in our case, $\operatorname{deg}(\Delta)=n+m+1$ and it is larger than the degree of the polynomials $P_{0}$ and $P_{\tau}$. It is easy to observe that the quasipolynomial degree corresponds also to the maximal allowable multiplicity that a characteristic root of (4) may have. To reach such a bound, the characteristic root should be real.

### 2.2 Degenerate hypergeometric functions and the corresponding contiguous relations

To develop our results, we need to use some properties of classical hypergoemetric functions. The first such a function we introduce is the Kummer (confluent) hypergeometric function, which, for $a, b \in \mathbb{C}$ such that $-b \notin \mathbb{N}$, is the entire function $\Phi(a, b, \cdot): \mathbb{C} \rightarrow \mathbb{C}$ defined by the series

$$
\begin{equation*}
\Phi(a, b, z)=\sum_{k=0}^{\infty} \frac{(a)_{k}}{(b)_{k}} \frac{z^{k}}{k!} . \tag{5}
\end{equation*}
$$

The series in (5) converges for every $z \in \mathbb{C}$ and, as presented in [26,34, 62], it satisfies the (second-order) Kummer differential equation

$$
\begin{equation*}
z \frac{\partial^{2} \Phi}{\partial z^{2}}(a, b, z)+(b-z) \frac{\partial \Phi}{\partial z}(a, b, z)-a \Phi(a, b, z)=0 . \tag{6}
\end{equation*}
$$

[^4]As discussed in $[26,34,62]$, for every $a, b, z \in \mathbb{C}$ such that $\Re(b)>\Re(a)>0$, Kummer functions also admit the integral representation

$$
\begin{equation*}
\Phi(a, b, z)=\frac{\Gamma(b)}{\Gamma(a) \Gamma(b-a)} \int_{0}^{1} \mathrm{e}^{z t} t^{a-1}(1-t)^{b-a-1} \mathrm{~d} t \tag{7}
\end{equation*}
$$

where $\Gamma$ denotes the Gamma function. This integral representation has been exploited in [50] to characterize the spectrum of some DDEs of retarded type.

Notice that Kummer functions satisfy some recurrence relations often called contiguous relations, see for instance [62]. In particular, in our case, the following relations are of interest:

Lemma 1 ( [62, p. 325]). Let $a \neq b$ and $z \neq 0$ be three complex numbers. The following relations hold:

$$
\begin{align*}
& \Phi(a, b+1, z)=\frac{-b(a+z) \Phi(a, b, z)+a b \Phi(a+1, b, z)}{z(a-b)} \\
& \Phi(a+1, b+1, z)=-\frac{-b \Phi(a+1, b, z)+b \Phi(a, b, z)}{z} \tag{8}
\end{align*}
$$

Kummer confluent hypergeometric functions have close links with Whittaker functions. More precisely, for $k, l \in \mathbb{C}$ with $-2 l \notin \mathbb{N}^{*}$, the Whittaker function $\mathcal{M}_{k, l}$ is the function defined for $z \in \mathbb{C}$ by

$$
\begin{equation*}
\mathcal{M}_{k, l}(z)=\mathrm{e}^{-\frac{z}{2}} z^{\frac{1}{2}+l} \Phi\left(\frac{1}{2}+l-k, 1+2 l, z\right) \tag{9}
\end{equation*}
$$

(see, e.g., [62]). Note that, if $\frac{1}{2}+l$ is not an integer, the function $\mathcal{M}_{k, l}$ is a multi-valued complex function with branch point at $z=0$. The nontrivial roots of $\mathcal{M}_{k, l}$ coincide with those of $\Phi\left(\frac{1}{2}+l-k, 1+2 l, \cdot\right)$ and $\mathcal{M}_{k, l}$ satisfies the (second-order) Whittaker differential equation

$$
\begin{equation*}
\frac{\mathrm{d}^{2} \varphi}{\mathrm{~d} z^{2}}(z)=\left(\frac{1}{4}-\frac{k}{z}+\frac{l^{2}-\frac{1}{4}}{z^{2}}\right) \varphi(z) \tag{10}
\end{equation*}
$$

Since $\mathcal{M}_{k, l}$ is a nontrivial solution of the second-order linear differential equation (10), any nontrivial root of $\mathcal{M}_{k, l}$ is necessarily simple.

### 2.3 Some Insights on Linear Combinations of Two Kummer Functions

Notice that on beyond the standard contiguous relations recalled in Lemma 1 and the other contiguous relations from [62, p. 325], to the best of the authors' knowledge, there does not exist any result describing the distribution of the nonasymptotic zeros of linear combinations of Kummer functions.

The next result, which is established and shown in [18], provides a partial step towards our goal, by giving a non-autonomous second-order differential equation having a given linear combination of two Kummer functions as a solution.

Lemma 2. Let $a, b$ be two complex numbers and $\alpha$ and $\beta$ two real numbers. Then the complex function $F$ defined by

$$
\begin{equation*}
F(z)=\alpha \Phi(a, b, z)+\beta \Phi(a, b+1, z) \tag{11}
\end{equation*}
$$

with $z \notin\left\{0, \frac{\beta(\beta+\alpha) b^{2}}{((a-b) \alpha-\beta b) \alpha}\right\}$, satisfies the second-order differential equation

$$
\begin{equation*}
\frac{\mathrm{d}^{2} \varphi}{\mathrm{~d} z^{2}}(z)+Q(z) \frac{\mathrm{d} \varphi}{\mathrm{~d} z}(z)+R(z) \varphi(z)=0 \tag{12}
\end{equation*}
$$

where

$$
\begin{equation*}
Q(z)=-1+\frac{b+1}{z}-\frac{\alpha(a \alpha-\alpha b-\beta b)}{D(z)} \tag{13}
\end{equation*}
$$

$$
\begin{equation*}
R(z)=-\frac{N(z)}{D(z)} \tag{14}
\end{equation*}
$$

with

$$
\begin{aligned}
& N(z)=a\left(\left((a-b) \alpha^{2}-\alpha b \beta\right) z-\beta b(b+1) \alpha\right)-a b^{2} \beta^{2} \\
& D(z)=\left((a-b) \alpha^{2}-\alpha b \beta\right) z-\alpha b^{2} \beta-b^{2} \beta^{2}
\end{aligned}
$$

Lemma 2 can be proved by using a simple property of $\Phi$, namely, $\frac{\partial \Phi}{\partial z}(a, b, z)=\frac{a}{b} \Phi(a+1, b+1, z)$, which follows immediately from (5), and exploiting the contiguous relations from Lemma 1 . In the sequel, we shall refer to functions $F$ of the form (59) as Kummer-type functions.

Note that Whittaker functions are defined in terms of Kummer functions in (9) by using the multiplicative factor $\mathrm{e}^{-\frac{z}{2}} z^{\frac{1}{2}+l}$, thanks to which the Whittaker differential equation (10) has no first-order term. We now proceed similarly from Kummer-type functions in order to define Whittaker-type functions. The next lemma can be shown by straightforward computations.

Lemma 3. Let $a, b$ be two complex numbers, $\alpha, \beta$ be two real numbers, $F$ be the function defined in (11), and $Q$ and $R$ be given by (13) and (14), respectively.

Let $\mathcal{Q}$ be a primitive of $\frac{Q}{2}$ and define the function $W$ by

$$
\begin{equation*}
W(z)=\mathrm{e}^{\mathcal{Q}(z)} F(z) \tag{15}
\end{equation*}
$$

Then $W$ satisfies the second-order differential equation

$$
\begin{equation*}
\frac{\mathrm{d}^{2} W}{\mathrm{~d} z^{2}}(z)+G(z) W(z)=0 \tag{16}
\end{equation*}
$$

where

$$
\begin{equation*}
G(z)=R(z)-\frac{(Q(z))^{2}}{4}-\frac{1}{2} \frac{\mathrm{~d} Q}{\mathrm{~d} z}(z) \tag{17}
\end{equation*}
$$

In the sequel, we refer to functions $W$ of the form (15) as Whittaker-type functions.
As discussed in the previous section, beyond the standard contiguous relation, to the best of the authors' knowledge, there does not exist any result describing the distribution of the non-asymptotic zeros' of linear combinations of Kummer functions.

### 2.4 Hille oscillation theorems

In [40], Hille studied the distribution of zeros of functions of a complex variable satisfying linear secondorder homogeneous differential equations with variable coefficients, as is the case for the degenerate Whittaker function $\mathcal{M}_{k, l}$, which satisfies (10). Thanks to an integral transformation defined there and called Green-Hille transformation, and some further conditions on the behavior of the function, Hille showed how to discard regions in the complex plane from including complex roots. Consider, for instance, the general homogeneous second-order differential equation:

$$
\begin{equation*}
\frac{\mathrm{d}}{\mathrm{~d} z}\left[K(z) \frac{\mathrm{d} \varphi}{\mathrm{~d} z}(z)\right]+G(z) \varphi(z)=0 \tag{18}
\end{equation*}
$$

where $z$ is complex and the functions $G$ and $K$ are assumed analytic in some region $\Omega$ such that $K$ does not vanish in that region. Equation (18) can be written in $\Omega$ as a second-order system on the unknown functions $\varphi_{1}(z)=\varphi(z)$ and $\varphi_{2}(z)=K(z) \frac{d \varphi}{d z}(z)$, and the Green-Hille transformation consists on multiplying the equation on $\varphi_{1}$ by $\overline{\varphi_{2}(z)}$, that on $\varphi_{2}$ by $\overline{\varphi_{1}(z)}$, and integrating on $z$ along a path in $\Omega$, which yields:

$$
\begin{equation*}
\left[\overline{\varphi_{1}(z)} \varphi_{2}(z)\right]_{z_{1}}^{z_{2}}-\int_{z_{1}}^{z_{2}}\left|\varphi_{2}(z)\right|^{2} \frac{\overline{\mathrm{~d} z}}{\overline{K(z)}}+\int_{z_{1}}^{z_{2}}\left|\varphi_{1}(z)\right|^{2} G(z) \mathrm{d} z=0 \tag{19}
\end{equation*}
$$

where $z_{1}, z_{2} \in \Omega$ and both integrals are taken along the same arbitrary smooth path in $\Omega$ connecting $z_{1}$ to $z_{2}$.
The following result, which is proved in [17] using the Green-Hille transformation from [40], gives insights on the distribution of the non asymptotic zeros of Kummer hypergeometric functions with real arguments $a$ and $b$.

Proposition 2.1 ([17]). Let $a, b \in \mathbb{R}$ be such that $b \geq 2$.

1. If $b=2 a$, then all nontrivial roots $z$ of $\Phi(a, b, \cdot)$ are purely imaginary.
2. If $b>2 a$ (resp., $b<2 a$ ), then all nontrivial roots $z$ of $\Phi(a, b, \cdot)$ satisfy $\Re(z)>0$ (resp., $\Re(z)<0)$.
3. If $b \neq 2 a$, then all nontrivial roots $z$ of $\Phi(a, b, \cdot)$ satisfy

$$
(b-2 a)^{2} \Im(z)^{2}-(4 a(b-a)-2 b) \Re(z)^{2}>0
$$

Remark 1. It should be noted that a quasipolynomial admitting a characteristic root with intermediate multiplicity necessarily shares its remaining roots with an appropriate linear combination of Kummer functions. Unfortunately, to the best of the authors' knowledge, there does not exist any general result in the open literature to describe the distribution of the non-asymptotic zeros of such a function combination.

### 2.5 Delay Systems Frequency Bound in the Right Half-plane

Despite the unquestionable interest and insights of the Hille oscillation Theorems [40] and the Green-Hille transform in discarding regions in the complex plane from containing zeros of a meromorphic function which is a solution of a given second-order differential equation, such an approach lacks effectiveness from the numerical point of view.

In this preliminary subsection, we provide a numerical-oriented alternative for the Green-Hille transform leading to an effective algorithm that we are explicitly using in our framework. In this sense, we first establish conditions on the system's parameters which guarantee the existence of a multiple root. Second, perform an affine change of variable in the characteristic equation $\Delta(\lambda)=P_{0}(\lambda)+P_{\tau}(\lambda) \mathrm{e}^{-\lambda \tau}$ in order to reduce the corresponding quasipolynomial to a normalized form: $\tilde{\Delta}(z)=\tilde{P}_{0}(z)+\tilde{P}_{\tau}(z) \mathrm{e}^{-z}$. Next, we derive a bound on the imaginary part of roots of the normalized characteristic function in the complex right half-plane. Lastly, a certification of the dominance of the multiple root is demonstrated. In what follows, Algorithm 1 lists the steps to be followed to reach a suitable frequency bound (see $[13,51]$ ).

### 2.6 Multiplicity-induced-dominancy property in DDEs and related topics

As briefly explained in the Introduction, a characteristic root $\lambda_{0}$ (of $\Delta$ ) satisfies the MID property if
(i) its algebraic multiplicity (denoted $M=M\left(\lambda_{0}\right)$ ) is larger than one,
(ii) it is dominant in the sense that the remaining characteristic roots $\lambda_{\sigma}$ of the spectrum satisfy the condition $\Re\left(\lambda_{\sigma}\right) \leq \Re\left(\lambda_{0}\right)$.

Since the maximal allowable multiplicity is defined by the degree of the quasipolynomial $\Delta$ (see, e.g., [50]), it is clear that $2 \leq M \leq \operatorname{deg}(\Delta)$. The case $M=\operatorname{deg}(\Delta)$ is called generic multiplicity, and any multiplicity larger than one and smaller than $\operatorname{deg}(\Delta)$ denotes an intermediate multiplicity. Regarding these intermediate multiplicities, when compared to $\operatorname{deg}\left(P_{0}\right)=n$, there are two sub-cases:
(i.1) sub-order multiplicity, if $1<M \leq n=\operatorname{deg}\left(P_{0}\right)$, and
(i.2) over-order multiplicity, if $1+\operatorname{deg}\left(P_{0}\right)=1+n \leq M \leq \operatorname{deg}(\Delta)$.

```
Algorithm 1 Estimation of the MID frequency bound in time-delay differential equations with single delay
Require: \(\tilde{\Delta}(z)=\tilde{P}_{0}(z)+\tilde{P}_{\tau}(z) \mathrm{e}^{-z} / /\) Normalized quasipolynomial
    // Initialization
    ord \(=0\)
    // ord: order of truncation of the Taylor expansion of \(\mathrm{e}^{2 x}=\underbrace{\underbrace{1}_{\text {ord }=0}}_{\text {ord }=1}+2 x+2 x^{2}+\frac{4 x^{3}}{3}+\cdots\)
    dominance \(=\) false
    \(\exists z_{0}=x+i \omega \in \mathbb{R}_{+}^{*}+i \mathbb{R}_{+}^{*}\) s.t. \(\tilde{\Delta}\left(z_{0}\right)=0\)
    \(\left|\tilde{P}_{0}(x+i \omega)\right|^{2} \mathrm{e}^{2 x}=\left|\tilde{P}_{\tau}(x+i \omega)\right|^{2}\)
    while \(\sim\) dominance do
        ord \(=\) ord +1
        \(F(x, \omega)=\left|\tilde{P}_{\tau}(x+i \omega)\right|^{2}-\left|\tilde{P}_{0}(x+i \omega)\right|^{2} \mathrm{~T}_{\text {ord }}\left(\mathrm{e}^{2 x}\right)>0\)
        \(/ / \mathrm{T}_{\text {ord }}\left(\mathrm{e}^{2 x}\right)\) : Taylor expansion of \(\mathrm{e}^{2 x}\) of order \(=\) ord
        \(\omega^{2}=\Omega\)
        \(H(x, \Omega)\) // The polynomial characterizing the real roots of \(F\)
    end while
    \(\Omega_{k}(x) / / \mathrm{k}^{\text {th }}\) real root of \(H\), depend on free parameters
    if \(\max _{x}\left(\max _{k}\left(\Omega_{k}(x)\right)\right)<\pi^{2}\) then
        dominance \(=\) true
    end if
    return Frequency bound
```

With these definitions and notations, the smallest over-order multiplicity is given by one added to the degree of the polynomial $P$, and the largest is given by the degree of the characteristic function $\Delta$ and it corresponds to the so-called generic multiplicity. The second property states that $\lambda_{0}$ should be the rightmost root of the spectrum and defines the spectral abscissa of the corresponding quasipolynomial $\Delta$ (see, for instance, [54] and the references therein).

For a better understanding of the notions, consider the following simple example. In the case of secondorder DDEs of retarded type including one delay with $\operatorname{deg}\left(P_{0}\right)=2, \operatorname{deg}\left(P_{1}\right)=1$, we have $\operatorname{deg}(\Delta)=4$. Thus, in this case, the generic multiplicity is four, the intermediate multiplicities are two and three, and the only over-order multiplicities are three and four since the order of the DDE is two. In this configuration, the only sub-order multiplicity is two.

Our main objective is to address the MID property in the case of over-order multiplicities. It should be mentioned that the generic case, i.e., $M\left(\lambda_{0}\right)=\operatorname{deg}(\Delta)$, was already treated and explicit characterizations exist: the retarded case $m=n-1$ was characterized in [50] and a unified treatment of the retarded and neutral cases, where $m \leq n$, was provided in [15].

To the best of the authors' knowledge, in the open literature, the MID with over-order multiplicities has been addressed in two particular configurations $M\left(\lambda_{0}\right)=n+1$ and $M\left(\lambda_{0}\right)=n+m$. More precisely, in the "limit" case $M\left(\lambda_{0}\right)=n+1$, sufficient conditions for their validity have been proposed in [9], where the authors exploited the particular spectrum location of the open-loop plant. Finally, by using a different argument inspired by the Green-Hille transformation, [18] treated the case $M\left(\lambda_{0}\right)=n+m$ corresponding to the other "limit" case. Finally, it should be mentioned that in the case of second-order DDEs of retarded type with a single delay, these two limits cases coincide and they were treated in [24].

### 2.7 Motivating example: Boundary control of the transport equation

To illustrate the use of the MID property for hyperbolic PDE control purposes, let us revisit the problem of high-volume, multistage continuous production flow through a re-entrant factory control [6, 29], which, in its
linear version, amounts to the problem of exponential stabilization of the standard scalar conservation law:

$$
\begin{equation*}
\partial_{t} \varphi(t, x)+\eta \partial_{x} \varphi(t, x)=0, t \in[0, \infty), x \in(0, L) \tag{20}
\end{equation*}
$$

that is a "standard" linear transport equation where $\varphi(t, x)$ represents the density or the concentration at position $x \in(0, L)$, with $L>0$, and in time $t \in[0,+\infty)$ of the corresponding physical quantity of interest. Under the assumption that the diffusion is neglected, the linear mapping $\varphi \mapsto \eta \varphi$, with $\eta>0$ defines the flux function ${ }^{16}$. Finally, the quantity $\eta$ represents the so-called advection speed or the velocity of propagation. In view of boundary controlling the dynamics of Equation (20), the standard PI controller is proposed in [29]:

$$
\begin{equation*}
\varphi(t, 0)=k_{p} \varphi(t, L)+k_{i} \int_{0}^{t} \varphi(\nu, L) \mathrm{d} \nu \tag{21}
\end{equation*}
$$

where $k_{p}$ and $k_{i}$ are the feedback parameters representing proportional and integral control gains. By using the Laplace transform to both sides of the boundary condition (21) and multiplying by $\lambda$, one gets the following ${ }^{17}$ :

$$
\begin{equation*}
\Delta(\lambda)=\lambda-\left(k_{i}+k_{p} \lambda\right) \mathrm{e}^{-\frac{L}{\eta} \lambda} \tag{22}
\end{equation*}
$$

which is a scalar quasipolynomial of neutral type with the delay given by the ratio between the length and the speed of propagation $L / \eta$.

As discussed in the previous sections, the degree of the quasipolynomial $\Delta$ is equal to 3 . Recall also that, thanks to the results from [15], the maximal multiplicity can be achieved only by a real root. Furthermore, if such a maximal multiplicity is reached, the GMID property holds and the corresponding triple characteristic root $\lambda_{0}=-\frac{2 \eta}{L}$ is necessarily the spectral abscissa. Despite the interest of this property from a purely analytic view point, a control implementation based on the GMID lacks of robustness, see for instance [52]. As a matter of fact, for the sake of robustness with respect to the model's parametric uncertainties, it appears that it will be more appropriate to relax constraints on the choice of the closed-loop spectral abscissa. This can be carried out using the MID property by assigning a root with an over-order intermediate multiplicity (in this case, multiplicity two):

Proposition 2.2. For any $\lambda_{0}$ satisfying $-\eta / L<\lambda_{0}<0$, the PI controller given by:

$$
\begin{equation*}
k_{i}=-\frac{\lambda_{0}^{2} L \mathrm{e}^{\frac{L \lambda_{0}}{\eta}}}{\eta}, \quad k_{p}=\frac{\mathrm{e}^{\frac{L \lambda_{0}}{\eta}}\left(L \lambda_{0}+\eta\right)}{\eta} \tag{23}
\end{equation*}
$$

stabilizes the system (20). Furthermore, the intermediate MID property holds with an exponential decay rate $\lambda_{0}$ for the closed-loop system.

Before providing the proof of Proposition 2.2 , let us consider the generic quasipolynomial $\Delta: \mathbb{C} \rightarrow \mathbb{C}$ defined by

$$
\begin{equation*}
\Delta(\lambda)=\lambda+a+\mathrm{e}^{-\lambda \tau}\left(b_{1} \lambda+b_{0}\right) \tag{24}
\end{equation*}
$$

This corresponds to the scalar neutral systems studied in [12]. In this case, $\operatorname{deg}(\Delta)=3$ and we are interested in the dominance of roots attaining the intermediate multiplicity 2 , which is an over-order multiplicity since the neutral DDE is scalar.

A number $\lambda_{0} \in \mathbb{R}$ is a root of multiplicity at least 2 of $\Delta$ if and only if the coefficients $b_{1}$ and $b_{0}$ satisfy

$$
\begin{align*}
& b_{1}=-\mathrm{e}^{\lambda_{0} \tau}\left(1+\left(a+\lambda_{0}\right) \tau\right),  \tag{25}\\
& b_{0}=\mathrm{e}^{\lambda_{0} \tau}\left(\left(a+\lambda_{0}\right) \lambda_{0} \tau-a\right) .
\end{align*}
$$

We first remark that, under conditions (25), (24) can be rewritten as

$$
\Delta(\lambda)=\lambda+a-\mathrm{e}^{-\left(\lambda-\lambda_{0}\right) \tau}\left(\left(a+\lambda_{0}\right)\left(\lambda-\lambda_{0}\right) \tau+\lambda+a\right) .
$$

[^5]We perform the change of variables in $\mathbb{C}$ corresponding to the new variable $z=\tau\left(\lambda-\lambda_{0}\right)$. More precisely, consider $\widetilde{\Delta}(z)=\tau \Delta\left(\lambda_{0}+\frac{z}{\tau}\right)$, which satisfies

$$
\widetilde{\Delta}(z)=z+\alpha-\mathrm{e}^{-z}(z(1+\alpha)+\alpha)
$$

where $\alpha=\tau\left(a+\lambda_{0}\right)$. Note that $\lambda_{0}$ is a dominant root of (24) if and only if 0 is a dominant root of $\widetilde{\Delta}$.
A straightforward computation shows that

$$
\widetilde{\Delta}(z)=z^{2} \int_{0}^{1}[1+\alpha-\alpha(1-t)] \mathrm{e}^{-z t} \mathrm{~d} t
$$

and thus, using the integral representation (7) of Kummer functions, we deduce that

$$
\begin{equation*}
\widetilde{\Delta}(z)=z^{2}\left[(1+\alpha) \Phi(1,2,-z)-\frac{\alpha}{2} \Phi(1,3,-z)\right] . \tag{26}
\end{equation*}
$$

Let $F: \mathbb{C} \rightarrow \mathbb{C}$ be the function defined by

$$
F(z)=(1+\alpha) \Phi(1,2, z)-\frac{\alpha}{2} \Phi(1,3, z)
$$

Thanks to (26), the root 0 is a dominant root of $\widetilde{\Delta}$ if and only if all roots of $F$ have nonnegative real part. To study the sign of the real part of the roots of $F$, we apply Hille's method as described in Section 2.2. Notice first that, by Lemma 2, $F$ satisfies the second-order equation (12) with $\alpha$ replaced by $1+\alpha$ and $\beta$ replaced by $-\frac{\alpha}{2}$. We now perform a transformation similar to that used to obtain Whittaker functions from Kummer functions. More precisely, we multiply $F$ by a function of $z$ in such a way that the second-order ODE satisfied by the product has no first-order term. This can be achieved introducing the Whittaker-type function

$$
W(z)=\frac{\mathrm{e}^{-z / 2} z^{3 / 2}}{\sqrt{(1+\alpha) z-\alpha(\alpha+2)}} F(z)
$$

which satisfies

$$
\begin{equation*}
\frac{\mathrm{d}^{2} W}{\mathrm{~d} z^{2}}(z)+G(z) W(z)=0 \tag{27}
\end{equation*}
$$

where $G$ is given by

$$
\begin{equation*}
G(z)=-\frac{(\alpha+1)^{2} z^{4}-2 \alpha(\alpha+1)(\alpha+2) z^{3}+\alpha^{2}\left(1+(\alpha+1)^{2}\right) z^{2}+2 \alpha^{3}(\alpha+2) z+3 \alpha^{2}(\alpha+2)^{2}}{4 z^{2}((\alpha+1) z-\alpha(\alpha+2))^{2}} \tag{28}
\end{equation*}
$$

Note that $G$ can be rewritten as

$$
\begin{aligned}
G(z)= & -\frac{1}{4}-\frac{3}{4 z^{2}}-\frac{3(\alpha+1)^{2}}{4((\alpha+1) z-\alpha(\alpha+2))^{2}} \\
& -\frac{\alpha^{2}+3 \alpha+3}{2 \alpha(\alpha+2) z}+\frac{(\alpha+1)\left(\alpha^{2}+3 \alpha+3\right)}{2 \alpha(\alpha+2)((\alpha+1) z-\alpha(\alpha+2))}
\end{aligned}
$$

Applying Hille's method to (27), we obtain, by taking in (18) $z_{1}=0$ and $z_{2}$ equal to a root $z_{*}$ of $F$, we deduce that

$$
\int_{0}^{z_{*}}\left|W^{\prime}(z)\right|^{2} \overline{\mathrm{~d} z}=\int_{0}^{z_{*}}|W(z)|^{2} G(z) \mathrm{d} z
$$

We choose as integration path the line segment from 0 to $z_{*}$. Hence

$$
\overline{z_{*}} \int_{0}^{1}\left|W^{\prime}\left(t z_{*}\right)\right|^{2} \mathrm{~d} t=z_{*} \int_{0}^{1}\left|W\left(t z_{*}\right)\right|^{2} G\left(t z_{*}\right) \mathrm{d} t .
$$

Taking the real part, we get

$$
x_{*} \int_{0}^{1}\left|W^{\prime}\left(t z_{*}\right)\right|^{2} \mathrm{~d} t=\int_{0}^{1}\left|W\left(t z_{*}\right)\right|^{2} \Re\left[z_{*} G\left(t z_{*}\right)\right] \mathrm{d} t .
$$

where $x_{*}=\Re\left(z_{*}\right)$ and $y_{*}=\Im\left(z_{*}\right)$.
So, a sufficient condition for $F$ to admit only roots with positive real part is that

$$
\begin{equation*}
\Re\left[-\frac{z}{4}-\frac{3}{4 t^{2} z}-\frac{3 z(\alpha+1)^{2}}{4((\alpha+1) t z-\alpha(\alpha+2))^{2}}-\frac{\alpha^{2}+3 \alpha+3}{2 \alpha(\alpha+2) t}+\frac{z(\alpha+1)\left(\alpha^{2}+3 \alpha+3\right)}{2 \alpha(\alpha+2)((\alpha+1) t z-\alpha(\alpha+2))}\right] \geq 0 \tag{29}
\end{equation*}
$$

for every $t \in(0,1)$ and $z \in \mathbb{C}_{-}$, where $\alpha=\left(a+\lambda_{0}\right) \tau$. Then $\lambda_{0}$ is a dominant root of $\Delta$, i.e., $\Re(\lambda) \leq \lambda_{0}$ for every root $\lambda$ of $\Delta$.

As emphasized in the previous sections, Hille oscillation method provides only implicit sufficient conditions. So that, in the following proof we exploit the numerical alternative provided in Section 2.5.

Proof of Proposition 2.2. Using a standard elimination procedure (thanks to the linear dependency of the quasipolynomial with respect to its parameters) allows to show that a real $\lambda_{0}$ is a root of multiplicity 2 of $\Delta$ if and only if the PI gains $k_{p}$ and $k_{i}$ satisfy (23). Thus, the characteristic function can be written as:

$$
\begin{equation*}
\tilde{\Delta}(z)=\tilde{P}_{0}(z)+\tilde{P}_{\tau}(z) \mathrm{e}^{-z} \tag{30}
\end{equation*}
$$

where $\tilde{P}_{0}(z)=z+\rho$, and $\tilde{P}_{\tau}(z)=(-\rho-1) z-\rho$, with $\rho=L \lambda_{0} / \eta$. A necessary condition for the exponential stability is given by the stability of the corresponding delay-difference operator (see, e.g. [39]). Thus, $\rho$ should verify $|\rho+1|<1$, that is $\rho \in(-2,0)$. Next, $\tilde{\Delta}$ in (30) can be written as:

$$
\begin{equation*}
\tilde{\Delta}(z)=z^{2} \int_{0}^{1} q_{\rho}(t) \mathrm{e}^{-t z} \mathrm{~d} t \quad \text { where } \quad q_{\rho}(t)=\rho t+1 \tag{31}
\end{equation*}
$$

In our approach, $q_{\rho}$ should keep a constant sign for $t \in(0,1)$, which happens if an only if $\rho \geq-1$, and then we concentrate in the interval of interest $\rho \in(-1,0)$. Next, one follows the steps of Algorithm 1 by setting $z_{0}=x_{0}+i \omega_{0} \in \mathbb{R}_{+}+i \mathbb{R}_{+}$as a root of $\tilde{\Delta}(z)$, we then have

$$
\left|\tilde{P}_{0}\left(x_{0}+i \omega_{0}\right)\right|^{2} \mathrm{e}^{2 x_{0}}=\left|\tilde{P}_{\tau}\left(x_{0}+i \omega_{0}\right)\right|^{2}
$$

Since $\mathrm{e}^{2 x}>1+2 x$ for any $x \in \mathbb{R}_{+}$, the real-valued function $F_{\rho}(x, \omega)=\left|\tilde{P}_{\tau}(x+i \omega)\right|^{2}-\mid \tilde{P}_{0}(x+$ $i \omega)\left.\right|^{2}(1+2 x)$ satisfies $F_{\rho}\left(x_{0}, \omega_{0}\right)>0$. In fact, $F_{\rho}$ is nothing but the following quadratic polynomial in (real) $\omega$

$$
\begin{equation*}
F_{\rho}(x, \omega)=b_{\rho}(x) \omega^{2}+c_{\rho}(x) \tag{32}
\end{equation*}
$$

where $b_{\rho}=-2 x+2 \rho+\rho^{2}, c_{\rho}=-2 x^{3}+\left(\rho^{2}-2 \rho\right) x^{2}$. By setting $x^{ \pm}=\frac{\rho^{2}}{2} \pm \rho$, one easily checks that $F_{\rho}$ is positive for $x \in\left(x^{+}, x^{-}\right)$if $\rho \in(-1,0)$.

The next step consists in characterizing the frequency bound for potential unstable roots if $\rho \in(-1,0)$. Since the discriminant of $F_{\rho}$ defined in (32) is positive, then $F_{\rho}$ admits two real roots:

$$
\omega_{\rho}^{ \pm}(x)=\mp \frac{\sqrt{-\left(\rho^{2}+2 \rho-2 x\right)\left(\rho^{2}-2 \rho-2 x\right)} x}{\rho^{2}+2 \rho-2 x}
$$

where $\omega_{\rho}^{+}$denotes the greatest solution. Since $\rho \in(-1,0)$ and $x>0, \omega_{\rho}^{+}$is upper bounded with respect to $\rho$ by the parameter-free expression $\omega^{+}(x)=\frac{x \sqrt{-4 x^{2}+3}}{1+2 x}$, which reaches a maximum value at $x^{*}=\frac{\sqrt{3}}{2}$. Thus, $\omega=\omega_{\rho}^{+}(x) \leq \omega^{+}\left(x^{*}\right) \approx 0.5899<\pi$. In other words, all unstable solutions $z_{u}$ of $\tilde{\Delta}$ should satisfy the condition $0<\Im\left(z_{u}\right)<\pi$.

Finally, by a contradiction argument, one assumes that such an unstable root $z_{u}$ exists. Then, the integral representation yields $\int_{0}^{1}(\rho t+1) \mathrm{e}^{-t z_{u}} \mathrm{~d} t=0$, the imaginary part of which is $\int_{0}^{1} t(\rho t+1) \mathrm{e}^{-t x} \sin (\omega t) \mathrm{d} t=0$. Now, the frequency bound $0<\omega \leq \pi$ of the previous step entails that the function $t \mapsto t(\rho t+1) \mathrm{e}^{-x t} \sin (\omega t)$ is strictly positive in $(0,1)$, thereby contradicting the last equality. This ends the proof.


Figure 1: Numerical simulation of a solution of (20)-(21) satisfying (23) with parameters $L=1 \eta=1$ and initial condition $\varphi(0, x)=\sin (\pi x)$.

As an illustration of Proposition 2.2, Fig. 1 represents a solution $\varphi$ of (20) under the PI controller (21) with parameters $k_{i}$ and $k_{p}$ computed from (23) with the choice $\lambda_{0}=-\frac{\eta}{2 L}$. The numerical solution was computed using a standard finite difference upwind scheme to discretize (20) and using Simpson's rule to approximate the integral in (21).

### 2.8 Problem Formulation

As emphasized in the Introduction, the aim of this paper is to further exploit the zeros distribution of linear combination of Kummer/Whittaker functions in the control of dynamical systems represented by DDEs by using the partial pole placement. As a matter of fact, our goal is to characterize quasipolynomials with multiple roots as a particular linear combinations of Kummer functions $\sum_{k=0}^{M-1} \zeta_{k} \Phi(a, b+k,-z)$. Next, for a prescribed $\lambda_{0}$ we aim to establish conditions discarding the right half-place $\left\{\lambda \in \mathbb{C} \mid \Re(\lambda)>\lambda_{0}\right\}$ from containing zeros of such functions. This step allows to show the validity of the so-called MID property with over-order intermediate multiplicities i.e. to show the dominancy of the multiple root among the set of all the remaining zeros of the considered quasipolynomial. Consider now the DDE (1) with the characteristic function:

$$
\Delta: \mathbb{C} \rightarrow \mathbb{C}, \quad \Delta(\lambda)=P_{0}(\lambda)+P_{\tau}(\lambda) \mathrm{e}^{-\lambda \tau}
$$

where $\operatorname{deg}\left(P_{0}\right)=n, \operatorname{deg}\left(P_{\tau}\right)=m \leq n$. As indicated in [23], $\operatorname{deg}(\Delta)=n+m+1$.
The problem addressed in this paper can be formulated as follows: finding conditions on the parameters of the dynamical system (1) such that a (real) characteristic root $\lambda_{0}$ with intermediate algebraic multiplicity $n+1 \leq M\left(\lambda_{0}\right) \leq n+m$ satisfies the MID property, i.e., it corresponds to the rightmost characteristic root.

## 3 Main Results

With the definitions, notations and prerequisites above, we are now able to state our main results.

### 3.1 Necessary and Sufficient Conditions guaranteeing Multiplicity $n+1 \leq M\left(\lambda_{0}\right) \leq$

 $n+m+1$Thanks to the preliminary results of Section 2.3, we are now in position to prove the following result, providing a necessary and sufficient condition for a given real number $\lambda_{0}$ to be a root of a quasipolynomial $\Delta$ with multiplicity between $n+1$ and $n+m+1$.

Theorem 3.1. Let $n$, $m$ be nonnegative integers with $n \geq m, \tau>0, \lambda_{0} \in \mathbb{R}, M \in[n+1, n+m+1]$ be an integer, and consider the quasipolynomial $\Delta$ from (3)-(4). The number $\lambda_{0}$ is a root of multiplicity at least $M$ of $\Delta$ if and only if there exists a polynomial pof degree $n+m+1-M$ with $p(0)=1$ such that

$$
\begin{equation*}
\Delta(\lambda)=\frac{\tau^{M-n}\left(\lambda-\lambda_{0}\right)^{M}}{(M-n-1)!} \int_{0}^{1} t^{M-n-1}(1-t)^{M-m-1} p(t) \mathrm{e}^{-t \tau\left(\lambda-\lambda_{0}\right)} \mathrm{d} t . \tag{33}
\end{equation*}
$$

Proof. Let $\mathcal{V}$ be the set of all functions $\Delta$ of the form $\Delta(\lambda)=P_{0}(\lambda)+\mathrm{e}^{-\lambda \tau} P_{\tau}(\lambda)$ with $P_{0}$ and $P_{\tau}$ given by (4). Note that $\mathcal{V}$ is is a real affine space with $\operatorname{dim} \mathcal{V}=n+m+1$, which is an affine subspace of the space of all entire complex functions, seen as a real vector space. In addition, $\mathcal{V}$ can be canonically identified with $\mathbb{R}^{2 n}$ by identifying a quasipolynomial $\Delta$ with its coefficients $a_{0}, \ldots, a_{n-1}, b_{0}, \ldots, b_{m}$.

Let us denote by $\mathcal{V}_{\lambda_{0}}^{M}$ the subset of $\mathcal{V}$ of those functions $\Delta$ admitting $\lambda_{0}$ as a root of multiplicity at least $M$, i.e.,

$$
\mathcal{V}_{\lambda_{0}}^{M}=\left\{\Delta \in \mathcal{V} \mid \Delta^{(k)}\left(\lambda_{0}\right)=0 \text { for all } k \in\{0, \ldots, M-1\}\right\} .
$$

Each equation $\Delta^{(k)}\left(\lambda_{0}\right)=0, k \in\{0, \ldots, M-1\}$, defines a hyperplane in $\mathcal{V}$ and, when identifying $\mathcal{V}$ with the Euclidean space $\mathbb{R}^{2 n}$, the normal vectors to all such hyperplanes are linearly independent. Hence $\mathcal{V}_{\lambda_{0}}^{M}$ is a subspace of $\mathcal{V}$ of codimension $M$, i.e., $\operatorname{dim} \mathcal{V}_{\lambda_{0}}=n+m+1-M$.

Introduce now $\mathcal{W}_{\lambda_{0}}^{M}$ as the space of all functions $\Delta$ of the form (33) for some polynomial $p$ of degree $n+m+1-M$ with $p(0)=1$. The set $\mathcal{W}_{\lambda_{0}}^{M}$ is an affine subspace of the space of all entire complex functions, seen once again as a real vector space, with $\operatorname{dim} \mathcal{W}_{\lambda_{0}}^{M}=n+m+1-M$.

As a first step, we will prove that $\mathcal{W}_{\lambda_{0}}^{M} \subset \mathcal{V}$, i.e., we show that every function $\Delta$ of the form (33) is indeed a quasipolynomial of the form (3)-(4). To do so, we first notice that, by an immediate inductive integration by parts, we have (see also [50, Proposition 2.1])

$$
\begin{equation*}
\int_{0}^{1} q(t) \mathrm{e}^{-z t} \mathrm{~d} t=\sum_{k=0}^{d} \frac{q^{(k)}(0)-q^{(k)}(1) \mathrm{e}^{-z}}{z^{k+1}} \tag{34}
\end{equation*}
$$

for every $z \in \mathbb{C} \backslash\{0\}, d \in \mathbb{N}$, and $q$ a polynomial of degree $d$. Now, let $\Delta \in \mathcal{W}_{\lambda_{0}}^{M}$ and $p$ be a polynomial of degree $n+m+1-M$ with $p(0)=1$ be such that $\Delta$ is given by (33). Define $q(t)=t^{M-n-1}(1-t)^{M-m-1} p(t)$ and notice that $q(0)=q^{\prime}(0)=\cdots=q^{(M-n-2)}(0)=0, q^{(M-n-1)}(0)=(M-n-1)!$, and $q(1)=q^{\prime}(1)=$ $\cdots=q^{(M-m-2)}(1)=0$. By using (34), we deduce that

$$
\begin{align*}
\Delta(\lambda)= & \frac{\tau^{M-n}\left(\lambda-\lambda_{0}\right)^{M}}{(M-n-1)!} \sum_{k=0}^{M-1} \frac{q^{(k)}(0)-q^{(k)}(1) \mathrm{e}^{-\tau\left(\lambda-\lambda_{0}\right)}}{\tau^{k+1}\left(\lambda-\lambda_{0}\right)^{k+1}} \\
= & \left(\lambda-\lambda_{0}\right)^{n}+\sum_{k=0}^{n-1} \frac{q^{(M-k-1)}(0)}{\tau^{n-k}(M-n-1)!}\left(\lambda-\lambda_{0}\right)^{k}  \tag{35}\\
& -\mathrm{e}^{-\tau\left(\lambda-\lambda_{0}\right)} \sum_{k=0}^{m} \frac{q^{(M-k-1)}(1)}{\tau^{n-k}(M-n-1)!}\left(\lambda-\lambda_{0}\right)^{k},
\end{align*}
$$

and thus, as required, $\Delta \in \mathcal{V}$.
We now notice that $\mathcal{W}_{\lambda_{0}}^{M} \subset \mathcal{V}_{\lambda_{0}}^{M}$, since, by construction, for any $\Delta$ given by (33), $\lambda_{0}$ is clearly a root of multiplicity at least $M$ of $\Delta$. Finally, since $\mathcal{W}_{\lambda_{0}}^{M}$ and $\mathcal{V}_{\lambda_{0}}^{M}$ are both affine spaces with the same dimension, we conclude that $\mathcal{W}_{\lambda_{0}}^{M}=\mathcal{V}_{\lambda_{0}}^{M}$, yielding the conclusion.

Remark 2. Note that (35) provides explicit expressions for the polynomials $P_{0}$ and $P_{\tau}$ from (4) in terms of the polynomial $q$ introduced in the above proof. More precisely, we have

$$
\begin{aligned}
& P_{0}(\lambda)=\left(\lambda-\lambda_{0}\right)^{n}+\sum_{k=0}^{n-1} \frac{q^{(M-k-1)}(0)}{\tau^{n-k}(M-n-1)!}\left(\lambda-\lambda_{0}\right)^{k} \\
& P_{\tau}(\lambda)=-\mathrm{e}^{\tau \lambda_{0}} \sum_{k=0}^{m} \frac{q^{(M-k-1)}(1)}{\tau^{n-k}(M-n-1)!}\left(\lambda-\lambda_{0}\right)^{k} .
\end{aligned}
$$

Recalling that $q(t)=t^{M-n-1}(1-t)^{M-m-1} p(t)$, one may further express $P_{0}$ and $P_{\tau}$ in terms of $p$. Indeed, we have

$$
\begin{aligned}
\frac{q^{(M-k-1)}(0)}{(M-n-1)!} & =\sum_{j=0}^{n-k}\binom{M-k-1}{n-k}\binom{n-k}{j} \frac{(-1)^{j}(M-m-1)!}{(M-m-j-1)!} p^{(n-k-j)}(0), \\
\frac{q^{(M-k-1)}(0)}{(M-n-1)!} & =\sum_{j=0}^{m-k}\binom{M-k-1}{m-k}\binom{m-k}{j} \frac{(-1)^{M-m-1}(M-m-1)!}{(M-n-j-1)!} p^{(m-k-j)}(1),
\end{aligned}
$$

and thus

$$
\begin{align*}
& P_{0}(\lambda)=\left(\lambda-\lambda_{0}\right)^{n}+\sum_{k=0}^{n-1} \sum_{j=0}^{n-k} \frac{(-1)^{j}}{\tau^{n-k}}\binom{M-k-1}{n-k}\binom{n-k}{j} \frac{(M-m-1)!}{(M-m-j-1)!} p^{(n-k-j)}(0)\left(\lambda-\lambda_{0}\right)^{k}  \tag{36}\\
& P_{\tau}(\lambda)=-\mathrm{e}^{\tau \lambda} \sum_{k=0}^{m} \sum_{j=0}^{m-k} \frac{(-1)^{M-m-1}}{\tau^{n-k}}\binom{M-k-1}{m-k}\binom{m-k}{j} \frac{(M-m-1)!}{(M-n-j-1)!} p^{(m-k-j)}(1)\left(\lambda-\lambda_{0}\right)^{k} . \tag{37}
\end{align*}
$$

Proposition 3.1. Let $n$, $m$ be nonnegative integers with $n \geq m, \tau>0, \lambda_{0} \in \mathbb{R}, M \in[n+1, n+m+1]$ be an integer, and consider the quasipolynomial $\Delta$ from (3)-(4). The number $\lambda_{0}$ is a root of multiplicity at least $M$ of $\Delta$ if and only if there exist real numbers $\pi_{0}, \pi_{1}, \ldots, \pi_{n+m+1-M}$ with $\sum_{k=0}^{n+m+1-M} \pi_{k}=1$ such that

$$
\begin{equation*}
\Delta(\lambda)=\tau^{M-n}\left(\lambda-\lambda_{0}\right)^{M} \sum_{k=0}^{n+m+1-M} \pi_{k} \frac{(M-m+k-1)!}{(2 M-m-n+k-1)!} \Phi\left(M-n, 2 M-m-n+k,-\tau\left(\lambda-\lambda_{0}\right)\right) . \tag{38}
\end{equation*}
$$

Proof. By Theorem 3.1, $\lambda_{0}$ is a root of multiplicity at least $M$ of $\Delta$ if and only if there exists a polynomial $p$ of degree $n+m+1-M$ with $p(0)=1$ and such that (33) holds. Such a polynomial can be written in a unique manner as

$$
\begin{equation*}
p(t)=\sum_{k=0}^{n+m+1-M} \pi_{k}(1-t)^{k} \tag{39}
\end{equation*}
$$

for some real numbers $\pi_{0}, \ldots, \pi_{n+m+1-M}$, and the condition $p(0)=1$ is satisfied if and only if

$$
\sum_{k=0}^{n+m+1-M} \pi_{k}=1
$$

Hence, (33) can be rewritten as

$$
\Delta(\lambda)=\frac{\tau^{M-n}\left(\lambda-\lambda_{0}\right)^{M}}{(M-n-1)!} \sum_{k=0}^{n+m+1-M} \pi_{k} \int_{0}^{1} t^{M-n-1}(1-t)^{k+M-m-1} \mathrm{e}^{-t \tau\left(\lambda-\lambda_{0}\right)} \mathrm{d} t
$$

The equivalence between this formula and (38) follows from (7), yielding the conclusion.

Remark 3. It can be observed that the polynomial $p$ involved in the kernel defining $\Delta$ and given in (33) can be alternatively written as:

$$
\begin{equation*}
p(t)=\sum_{k=0}^{n+m+1-M} \alpha_{k} t^{k} \tag{40}
\end{equation*}
$$

with

$$
\begin{equation*}
\alpha_{k}=(-1)^{k} \sum_{\ell=k}^{n+m+1-M}\binom{k}{\ell} \pi_{\ell} \quad \text { for } \quad k \in\{0, \ldots, n+m+1-M\} \tag{41}
\end{equation*}
$$

Furthermore, using the first equality from (36), one easily recovers the expression of $a_{k}$ for $k \in\{0, \ldots, n-1\}$ as a function of $\alpha_{\ell}$ for $\ell \in\{0, \ldots, n+m+1-M\}$ or equivalently, by using (39), a function of $\pi_{\ell}$ for $\ell \in\{0, \ldots, n+m+1-M\}$.

### 3.2 Elimination-produced function

In concrete control problems, typically, some of the coefficients of the polynomials $P_{0}$ and $P_{\tau}$ from (4) are fixed, corresponding to parameters coming from the physical modeling of the system, while other coefficients can be freely chosen, corresponding to parameters coming, for instance, from an implemented linear feedback controller (see [14] for some illustrations of such applications). In this section, we discuss how one can compute the "free" coefficients of the polynomials (4) in order to impose that a given real number $\lambda_{0}$ is a root of multiplicity at least $M$ of the quasipolynomial $\Delta$ from (3).

We will assume in this section that $n$ and $m$ are nonnegative integers with $n \geq m, M \in[n+1, n+m+1]$ is an integer representing the desired multiplicity of the root $\lambda_{0}$, and that $M-1$ coefficients of the polynomials $P_{0}$ and $P_{\tau}$ from (4) can be freely chosen, the other $n+m+2-M$ coefficients being fixed. In this configuration, which appears in several practical applications (see, e.g., [67], as well as the examples from Section 4 below), the $M-1$ degrees of freedom in the coefficients of the system are typically not enough to ensure the existence of a root of multiplicity $M$ by themselves, and an additional constraint on the root $\lambda_{0}$ and the delay $\tau$ must be imposed. This constraint can be formulated as a certain meromorphic function on $\left(\lambda_{0}, \tau\right)$, depending on the known coefficients of $P_{0}$ and $P_{\tau}$, being equal to 0 . This meromorphic function will be called eliminationproduced function, and the main goal of this section is to describe it, first in a general setting, and then in the particular case where the known coefficients are the coefficients of the monomials of highest degree in $P_{0}$. For that purpose, we start with some appropriate notations and definitions.

Let $n$ and $m$ be nonnegative integers with $n \geq m, D=n+m+1, M \in[n+1, n+m+1]$ be an integer, $I \subset\{1, \ldots, D\}$ be a set of cardinality $M-1$, and $\bar{I}=\{1, \ldots, D\} \backslash I$.

We define the holomorphic functions

$$
\begin{array}{lll}
\ell: \mathbb{C}^{2} \rightarrow \mathcal{M}_{1, D}(\mathbb{C}), & \ell_{I}: \mathbb{C}^{2} \rightarrow \mathcal{M}_{1, M-1}(\mathbb{C}), & \ell_{\bar{I}}: \mathbb{C}^{2} \rightarrow \mathcal{M}_{1, D-M+1}(\mathbb{C}), \\
A: \mathbb{C}^{2} \rightarrow \mathcal{M}_{M-1, D}(\mathbb{C}), & A_{I}: \mathbb{C}^{2} \rightarrow \mathcal{M}_{M-1, M-1}(\mathbb{C}), & A_{\bar{I}}: \mathbb{C}^{2} \rightarrow \mathcal{M}_{M-1, D-M+1}(\mathbb{C}), \\
b: \mathbb{C} \rightarrow \mathcal{M}_{M-1,1}(\mathbb{C}), & &
\end{array}
$$

and the real number $b_{M}$ as follows:
For $\left(\lambda_{0}, \tau\right) \in \mathbb{C}^{2}$, we set

$$
\ell\left(\lambda_{0}, \tau\right)=\left(\begin{array}{llllllll}
\lambda_{0}^{n-1} & \cdots & \lambda_{0} & 1 & \lambda_{0}^{m} \mathrm{e}^{-\lambda_{0} \tau} & \cdots & \lambda_{0} \mathrm{e}^{-\lambda_{0} \tau} & \mathrm{e}^{-\lambda_{0} \tau} \tag{42}
\end{array}\right)
$$

$\ell_{I}\left(\lambda_{0}, \tau\right)$ (respectively, $\ell_{\bar{I}}\left(\lambda_{0}, \tau\right)$ ) is the row matrix whose entries are the columns of $\frac{\partial^{M-1}}{\partial \lambda_{0}^{M-1}} \ell\left(\lambda_{0}, \tau\right)$ whose indices belong to $I$ (respectively, to $\bar{I}$ ),

$$
A\left(\lambda_{0}, \tau\right)=\left(\begin{array}{c}
\ell\left(\lambda_{0}, \tau\right)  \tag{43}\\
\frac{\partial \ell}{\partial \lambda_{0}}\left(\lambda_{0}, \tau\right) \\
\vdots \\
\frac{\partial^{M-2} \ell}{\partial \lambda_{0}^{M-2}}\left(\lambda_{0}, \tau\right)
\end{array}\right)
$$

$A_{I}\left(\lambda_{0}, \tau\right)$ (respectively, $A_{\bar{I}}\left(\lambda_{0}, \tau\right)$ ) is the matrix whose columns are the columns of $A\left(\lambda_{0}, \tau\right)$ whose indices belong to $I$ (respectively, to $\bar{I}$ ), $b\left(\lambda_{0}\right)$ is the column matrix with $M-1$ rows whose first entry is $-\lambda_{0}^{n}$ and such that its $k$-th entry $b_{k}\left(\lambda_{0}\right)$, for an integer $k \in[2, M-1]$, satisfies $b_{k}\left(\lambda_{0}\right)=\frac{d}{d \lambda_{0}} b_{k-1}\left(\lambda_{0}\right)$. We also set

$$
b_{M}= \begin{cases}-n! & \text { if } \quad M=n+1  \tag{44}\\ 0 & \text { otherwise }\end{cases}
$$

The set $I$ introduced above represents the indices of the coefficients of the polynomials $P_{0}$ and $P_{\tau}$ from (4) that are assumed to be free, when arranged in the column vector

$$
X=\left(\begin{array}{llllllll}
a_{n-1} & \cdots & a_{1} & a_{0} & b_{m} & \cdots & b_{1} & b_{0} \tag{45}
\end{array}\right)^{T}
$$

and $\bar{I}$ represents the indices of the known coefficients of $P_{0}$ and $P_{\tau}$ in this same vector. Note that the order of the coefficients in the vector $X$ corresponds to the order of the entries of the row matrix $\ell\left(\lambda_{0}, \tau\right)$, so that we have the equality $\Delta\left(\lambda_{0}\right)=\lambda_{0}^{n}+\ell\left(\lambda_{0}, \tau\right) X$.

Definition 3.1. With the notations above, if $\operatorname{det} A_{I}(\cdot, \cdot)$ is not the zero function, for $\left(\lambda_{0}, \tau\right) \in \mathbb{C} \times \mathbb{R}_{+}$such that $\operatorname{det} A_{I}\left(\lambda_{0}, \tau\right) \neq 0$ and $\gamma \in \mathbb{R}^{D-M+1}$, we define $\mathbb{P}\left(\lambda_{0}, \tau ; \gamma\right) \in \mathbb{C}$ by

$$
\begin{equation*}
\mathbb{P}\left(\lambda_{0}, \tau ; \gamma\right)=\ell_{\bar{I}}\left(\lambda_{0}, \tau\right) \gamma+\ell_{I}\left(\lambda_{0}, \tau\right) A_{I}\left(\lambda_{0}, \tau\right)^{-1}\left(b\left(\lambda_{0}\right)-A_{\bar{I}}\left(\lambda_{0}, \tau\right) \gamma\right)-b_{M} \tag{46}
\end{equation*}
$$

The function $\mathbb{P}$ is called elimination-produced function.
The vector $\gamma$ in Definition 3.1 represents the numerical values of the known coefficients of $P_{0}$ and $P_{\tau}$ from (4): it corresponds to taking, in the above vector $X$, only the entries corresponding to indices in $\bar{I}$.

Remark 4. Note that $\operatorname{det} A_{I}\left(\lambda_{0}, \tau\right)$ is a quasipolynomial in the variable $\lambda_{0}$, depending only on $\tau, n, m, M$, and $I$. In particular, when $\operatorname{det} A_{I}$ is not identically zero, for every $\gamma \in \mathbb{R}^{D-M+1}$, the function $\mathbb{P}(\cdot, \cdot ; \gamma)$ is meromorphic, and can be expressed as a fraction of quasipolynomials, its denominator being $\operatorname{det} A_{I}\left(\lambda_{0}, \tau\right)$.

Example 3.1. To illustrate Definition 3.1 and notations above, consider the case $n=2$ and $m=1$, i.e.,

$$
\Delta(\lambda)=\lambda^{2}+a_{1} \lambda+a_{0}+\mathrm{e}^{-\lambda \tau}\left(b_{1} \lambda+b_{0}\right),
$$

in which case $D=n+m+1=4$. We are interested here in roots of multiplicity $M=3$. We assume that $a_{1}$ and $b_{0}$ are known, and $a_{0}$ and $b_{1}$ are free. In this case, we have $I=\{2,3\}$, which correspond to the indices of $a_{0}$ and $b_{1}$ in the vector $\left(a_{1}, a_{0}, b_{1}, b_{0}\right)$, and thus $\bar{I}=\{1,4\}$. In this case, we compute

$$
\begin{aligned}
& \ell\left(\lambda_{0}, \tau\right)=\left(\begin{array}{llll}
\lambda_{0} & 1 & \lambda_{0} \mathrm{e}^{-\lambda_{0} \tau} & \mathrm{e}^{-\lambda_{0} \tau}
\end{array}\right), \\
& \ell_{I}\left(\lambda_{0}, \tau\right)=\left(\begin{array}{ll}
0 & \left(\tau^{2} \lambda_{0}-2 \tau\right) \mathrm{e}^{-\lambda_{0} \tau}
\end{array}\right), \\
& \ell_{\bar{I}}\left(\lambda_{0}, \tau\right)=\left(\begin{array}{ll}
0 & \tau^{2} \mathrm{e}^{-\lambda_{0} \tau}
\end{array}\right), \\
& A\left(\lambda_{0}, \tau\right)=\left(\begin{array}{cccc}
\lambda_{0} & 1 & \lambda_{0} \mathrm{e}^{-\lambda_{0} \tau} & \mathrm{e}^{-\lambda_{0} \tau} \\
1 & 0 & \left(1-\lambda_{0} \tau\right) \mathrm{e}^{-\lambda_{0} \tau} & -\tau \mathrm{e}^{-\lambda_{0} \tau}
\end{array}\right), \\
& A_{I}\left(\lambda_{0}, \tau\right)=\left(\begin{array}{cc}
1 & \lambda_{0} \mathrm{e}^{-\lambda_{0} \tau} \\
0 & \left(1-\lambda_{0} \tau\right) \mathrm{e}^{-\lambda_{0} \tau}
\end{array}\right), \\
& A_{\bar{I}}\left(\lambda_{0}, \tau\right)=\left(\begin{array}{cc}
\lambda_{0} & \mathrm{e}^{-\lambda_{0} \tau} \\
1 & -\tau \mathrm{e}^{-\lambda_{0} \tau}
\end{array}\right), \\
& b\left(\lambda_{0}\right)=\binom{-\lambda_{0}^{2}}{-2 \lambda_{0}}, \\
& b_{M}=-2 \text {. }
\end{aligned}
$$

In particular, we have

$$
\operatorname{det} A_{I}\left(\lambda_{0}, \tau\right)=\left(1-\lambda_{0} \tau\right) \mathrm{e}^{-\lambda_{0} \tau}
$$

which is not identically zero. Thus, letting $\gamma=\left(a_{1}^{*}, b_{0}^{*}\right)$ be the vector containing the numerical values of the fixed parameters $a_{1}$ and $b_{0}$, the elimination-produced function $\mathbb{P}$ is given by

$$
\mathbb{P}\left(\lambda_{0}, \tau ;\left(a_{1}^{*}, b_{0}^{*}\right)\right)=\frac{a_{1}^{*} \lambda_{0} \tau^{2}-2 a_{1}^{*} \tau+b_{0}^{*} \tau^{2} \mathrm{e}^{-\lambda_{0} \tau}+2 \lambda_{0}^{2} \tau^{2}-2 \lambda_{0} \tau-2}{\lambda_{0} \tau-1}
$$

The main interest of the notation we introduced above, and in particular of the elimination-produced function from Definition 3.1, is that it allows one to express conditions for a real number $\lambda_{0}$ to be a root of multiplicity $M$ of $\Delta$ in terms of the free coefficients of $P_{0}$ and $P_{\tau}$, as expressed in the following theorem.

Theorem 3.2. Let $n, m$ be nonnegative integers with $n \geq m, \tau>0, M \in[n+1, n+m+1]$ be an integer, and consider the quasipolynomial $\Delta$ from (3)-(4).

Let $I \subset\{1, \ldots, D\}$ be a set of cardinality $M-1$ and $\bar{I}=\{1, \ldots, D\} \backslash I$. Let $X$ be the parameter vector from (45) and denote by $X_{I}$ and $X_{\bar{I}}$ the vectors obtained by keeping in $X$ only the entries with indices in $I$ and $\bar{I}$, respectively. Assume that the values of the parameters in $X_{\bar{I}}$ are known and denote by $\gamma$ the vector with those values. Consider also the functions $\ell_{I}, \ell_{\bar{I}}, A_{I}, A_{\bar{I}}, b, \mathbb{P}$, and the real number $b_{M}$ defined in (42)-(44).

Let $\lambda_{0} \in \mathbb{R}$ be such that $\operatorname{det} A_{I}\left(\lambda_{0}, \tau\right) \neq 0$. Then $\lambda_{0}$ is a root of multiplicity at least $M$ of $\Delta$ if and only if

$$
\begin{equation*}
X_{I}=A_{I}\left(\lambda_{0}, \tau\right)^{-1}\left(b\left(\lambda_{0}\right)-A_{\bar{I}}\left(\lambda_{0}, \tau\right) \gamma\right) \quad \text { and } \quad \mathbb{P}\left(\lambda_{0}, \tau ; \gamma\right)=0 \tag{47}
\end{equation*}
$$

Proof. The number $\lambda_{0}$ is a root of multiplicity at least $M$ of $\Delta$ if and only if $\Delta^{(k)}\left(\lambda_{0}\right)=0$ for every $k \in$ $\{0, \ldots, M-1\}$, and, recalling that $\Delta(\lambda)=\lambda^{n}+\ell(\lambda, \tau) X$, these $M$ equations can be rewritten as

$$
\left\{\begin{aligned}
A\left(\lambda_{0}, \tau\right) X & =b\left(\lambda_{0}\right), \\
\frac{\partial^{M-1} \ell}{\partial \lambda_{0}^{M-1}} \ell\left(\lambda_{0}, \tau\right) X & =b_{M},
\end{aligned}\right.
$$

where $\ell$ and $A$ are defined in (42) and (43). Splitting $X$ into $X_{I}$ and $X_{\bar{I}}$ and noticing that $X_{\bar{I}}=\gamma$, the above system is equivalent to

$$
\left\{\begin{aligned}
A_{I}\left(\lambda_{0}, \tau\right) X_{I}+A_{\bar{I}}\left(\lambda_{0}, \tau\right) \gamma & =b\left(\lambda_{0}\right) \\
\ell_{I}\left(\lambda_{0}, \tau\right) X_{I}+\ell_{\bar{I}}\left(\lambda_{0}, \tau\right) \gamma & =b_{M}
\end{aligned}\right.
$$

and, using the assumption $\operatorname{det} A_{I}\left(\lambda_{0}, \tau\right) \neq 0$, we obtain that the above system is equivalent to (47).
Remark 5. In the above framework, we have considered that the values of some of the coefficients of the quasipolynomial $\Delta$ from (3)-(4) are known. Our approach can also be generalized to the case where, instead of knowing values of the coefficients, we have other information on them, such as the fact that they satisfy some linear dependence relation, written under the form $C X=d$ for some matrix $C \in \mathcal{M}_{D-M+1, D}(\mathbb{C})$ of full rank and some $d \in \mathbb{R}^{D-M+1}$. Indeed, in this case, we decompose the space $\mathbb{R}^{D}$ into the direct sum $\operatorname{Ker} C \oplus(\operatorname{Ker} C)^{\perp}$, and we replace the decomposition into indices in $I$ and $\bar{I}$ done in the construction above by projections into $\operatorname{Ker} C$ and $(\operatorname{Ker} C)^{\perp}$.

### 3.3 A control-oriented setting

In Section $3.2, I$ is an arbitrary subset of $\{1, \ldots, D\}$ with $M-1$ elements, meaning that we can consider any choice of free parameters of the quasipolynomial $\Delta$ from (3)-(4). With the aim of obtaining more precise results in a particular situation, we now consider the case where $I=\{D-M+2, \ldots, D\}$, and thus $\bar{I}=$ $\{1, \ldots, D-M+1\}$, meaning that the coefficients of highest degree of $P_{0}$ are all known, and the $M-1$ free parameters are the coefficients of $P_{\tau}$ and those of lowest degree of $P_{0}$ (except in the particular case $n=m$ and $M=n+1$, in which case the coefficient $b_{m}$ of $P_{\tau}$ is also assumed to be known).

Our first result in this setting is the following characterization of the elimination-produced function.

Proposition 3.2. Let $n$, $m$ be nonnegative integers with $n \geq m, \tau>0, M \in[n+1, n+m+1]$ be an integer satisfying $M \geq m+2$, and consider the quasipolynomial $\Delta$ from (3)-(4). Let $\mathbb{P}$ be the elimination-produced function in the case $I=\{D-M+2, \ldots, D\}$. Then

$$
\begin{equation*}
\mathbb{P}\left(\lambda_{0}, \tau ; \gamma\right)=\sum_{k=0}^{m+1}\binom{m+1}{k} \tau^{k} P_{F, M}^{(M-1-k)}\left(\lambda_{0}\right) \tag{48}
\end{equation*}
$$

where

$$
\begin{equation*}
P_{F, M}(\lambda)=\lambda^{n}+\sum_{k=1}^{D-M+1} a_{n-k} \lambda^{n-k} \tag{49}
\end{equation*}
$$

Proof. Note that, with the notations of Section 3.2, we have

$$
b\left(\lambda_{0}\right)-A_{\bar{I}}\left(\lambda_{0}, \tau\right) \gamma=\left(\begin{array}{c}
-P_{F, M}\left(\lambda_{0}\right)  \tag{50}\\
-P_{F, M}^{\prime}\left(\lambda_{0}\right) \\
\vdots \\
-P_{F, M}^{(M-2)}\left(\lambda_{0}\right)
\end{array}\right)
$$

and

$$
\begin{equation*}
\ell_{\bar{I}}\left(\lambda_{0}, \tau\right) \gamma-b_{M}=P_{F, M}^{(M-1)}\left(\lambda_{0}\right) \tag{51}
\end{equation*}
$$

Let

$$
\tilde{\ell}\left(\lambda_{0}, \tau\right)=\left(\begin{array}{llllllll}
\lambda_{0}^{M-m-3} & \cdots & \lambda_{0} & 1 & \lambda_{0}^{m} \mathrm{e}^{-\lambda_{0} \tau} & \cdots & \lambda_{0} \mathrm{e}^{-\lambda_{0} \tau} & \mathrm{e}^{-\lambda_{0} \tau}
\end{array}\right)
$$

and note that

$$
A_{I}\left(\lambda_{0}, \tau\right)=\left(\begin{array}{c}
\tilde{\ell}\left(\lambda_{0}, \tau\right) \\
\frac{\partial \widetilde{\ell}}{\partial \lambda_{0}}\left(\lambda_{0}, \tau\right) \\
\vdots \\
\frac{\partial^{M-2} \tilde{\ell}}{\partial \lambda_{0}^{M-2}}\left(\lambda_{0}, \tau\right)
\end{array}\right)
$$

The matrix $A_{I}\left(\lambda_{0}, \tau\right)$ admits the block decomposition

$$
A_{I}\left(\lambda_{0}, \tau\right)=\left(\begin{array}{cc}
A_{I, 0}\left(\lambda_{0}\right) & A_{I, 1}\left(\lambda_{0}, \tau\right)  \tag{52}\\
0 & A_{I, 2}\left(\lambda_{0}, \tau\right)
\end{array}\right)
$$

where

$$
\begin{aligned}
A_{I, 0}\left(\lambda_{0}\right) & =\left(\begin{array}{cccc}
\lambda_{0}^{M-m-3} & \cdots & \lambda_{0} & 1 \\
(M-m-3) \lambda_{0}^{M-m-4} & \cdots & 1 & 0 \\
\vdots & . & \vdots & \vdots \\
(M-m-3)! & \cdots & 0 & 0
\end{array}\right) \in \mathcal{M}_{M-m-2}(\mathbb{R}), \\
A_{I, 1}\left(\lambda_{0}, \tau\right) & =\left(\begin{array}{c}
\widehat{\ell}\left(\lambda_{0}, \tau\right) \\
\frac{\partial \widehat{\ell}}{\partial \lambda_{0}}\left(\lambda_{0}, \tau\right) \\
\vdots \\
\frac{\partial^{M-m-3} \widehat{\ell}}{\partial \lambda_{0}^{M-m-3}}\left(\lambda_{0}, \tau\right)
\end{array}\right) \in \mathcal{M}_{M-m-2, m+1}(\mathbb{R}), \\
A_{I, 2}\left(\lambda_{0}, \tau\right) & =\left(\begin{array}{c}
\frac{\partial^{M-m-2} \widehat{\ell}}{\partial \lambda_{0}^{M-m-2}}\left(\lambda_{0}, \tau\right) \\
\vdots \\
\frac{\partial^{M-2} \widehat{\ell}}{\partial \lambda_{0}^{M-2}}\left(\lambda_{0}, \tau\right)
\end{array}\right) \in \mathcal{M}_{m+1, m+1}(\mathbb{R})
\end{aligned}
$$

$$
\widehat{\ell}\left(\lambda_{0}, \tau\right)=\left(\begin{array}{lll}
\lambda_{0}^{m} \mathrm{e}^{-\lambda_{0} \tau} & \cdots & \lambda_{0} \mathrm{e}^{-\lambda_{0} \tau}
\end{array} \mathrm{e}^{-\lambda_{0} \tau}\right) \in \mathcal{M}_{1, m+1}(\mathbb{R}),
$$

Clearly, $A_{I, 0}\left(\lambda_{0}\right)$ is invertible. Set $\rho\left(\lambda_{0}, \tau\right)=\mathrm{e}^{\lambda_{0} \tau} \frac{\partial^{M-m-2} \widehat{\ell}}{\partial \lambda_{0}^{M-m-2}}\left(\lambda_{0}, \tau\right)$ and note that

$$
\begin{equation*}
\rho\left(\lambda_{0}, \tau\right)=\left(\mathfrak{q}_{m}\left(\lambda_{0}, \tau\right) \quad \cdots \quad \mathfrak{q}_{1}\left(\lambda_{0}, \tau\right) \quad \mathfrak{q}_{0}\left(\lambda_{0}, \tau\right)\right), \tag{53}
\end{equation*}
$$

where, for $j \in\{0, \ldots, m\}$, the function $\lambda_{0} \mapsto \mathfrak{q}_{j}\left(\lambda_{0}, \tau\right)$ is a polynomial of degree $j$, with coefficients depending on $\tau$, and its term of highest degree is $(-\tau)^{M-m-2} \lambda_{0}^{j}$. Thus

$$
\left(\begin{array}{c}
\rho\left(\lambda_{0}, \tau\right) \\
\frac{\partial \rho}{\partial \lambda_{0}}\left(\lambda_{0}, \tau\right) \\
\vdots \\
\frac{\partial^{m} \rho}{\partial \lambda_{0}^{m}}\left(\lambda_{0}, \tau\right)
\end{array}\right)=\left(\begin{array}{ccccc}
\mathfrak{q}_{m}\left(\lambda_{0}, \tau\right) & \cdots & \mathfrak{q}_{2}\left(\lambda_{0}, \tau\right) & \mathfrak{q}_{1}\left(\lambda_{0}, \tau\right) & (-\tau)^{M-m-2} \\
\frac{\partial \mathfrak{q}_{m}}{\partial \lambda_{0}}\left(\lambda_{0}, \tau\right) & \cdots & \frac{\partial \mathfrak{q}_{2}}{\partial \lambda_{0}}\left(\lambda_{0}, \tau\right) & (-\tau)^{M-m-2} & 0 \\
\frac{\partial^{2} \mathfrak{q}_{m}}{\partial \lambda_{0}^{2}}\left(\lambda_{0}, \tau\right) & \cdots & 2(-\tau)^{M-m-2} & 0 & 0 \\
\vdots & . \cdot & \vdots & \vdots & \vdots \\
m!(-\tau)^{M-m-2} & \cdots & 0 & 0 & 0
\end{array}\right)
$$

and, in particular, since $\tau>0$, the above matrix is invertible. On the other hand, taking derivatives directly in the definition of $\rho$, we have

$$
\left(\begin{array}{c}
\rho\left(\lambda_{0}, \tau\right) \\
\frac{\partial \rho}{\partial \lambda_{0}}\left(\lambda_{0}, \tau\right) \\
\vdots \\
\frac{\partial^{m} \rho}{\partial \lambda_{0}^{m}}\left(\lambda_{0}, \tau\right)
\end{array}\right)=e^{\lambda_{0} \tau}\left(\begin{array}{cccccc}
1 & 0 & 0 & \cdots & 0 & 0 \\
\tau & 1 & 0 & \cdots & 0 & 0 \\
\vdots & \ddots & \ddots & \ddots & \ddots & \vdots \\
\binom{k}{k} \tau^{k} & \binom{k}{k-1} \tau^{k-1} & \ddots & 1 & \ddots & 0 \\
\vdots & \vdots & \ddots & \ddots & \ddots & \vdots \\
\binom{m}{m} \tau^{m} & \binom{m}{m-1} \tau^{m-1} & \cdots & \binom{m}{2} \tau^{2} & m \tau & 1
\end{array}\right) A_{I, 2}(\rho, \tau)
$$

Hence, $A_{I, 2}(\rho, \tau)$ and, thanks to the block decomposition (52), $A_{I}\left(\lambda_{0}, \tau\right)$ is invertible.
Thanks to (53), we have $\frac{\partial^{m+1} \rho}{\partial \lambda_{0}^{m+1}}\left(\lambda_{0}, \tau\right)=0$ and, using the definition of $\rho$, this implies that

$$
\sum_{k=0}^{m+1}\binom{m+1}{k} \tau^{k} \frac{\partial^{M-k-1} \widehat{\ell}}{\partial \lambda_{0}^{M-k-1}}\left(\lambda_{0}, \tau\right)=0
$$

which can be rewritten in matrix form as

$$
\left(\binom{m+1}{m+1} \tau^{m+1} \quad\binom{m+1}{m} \tau^{m} \quad \ldots \quad\binom{m+1}{1} \tau\right) A_{I, 2}\left(\lambda_{0}, \tau\right)=-\frac{\partial^{M-1} \widehat{\ell}}{\partial \lambda_{0}^{M-1}}\left(\lambda_{0}, \tau\right)
$$

Hence

$$
\left(\begin{array}{llll}
0 & \cdots & 0 & \binom{m+1}{m+1} \tau^{m+1}
\end{array}\binom{m+1}{m} \tau^{m} \quad \cdots \quad\binom{m+1}{1} \tau\right) A_{I}\left(\lambda_{0}, \tau\right)=-\left(\begin{array}{cc}
0 & \frac{\partial^{M-1} \widehat{\ell}}{\partial \lambda_{0}^{M-1}}\left(\lambda_{0}, \tau\right)
\end{array}\right)=-\ell_{I}\left(\lambda_{0}, \tau\right)
$$

so that

$$
\ell_{I}\left(\lambda_{0}, \tau\right) A_{I}\left(\lambda_{0}, \tau\right)^{-1}=-\left(\begin{array}{lllll}
0 & \cdots & 0 & \binom{m+1}{m+1} \tau^{m+1} & \cdots \tag{54}
\end{array}\binom{m+1}{1} \tau\right) .
$$

Inserting (50), (51), and (54) into (46), we finally deduce (48).
Remark 6. Note that the elimination-produced function from (48) is a polynomial in $\left(\lambda_{0}, \tau\right)$. We refer to this function as the elimination-produced polynomial.

Remark 7. Proposition 3.2 requires the additional assumption that $M \geq m+2$. This is always satisfied if $n>m$, since $M \geq n+1$, and hence the only case not covered by Proposition 3.2 is the case of neutral systems
with the smallest over-order multiplicity, $M=n+1$. In this case, the conclusion of Remark 6 does not hold. Indeed, consider the case $n=m=1$, i.e.,

$$
\Delta(\lambda)=\lambda+a_{0}+e^{-\lambda \tau}\left(b_{1} \lambda+b_{0}\right)
$$

in which case $D=3$. The lowest over-order multiplicity is $M=2$ and, assuming $a_{0}$ and $b_{1}$ to be known and $b_{0}$ to be free (i.e., $I=\{3\}$ and $\bar{I}=\{1,2\}$ ), straightforward computations from its definition show that the elimination-produced function is

$$
\mathbb{P}\left(\lambda_{0}, \tau ; \gamma\right)=1+\tau \lambda_{0}+\tau a_{0}+b_{1} e^{-\lambda_{0} \tau}
$$

which is not a polynomial in $\left(\lambda_{0}, \tau\right)$.
An interesting fact about the elimination-produced polynomial from Proposition 3.2 is that the eliminationproduced polynomial for a root of multiplicity $M+1$ can be obtained by derivating the elimination-produced polynomail for a root of multiplicity $M$, as detailed in the next result.

Proposition 3.3. Let $n, m$ be nonnegative integers with $n \geq m, \tau>0, M \in[n+1, n+m]$ be an integer satisfying $M \geq m+2$, and consider the quasipolynomial $\Delta$ from (3)-(4). Let $\mathbb{P}_{M}$ (respectively, $\mathbb{P}_{M+1}$ ) be the elimination-produced polynomial from Proposition 3.2 for a root of multiplicity at least $M$ (respectively, at least $M+1$ ). Then

$$
\mathbb{P}_{M+1}\left(\lambda_{0}, \tau ; \gamma\right)=\frac{\partial \mathbb{P}_{M}}{\partial \lambda_{0}}\left(\lambda_{0}, \tau ; \gamma\right)
$$

Proof. Let $P_{F, M}$ and $P_{F, M+1}$ be defined as in (49). Then

$$
P_{F, M}(\lambda)=P_{F, M+1}(\lambda)+a_{M-m-2} \lambda^{M-m-2}
$$

Hence, for every integer $k \in[0, m+1]$, we have $M-m-2<M-k$, and thus

$$
P_{F, M}^{(M-k)}(\lambda)=P_{F, M+1}^{(M-k)}(\lambda) .
$$

It then follows from (48) that

$$
\begin{aligned}
\mathbb{P}_{M+1}\left(\lambda_{0}, \tau ; \gamma\right) & =\sum_{k=0}^{m+1}\binom{m+1}{k} \tau^{k} P_{F, M+1}^{(M-k)}\left(\lambda_{0}\right) \\
& =\sum_{k=0}^{m+1}\binom{m+1}{k} \tau^{k} P_{F, M}^{(M-k)}\left(\lambda_{0}\right)=\frac{\partial \mathbb{P}_{M}}{\partial \lambda_{0}}\left(\lambda_{0}, \tau ; \gamma\right)
\end{aligned}
$$

We conclude this section by the following result, which provides an explicit expression of the eliminationproduced polynomial $\mathbb{P}$ from (48) in terms of the known coefficients $a_{M-m-2}, \ldots, a_{n-1}$ of $\Delta$.

Proposition 3.4. Let $n$, $m$ be nonnegative integers with $n \geq m, \tau>0, M \in[n+1, n+m+1]$ be an integer satisfying $M \geq m+2$, and consider the quasipolynomial $\Delta$ from (3)-(4). Let $\mathbb{P}$ be the elimination-produced function from Definition 3.1 in the case $I=\{D-M+2, \ldots, D\}$. Then

$$
\begin{equation*}
\mathbb{P}\left(\lambda_{0}, \tau ; \gamma\right)=\sum_{j=0}^{D-M+1} \sum_{\ell=j+M-m-2}^{n}\binom{m+1}{M-\ell-1+j} \frac{\ell!}{j!} \tau^{M-\ell-1+j} a_{\ell} \lambda_{0}^{j} \tag{55}
\end{equation*}
$$

with the convention $a_{n}=1$.
The proof of the above proposition follows by inserting (49) into (48) and standard changes of summation variables in order to group the terms in powers of $\lambda_{0}$.

### 3.4 Dominance of multiple roots and its consequences for stability

Note that (38) factorizes $\Delta$ in terms of a linear combination of $D-M+1$ Kummer functions with real coefficients. However, in the following Proposition an equivalent combination of two Kummer functions with coefficients which are rational functions in the terms of the complex variable $\lambda$ will be considered.

Proposition 3.5. Consider the quasipolynomial $\Delta$ from (3)-(4) with $m \leq n$. The real number $\lambda_{0}$ is a root of multiplicity at least $n+1 \leq M \leq n+m+1$ of $\Delta$ if, and only if, $\mathbb{P}\left(\lambda_{0}, \tau\right)=0$ and

$$
\begin{equation*}
\Delta(\lambda)=\frac{\tau^{M-n}\left(\lambda-\lambda_{0}\right)^{M}}{(M-n-1)!}\left(\beta(\lambda) \Phi\left(0,1,-\tau\left(\lambda-\lambda_{0}\right)\right)+\gamma(\lambda) \Phi\left(1,1,-\tau\left(\lambda-\lambda_{0}\right)\right)\right) \tag{56}
\end{equation*}
$$

where

$$
\begin{align*}
& \beta(\lambda)=\frac{(M-n-1)!P_{0}(\lambda)}{\tau^{M-n}\left(\lambda-\lambda_{0}\right)^{M}}  \tag{57}\\
& \gamma(\lambda)=\frac{(M-n-1)!P_{\tau}(\lambda)}{\tau^{M-n}\left(\lambda-\lambda_{0}\right)^{M}} \tag{58}
\end{align*}
$$

and $P_{0}$ and $P_{\tau}$ are given in (36)-(37).
Since for every $a \in \mathbb{C}$ with $-a \notin \mathbb{N}$, we have $\Phi(a, a, z)=\Phi(1,1, z)=\mathrm{e}^{z}$ and $\Phi(0,1, z)=1$, thus an immediate representation of the quasipolynomial $\Delta$ is given by (56).

Beyond the standard contiguous relation (see for instance [62]), to the best of the authors' knowledge, there does not exist any result describing the distribution of the nonasymptotic zeros of linear combinations of Kummer functions. The next lemma provides a partial step towards that goal, by providing a non-autonomous second-order differential equation admitting a given linear combination of Kummer functions as a solution.

Lemma 4. Let $\widetilde{\beta}$ and $\widetilde{\gamma}$ be two meromorphic functions. Then, the complex function $F$ defined by

$$
\begin{equation*}
F(z)=\widetilde{\beta}(z) \Phi(0,1, z)+\widetilde{\gamma}(z) \Phi(1,1, z) \tag{59}
\end{equation*}
$$

with $\widetilde{\beta}(z) \widetilde{\gamma}^{\prime}(z)+\widetilde{\gamma}(z)\left(\widetilde{\beta}(z) \tau-\widetilde{\beta}^{\prime}(z)\right) \neq 0$ satisfies the following second-order differential equation

$$
\begin{equation*}
F^{\prime \prime}(z)+Q(z) F^{\prime}(z)+R(z) F(z)=0 \tag{60}
\end{equation*}
$$

where $Q$ and $R$ are given in (61) and (62).

$$
\begin{align*}
& Q(z)=\frac{\left(\widetilde{\beta}^{\prime \prime}(z)+\tau\left(\widetilde{\beta}(z) \tau-2 \widetilde{\beta}^{\prime}(z)\right)\right) \widetilde{\gamma}(z)-\left(\widetilde{\gamma}^{\prime \prime}(z)\right) \widetilde{\beta}(z)}{\widetilde{\beta}(z) \widetilde{\gamma}^{\prime}(z)+\widetilde{\gamma}(z)\left(\widetilde{\beta}(z) \tau-\widetilde{\beta}^{\prime}(z)\right)}  \tag{61}\\
& R(z)=\frac{\left(-\widetilde{\beta}(z) \tau+\widetilde{\beta}^{\prime}(z)\right) \widetilde{\gamma}^{\prime \prime}(z)-\widetilde{\gamma}^{\prime}(z)\left(\widetilde{\beta}^{\prime \prime}(z)+\tau\left(\widetilde{\beta}(z) \tau-2 \widetilde{\beta}^{\prime}(z)\right)\right)}{\widetilde{\beta}(z) \widetilde{\gamma}^{\prime}(z)+\widetilde{\gamma}(z)\left(\widetilde{\beta}(z) \tau-\widetilde{\beta}^{\prime}(z)\right)} . \tag{62}
\end{align*}
$$

In what follows, we shall refer to functions $F$ of the form (59) as Kummer-type functions. Similarly to what was done in Lemma 3, one can also define a Whittaker-type function $W$ from the Kummer-type function $F$ defined in (59) by the same formula (15), where $\mathcal{Q}$ is a primitive of $\frac{Q}{2}$ for $Q$ given by (61), and $W$ satisfies (16) with $G$ defined from $Q$ and $R$ from (61)-(62) using the same expression (17) as in Lemma 3.

### 3.5 MID Validity for Over-order Multiplicities

Now, we shall use the results of Section 2.3 relating quasipolynomials with roots of over-order multiplicity and Kummer and Whittaker functions in order to provide sufficient conditions under which the MID property is valid for characteristic roots of multiplicity at least $n+1 \leq M \leq m+n+1$ of $\Delta$.

Theorem 3.3. Consider the quasipolynomial $\Delta$ from (3)-(4) with $m \leq n$, and assume that $\Delta$ admits a real root $\lambda_{0}$ of multiplicity at least $n+1 \leq M \leq m+n+1$. Let $\beta$ and $\gamma$ be the meromorphic functions defined in (57) and (58), respectively, and define the meromorphic functions $\widetilde{\beta}$ and $\widetilde{\gamma}$ by

$$
\begin{equation*}
\widetilde{\beta}(z)=\beta\left(\lambda_{0}-\frac{z}{\tau}\right), \quad \widetilde{\gamma}(z)=\gamma\left(\lambda_{0}-\frac{z}{\tau}\right) . \tag{63}
\end{equation*}
$$

Let $F, Q, R$, and $G$ be defined by (59), (61), (62), and (17), respectively. Assume that, for every $t \in(0,1)$ and every root $z$ of $F$ in $\mathbb{C}_{-}$, we have $\Re[z G(t z)] \geq 0$. Then, $\lambda_{0}$ is a dominant root of $\Delta$, i.e., $\lambda_{0}$ satisfies the MID property.

A result similar to Theorem 3.3 was already shown in [18, Theorem 10] for the case of roots of multiplicity $n+m$. The proof of the former can be obtained by an easy adaptation of that of the latter, and we detail this argument here for the sake of completeness.

Proof. We deduce from Proposition 3.5 and Lemma 4 that

$$
\begin{equation*}
\Delta(\lambda)=\frac{\tau^{M-n}\left(\lambda-\lambda_{0}\right)^{M}}{(M-n-1)!} F\left(-\tau\left(\lambda-\lambda_{0}\right)\right) . \tag{64}
\end{equation*}
$$

Since our objective is to investigate zeros of $\Delta$ which are different from $\lambda_{0}$, then we focus on its second factor $F$. In particular, the result is thereby proved if we show that all roots of the Kummer-type function $F$ have nonnegative real part.

To do so, we consider the Whittaker-type function $W(\cdot)$ defined from $F$ as in (15). Note that the differential equation (16) satisfied by $W$ is of the form (18), with $K(z)=1$. As a consequence, one can apply Hille's method to (16). By taking $z_{1}=0$ and $z_{2}$ equal to a root $z_{*}$ of $F(\cdot)$ in (19), we obtain:

$$
\int_{0}^{z_{*}}\left|W^{\prime}(z)\right|^{2} \overline{\mathrm{~d} z}=\int_{0}^{z_{*}}|W(z)|^{2} G(z) \mathrm{d} z
$$

We choose as integration path the line segment from 0 to $z_{*}$. Hence

$$
\overline{z_{*}} \int_{0}^{1}\left|W^{\prime}\left(t z_{*}\right)\right|^{2} \mathrm{~d} t=z_{*} \int_{0}^{1}\left|W\left(t z_{*}\right)\right|^{2} G\left(t z_{*}\right) \mathrm{d} t .
$$

Taking the real part, we get

$$
\begin{equation*}
x_{*} \int_{0}^{1}\left|W^{\prime}\left(t z_{*}\right)\right|^{2} \mathrm{~d} t=\int_{0}^{1}\left|W\left(t z_{*}\right)\right|^{2} \Re\left[z_{*} G\left(t z_{*}\right)\right] \mathrm{d} t \tag{65}
\end{equation*}
$$

where $x_{*}=\Re\left(z_{*}\right)$ and $y_{*}=\Im\left(z_{*}\right)$.
Assume now, by contradiction, that $F(\cdot)$ admits a root with negative real part, and take $z_{*}$ in (65) as equal to this root. The left-hand side of (65) is negative, however its right-hand side is nonnegative by assumption, yielding the desired contradiction. Hence, all roots of $F$ have nonnegative real parts, entailing the conclusion thanks to (64).

## 4 Comprehensive illustrative examples with insights on numerics

In this section, we provide some applications of the MID-based design in both retarded and neutral cases: the control of a transonic flow in a wind tunnel as well as the design of the standard PID controller for a prescribed stabilization of unstable delayed plants.

### 4.1 Control of a transonic flow in a wind tunnel

As a first application of the IMID-based design, let us revisit the regulation problem of the transonic flow in a wind tunnel. Transonic flows in a compressible fluid exhibit complex dynamics, making their analysis challenging since a full model of the flow is governed by a Navier-Stokes equation in a three-dimensional domain and boundary controls for temperature and pressure regulation. A simplified model can be found in [73], assuming that the flow is uniform across every cross section and the tunnel is a one-dimensional tube of varying cross-sectional area, yielding a coupled model of nonlinear partial differential equations in one space dimension.

In order to study the response of the Mach number of the flow to changes in the guide vane angle a further simplified model has been proposed in [7]. Propagation phenomena are modeled through a time delay, leading to the system of delay differential equations

$$
\left\{\begin{align*}
\kappa m^{\prime}(t)+m(t) & =k \vartheta\left(t-\tau_{0}\right),  \tag{66}\\
\vartheta^{\prime \prime}(t)+2 \zeta \omega \vartheta^{\prime}(t)+\omega^{2} \vartheta(t) & =\omega^{2} u(t)
\end{align*}\right.
$$

where $m, \vartheta$, and $u$ represent, respectively, perturbations of the Mach number of the flow, the guide vane angle, and the input of the guide vane actuator with respect to steady-state values. The parameters $\kappa$ and $k$ depend on the steady-state operating point and are assumed to be constant as long as $m, \vartheta$, and $u$ remain small, and satisfy $\kappa>0$ and $k<0$. The parameters $\zeta \in(0,1)$ and $\omega>0$ come from the design of the guide vane angle actuator and are thus independent of the operating point. The time-delay $\tau_{0}$ is assumed to depend only on the temperature of the flow. In the absence of control $(u(t)=0)$, the open-loop system (66) is exponentially stable.

The design of exponentially stabilizing feedback laws for (66) improving its exponential decay rate has been considered, for instance, in [48], in which the author designs a predictor of the state over an interval of length equal to the time-delay, yielding a closed-loop system with finite spectrum. However, the practical implementation of this type of controllers suffers from robustness issues [33,57], which motivates the research for control laws with reduced implementation complexity.

The goal of this section is to illustrate how the main result of this work can be used to obtain a feedback controller for (66) improving its open-loop characteristics, with a reduced complexity with respect to the controller proposed in [48]. We assume that the parameters $\zeta \in(0,1)$ and $\omega>0$ are fixed. Here we exploit the control law proposed in [50]

$$
\begin{equation*}
u(t)=-\frac{\beta_{0} \kappa}{k \omega^{2}} m\left(t-\tau_{1}\right)-\frac{\beta_{1} \kappa}{k \omega^{2}} m^{\prime}\left(t-\tau_{1}\right)-\frac{\beta_{2} \kappa}{k \omega^{2}} m^{\prime \prime}\left(t-\tau_{1}\right) \tag{67}
\end{equation*}
$$

where $\tau_{1}>0$ should be greater than or equal to the time-delay corresponding to measuring $m$ and its first two derivatives. Notice that the tuning of (67) made in [50] relies on the GMID property, making such a design sensitive with respect to parametric uncertainties, and it also assumed that one could choose the parameters $\zeta$ and $\omega$.

Substituting the control law (67) into (66), one obtains that the closed-loop characteristic quasipolynomial $\tilde{\Delta}$ is given by

$$
\begin{aligned}
\Delta(\lambda) & =\frac{\tilde{\Delta}(\lambda)}{\kappa} \\
& =\lambda^{3}+\left(2 \omega \zeta+\frac{1}{\kappa}\right) \lambda^{2}+\left(\omega^{2}+\frac{2 \omega \zeta}{\kappa}\right) \lambda+\frac{\omega^{2}}{\kappa}+\left(\beta_{2} \lambda^{2}+\beta_{1} \lambda+\beta_{0}\right) e^{-\lambda(\tau)}
\end{aligned}
$$

where $\tau=\tau_{0}+\tau_{1}$ and the division by $\kappa$ is performed in order to obtain a quasipolynomial under the form (3), for which the polynomial $P_{0}$ is monic.

As a consequence of the main results, one gets the following:
Theorem 4.1. A given complex number $\lambda_{0}$ is a root of multiplicity 4 of the quasipolynomial $\Delta$ from (68) if, and only if, $\lambda_{0}$ is a root of the elimination-produced polynomial $\mathbb{P}$ where

$$
\begin{align*}
\mathbb{P}(\lambda)= & \kappa \lambda^{3} \tau^{3}+\left(2 \zeta \kappa \tau^{3} \omega+9 \kappa \tau^{2}+\tau^{3}\right) \lambda^{2} \\
& +\left(\kappa \omega^{2} \tau^{3}+\left(12 \zeta \kappa \tau^{2}+2 \zeta \tau^{3}\right) \omega+18 \kappa \tau+6 \tau^{2}\right) \lambda  \tag{68}\\
& +\left(3 \kappa \tau^{2}+\tau^{3}\right) \omega^{2}+\left(12 \zeta \kappa \tau+6 \zeta \tau^{2}\right) \omega+6 \kappa+6 \tau
\end{align*}
$$

and the controller's gains satisfy

$$
\begin{aligned}
\frac{\beta_{0}}{\mathrm{e}^{\lambda_{0} \tau}}= & \left(\left(2 \tau \zeta^{2}-\tau\right) \omega^{2}+14 \omega \zeta+\frac{177}{2 \tau}+\frac{7}{\kappa}+\frac{\tau}{2 \kappa^{2}}\right) \lambda_{0}^{2} \\
& +\left(\zeta \tau \omega^{3}+\left(\left(12+\frac{2 \tau}{\kappa}\right) \zeta^{2}+7-\frac{\tau}{\kappa}\right) \omega^{2}+\left(\frac{\tau}{\kappa^{2}}+\frac{26}{\kappa}+\frac{162}{\tau}\right) \zeta \omega+\frac{246}{\tau^{2}}+\frac{81}{\kappa \tau}+\frac{3}{\kappa^{2}}\right) \lambda_{0} \\
& +\left(3+\frac{\tau}{\kappa}\right) \zeta \omega^{3}+\left(\left(\frac{6}{\kappa}+\frac{12}{\tau}\right) \zeta^{2}+\frac{\tau}{2 \kappa^{2}}+\frac{15}{\kappa}+\frac{87}{2 \tau}\right) \omega^{2} \\
& +\left(\frac{3}{\kappa^{2}}+\frac{99}{\kappa \tau}+\frac{180}{\tau^{2}}\right) \zeta \omega+\frac{87}{\tau^{3}}+\frac{90}{\kappa \tau^{2}}+\frac{3}{\kappa^{2} \tau} \\
\frac{\beta_{1}}{\mathrm{e}^{\lambda_{0} \tau}}= & \left(2 \omega \tau \zeta+21+\frac{\tau}{\kappa}\right) \lambda_{0}^{2}+\left(\left(36+\frac{4 \tau}{\kappa}\right) \omega \zeta+2 \omega^{2} \tau+\frac{18}{\kappa}+\frac{66}{\tau}\right) \lambda_{0} \\
& +\left(\frac{22}{\kappa}+\frac{48}{\tau}\right) \omega \zeta+\left(11+\frac{3 \tau}{\kappa}\right) \omega^{2}+\frac{24}{\kappa \tau}+\frac{24}{\tau^{2}} \\
\frac{\beta_{2}}{\mathrm{e}^{\lambda_{0} \tau}}= & \frac{3 \tau \lambda_{0}^{2}}{2}+\left(2 \omega \tau \zeta+6+\frac{\tau}{\kappa}\right) \lambda_{0}+\left(4+\frac{\tau}{\kappa}\right) \omega \zeta+\frac{\omega^{2} \tau}{2}+\frac{2}{\kappa}+\frac{3}{\tau}
\end{aligned}
$$

where $\tau=\tau_{0}+\tau_{1}$. Moreover, if the discriminant of $\mathbb{P}$ is positive and for every $t \in(0,1)$ and every root $z$ of $F$ in $\mathbb{C}_{-}$, we have $\Re[z G(t z)] \geq 0$ is satisfied where $G$ is defined by (61)-(62)-(17), then $s_{0}<0, s_{0}$ is a strictly dominant root of $\Delta$, and the trivial solution of (66) with the control law (67) is exponentially stable with exponential decay $-\lambda_{0}$.

Proof. The result is a direct consequence of Proposition 3.5 and Theorem 3.3. The realness of $\lambda_{0}$ is guaranteed by the positivity of the discriminant of $\mathbb{P}$ which is a of cubic polynomial in $\lambda$. The negativity of $\lambda_{0}$ is ensured by the fact that $\mathbb{P}$ is Hurwitz for any $(\tau, \kappa, \omega, \zeta) \in \mathbb{R}_{+} \times \mathbb{R}_{+} \times \mathbb{R}_{+} \times \mathbb{R}_{+}$.

Remark 8. As illustrated through the above example, despite the sufficient conditions established in Theorem 4.1, the use of Hille oscillation theorems [40] and the Green-Hille transform lacks the explicit character when one is dealing with a parametric study. As a matter of fact, one should further investigate the parametric semialgebraic problem $\Re[z G(t z)] \geq 0$, which is of huge complexity when a large number of parameters is involved and left free. In the sequel we illustrate the effective superiority of the approach when a priori bound using the algorithmic alternative described in Section 2.5 giving rise to more explicit conditions.

### 4.2 PID design of unstable low-order plants

PID control is the most popular control technology and dominates industrial control systems [66] because it represents one of the simplest control laws with a small number of control parameters giving satisfactory behaviors for closed-loop systems for large classes of industrial processes. Despite the fact that the design of such a controller is nowadays well-mastered to meet the performance requirements in stabilization and tracking of finite-dimensional industrial processes, to the best of the authors' knowledge, there is no systematical method in tuning a PID controller for infinite-dimensional systems. Recently, in [47], the MID property has been proposed as a method in tuning such controllers. Nonetheless, the proposed solution in [47] lacks freedom on the prescribed stabilization of the delayed first-order case study. By this section, we first revisit the considered problem and results from [47], then we improve such a solution by allowing some additional freedom when assigning the closed-loop decay rate.

Consider the feedback system depicted in Figure 2, in which $K(\lambda)$ represents a finite-dimensional linear time-invariant (LTI) controller, $Q_{\tau}(\lambda)$ denotes the plant containing a constant but uncertain delay $\tau$, with a transfer function given by

$$
\begin{equation*}
Q_{\tau}(\lambda)=\frac{1}{\lambda-\kappa} \mathrm{e}^{-\tau \lambda}, \quad \tau \geq 0 \tag{70}
\end{equation*}
$$

where $\kappa \geq 0$. The controller $K$ of interest is the standard PID controller, i.e., $K(\lambda)=K_{P I D}(\lambda)$ where


Figure 2: A tracking control system.

$$
\begin{equation*}
K_{P I D}(\lambda)=k_{p}+\frac{k_{i}}{\lambda}+k_{d} \lambda \tag{71}
\end{equation*}
$$

Thus, the closed-loop characteristic quasipolynomial is found as

$$
\left\{\begin{align*}
\Delta(\lambda) & =P_{0}(\lambda)+P_{\tau}(\lambda) \mathrm{e}^{-\lambda \tau}  \tag{72}\\
P_{0}(\lambda) & =\lambda^{2}-\lambda p, \quad P_{\tau}(\lambda)=k_{d} \lambda^{2}+k_{p} \lambda+k_{i}
\end{align*}\right.
$$

The main result of [47] is stated as follows.
Theorem 4.2 ( [47]). Let $P_{\tau}$ and $K_{P I D}$ be given respectively by (70) and $K_{P I D}(\lambda)$ by (71). Then the following statements are true.
i) For arbitrary real parameters $k_{p}, k_{i}, k_{d}$ and arbitrary positive delay $\tau$, the multiplicity of any root of the quasipolynomial $\Delta(\lambda)$ is less than or equal to 5 .
ii) The quasipolynomial $\Delta$ admits a multiple real root at

$$
\begin{equation*}
\lambda_{ \pm}=\frac{\tau \kappa-6 \pm \sqrt{\tau^{2} \kappa^{2}+12}}{2 \tau} \tag{73}
\end{equation*}
$$

with algebraic multiplicity at least 4 if and only if

$$
\left\{\begin{array}{l}
k_{d}=\frac{\left(4+2 \tau \lambda_{ \pm}-\tau \kappa\right) \mathrm{e}^{\tau s_{ \pm}}}{2}  \tag{74}\\
k_{p}=-\frac{\left(\left(8 \tau+\tau^{2} \lambda_{ \pm}\right) \kappa-18-12 \tau \lambda_{ \pm}\right) \mathrm{e}^{\tau \lambda_{ \pm}}}{\tau} \\
k_{i}=\frac{\left(\left(\tau \lambda_{ \pm}+3\right) \tau^{2} \kappa^{2}-\left(12 \tau \lambda_{ \pm}+60\right) \tau \kappa+108+84 \tau \lambda_{ \pm}\right) \mathrm{e}^{\tau \lambda_{ \pm}}}{2 \tau^{2}}
\end{array}\right.
$$

iii) If (74) is satisfied, then, under the condition $\tau<\bar{\tau}_{P I D}=2 / \kappa, \lambda=\lambda_{+}$is the rightmost root of the quasipolynomial $\Delta(\lambda)$ with multiplicity equal to 4 .
As mentioned earlier, under the assumption that the delay $\tau$ and the unstable pole $\kappa$ are fixed, the assignable quadruple root (the spectral abscissa) $\lambda_{+}$given in (73) is fully characterized, i.e., it does not allow any degree of freedom in the choice of the spectral abscissa. In order to enable some additional freedom when assigning it, one can relax such a constraint by forcing an over-order multiplicity which is lower than four, which, in our case, gives exactly the only option of multiplicity three.

Theorem 4.3. Consider the quasipolynomial $\Delta$ given in (72). Then the following statements are true.
i) The quasipolynomial $\Delta(\lambda)$ admits a triple real root at $\lambda_{0}$ if and only if it is real and the gains $\left(k_{d}, k_{p}, k_{i}\right)$ satisfy:

$$
\left\{\begin{array}{l}
k_{d}=\frac{\left(\lambda_{0}\left(-\lambda_{0}+\kappa\right) \tau^{2}+\left(2 \kappa-4 \lambda_{0}\right) \tau-2\right) \mathrm{e}^{\tau \lambda_{0}}}{2}  \tag{75}\\
k_{p}=\left(\lambda_{0}^{2}\left(\kappa-\lambda_{0}\right) \tau^{2}-\lambda_{0}\left(\kappa-3 \lambda_{0}\right) \tau+\kappa\right) \mathrm{e}^{\tau \lambda_{0}} \\
k_{i}=\frac{\lambda_{0}^{3} \tau\left(\left(\kappa-\lambda_{0}\right) \tau-2\right) \mathrm{e}^{\tau \lambda_{0}}}{2}
\end{array}\right.
$$

ii) Furthermore, if

$$
\left\{\begin{array}{l}
\kappa \in\left(0, \frac{1}{\tau}\right)  \tag{76}\\
\lambda_{0} \in\left[\frac{\kappa}{2}-\frac{2}{\tau}+\frac{\sqrt{\kappa^{2} \tau^{2}+8}}{2 \tau}, 0\right)
\end{array}\right.
$$

then $\lambda_{0}$ corresponds to the spectral abscissa of (72).
Proof. An integration by part shows that under (75), the quasipolynomial $\Delta$ reads:

$$
\begin{equation*}
\Delta(\lambda)=\tau\left(\lambda-\lambda_{0}\right)^{3} \int_{0}^{1} \mathrm{e}^{-\tau\left(\lambda-\lambda_{0}\right) t}\left(\frac{1}{2}\left(\lambda_{0}-\kappa\right) \lambda_{0} \tau^{2} t^{2}+\left(2 \lambda_{0}-\kappa\right) \tau t+1\right) \mathrm{d} t \tag{77}
\end{equation*}
$$

Thus, the corresponding kernel polynomial (39) is given by:

$$
\begin{equation*}
p(t)=\frac{\lambda_{0}\left(\lambda_{0}-\kappa\right) \tau^{2} t^{2}}{2}+\left(2 \lambda_{0}-\kappa\right) \tau t+1 \tag{78}
\end{equation*}
$$

which is necessarily real-rooted since its discriminant (w.r.t. the variable $t$ ) is given by $\delta=\left(\left(\kappa-\lambda_{0}\right)^{2}+\lambda_{0}^{2}\right) \tau^{2}$ $\geq 0$. Clearly, the leading coefficient of the polynomial $p$ is positive. Furthermore, the corresponding zeros are given by

$$
\begin{equation*}
t_{ \pm}=\frac{\kappa-2 \lambda_{0} \pm \sqrt{\left(\kappa-\lambda_{0}\right)^{2}+\lambda_{0}^{2}}}{\lambda_{0}\left(\lambda_{0}-\kappa\right) \tau} \tag{79}
\end{equation*}
$$

Since, $\kappa \geq 0$ and the closed-loop multiple zero $\lambda_{0}$ is intended to correspond to the exponential decay of (77), then it needs to satisfy $\lambda_{0}<0$, hence the denominator of $t_{ \pm}$given in (79) is positive, which yields that $0<t_{-}<t_{+}$. Hence, it suffices to set $t_{-} \geq 1$ to guarantee the sign constancy of the kernel polynomial $p$ given in (78). This assumption is equivalent to:

$$
\begin{equation*}
\sqrt{\kappa^{2}-2 \lambda_{0} \kappa+2 \lambda_{0}^{2}} \leq-\lambda_{0}^{2} \tau+(\tau \kappa-2) \lambda_{0}+\kappa \tag{80}
\end{equation*}
$$

An immediate necessary condition for the last inequality to hold is to select $\lambda_{0}$ such that

$$
\begin{equation*}
\lambda_{0} \in\left[\frac{\kappa}{2}-\frac{1}{\tau}-\frac{\sqrt{\kappa^{2} \tau^{2}+4}}{2 \tau}, \frac{\kappa}{2}-\frac{1}{\tau}+\frac{\sqrt{\kappa^{2} \tau^{2}+4}}{2 \tau}\right] \tag{81}
\end{equation*}
$$

Furthermore, by squaring both sides of inequality (80), one gets

$$
\lambda_{0}\left(\kappa-\lambda_{0}\right)\left(-\lambda_{0}^{2} \tau^{2}+\tau(\tau \kappa-4) \lambda_{0}+2 \tau \kappa-2\right) \geq 0
$$

Again, since $\lambda_{0}<0$ and $\kappa \geq 0$, the third factor of the left-hand side of the above inequality has to be nonpositive, which enables to choose $\lambda_{0}$ such that

$$
\begin{equation*}
\lambda_{0} \in \mathbb{R}_{-}^{*} \backslash\left(\frac{\kappa}{2}-\frac{2}{\tau}-\frac{\sqrt{\kappa^{2} \tau^{2}+8}}{2 \tau}, \frac{\kappa}{2}-\frac{2}{\tau}+\frac{\sqrt{\kappa^{2} \tau^{2}+8}}{2 \tau}\right) . \tag{82}
\end{equation*}
$$

Taking into account the intervals (81) and (82), one concludes that the sign constancy (positive) of the kernel polynomial $p$ given by (78) is guaranteed by

$$
\lambda_{0} \in\left[\frac{\kappa}{2}-\frac{2}{\tau}+\frac{\sqrt{\kappa^{2} \tau^{2}+8}}{2 \tau}, \frac{\kappa}{2}-\frac{1}{\tau}+\frac{\sqrt{\kappa^{2} \tau^{2}+4}}{2 \tau}\right]
$$

Finally, since $\lambda_{0}<0$ and the corresponding lower-bound is strictly increasing with respect to $\kappa$ and nonnegative for $\kappa \tau \geq 1$ and the corresponding upper-bound is positive, then necessarily $\kappa \tau<1$. As illustrated in Figure 3, the appropriate choice of $\lambda_{0}$ is such that

$$
\begin{equation*}
\lambda_{0} \in\left[\frac{\kappa}{2}-\frac{2}{\tau}+\frac{\sqrt{\kappa^{2} \tau^{2}+8}}{2 \tau}, 0\right) \tag{83}
\end{equation*}
$$



Figure 3: Region in the parameters $\left(\kappa, \lambda_{0}\right)$ plane described in (83) where the sign constancy of (78) is guaranteed. The delay $\tau$ is taken unitary.

Next, one needs first to write $\Delta$, or, equivalently, the corresponding integral representation given in (77), into a normalized form. This is done using the following standard affine change of variables $[24,50]$

$$
\begin{equation*}
\lambda=\frac{s}{\tau}+\lambda_{0} \tag{84}
\end{equation*}
$$

and the new parametrization $\kappa \tau=\xi$ allowing to write (72) satisfying (75) as

$$
\left\{\begin{array}{l}
\hat{\Delta}(s)=\tau^{2} \Delta\left(\lambda=\frac{s}{\tau}+\lambda_{0}\right)=\hat{P}_{0}(s)+\hat{P}_{\tau}(s) \mathrm{e}^{-s}  \tag{85}\\
\hat{P}_{0}(s)=s^{2}+\left(2 \lambda_{0}-\xi\right) s+\lambda_{0}\left(-\xi+\lambda_{0}\right) \\
P_{\tau}(s)=\left(-\frac{\lambda_{0}^{2}}{2}+\frac{(\xi-4) \lambda_{0}}{2}+\xi-1\right) s^{2}+\left(-\lambda_{0}^{2}+\frac{(2 \xi-4) \lambda_{0}}{2}+\xi\right) s+\lambda_{0}\left(\xi-\lambda_{0}\right)
\end{array}\right.
$$

or, equivalently,

$$
\begin{equation*}
\hat{\Delta}(s)=\tau s^{3} \int_{0}^{1} \mathrm{e}^{-s t}\left(1-\frac{\lambda_{0}\left(\xi-\lambda_{0}\right) t^{2}}{2}+\frac{\left(4 \lambda_{0}-2 \xi\right) t}{2}\right) \mathrm{d} t \tag{86}
\end{equation*}
$$

Obviously, to show that $\lambda_{0}$ is the spectral abscissa of (72) with control parameters satisfying (75) amounts to show that 0 is the spectral abscissa of (85) or, equivalently, (86). Thanks to the algorithm from [13,51] also recalled in Section 2.5 as Algorithm 1, one is now able to investigate the frequency bound for potential spectral values with positive real parts for (85).

To do so, let us assume that there exists $s_{0}=x_{0}+i \omega_{0} \in \mathbb{C}_{+}$such that $\hat{\Delta}\left(s_{0}\right)=0$, then necessarily

$$
\begin{equation*}
\left|\hat{P}_{0}\left(x_{0}+i \omega_{0}\right)\right|^{2} \mathrm{e}^{2 x_{0}}=\left|\hat{P}_{\tau}\left(x_{0}+i \omega_{0}\right)\right|^{2} \tag{87}
\end{equation*}
$$

Consider then the family of auxiliary functions

$$
F_{n}(x, \omega)=\left|\hat{P}_{\tau}(x+i \omega)\right|^{2}-\mathcal{T}_{n}\left(\mathrm{e}^{2 x}\right)\left|\hat{P}_{0}(x+i \omega)\right|^{2}
$$

where $\mathcal{T}_{n}$ designates the $n$-th order Taylor approximation. Since $\mathrm{e}^{2 x}>\mathcal{T}_{n}\left(\mathrm{e}^{2 x}\right) \geq 1$ for any $x>0$, we then have $F_{n}\left(x_{0}, \omega_{0}\right)>0$. In particular, one has

$$
F_{0}(x, \omega)=\left|\hat{P}_{\tau}(x+i \omega)\right|^{2}-\left|\hat{P}_{0}(x+i \omega)\right|^{2}>0
$$

The idea then amounts to show that if $s_{0}=x_{0}+i \omega_{0}$ is a root of $\hat{\Delta}$ such that $x_{0}>0$ and $\omega_{0} \geq \pi$, then $F\left(x_{0}, \omega_{0}\right) \leq 0$, thus resulting a contradiction. Toward this end, we set $\omega=\sqrt{\Omega}$ and establish the explicit expression

$$
\begin{aligned}
F_{0}(x, \sqrt{\Omega})= & \frac{\left(\lambda_{0}+2\right)\left(-\lambda_{0}+\xi-2\right) \mu\left(\lambda_{0}, \xi\right)}{4} \Omega^{2} \\
& +\left(\frac{\left(\lambda_{0}+2\right)\left(-\lambda_{0}+\xi-2\right) \mu\left(\lambda_{0}, \xi\right)}{2} x^{2}+\nu\left(\lambda_{0}, \xi\right) x\right) \Omega \\
& +\frac{\left(\lambda_{0}+2\right)\left(-\lambda_{0}+\xi-2\right) \mu\left(\lambda_{0}, \xi\right)}{2} x^{4}+\nu\left(\lambda_{0}, \xi\right) x^{3} \\
& +\left(2 \lambda_{0}\left(\xi-\lambda_{0}\right)\left(-\lambda_{0}^{2}+(\xi-4) \lambda_{0}+2 \xi\right)\right) x^{2}+2 \lambda_{0}^{2}\left(\xi-\lambda_{0}\right)^{2} x
\end{aligned}
$$

with

$$
\begin{aligned}
\mu\left(\lambda_{0}, \xi\right) & =\left(-\lambda_{0}^{2}+(\xi-4) \lambda_{0}+2 \xi\right) \\
\nu\left(\lambda_{0}, \xi\right) & =\left(\lambda_{0}^{4}+(-2 \xi+6) \lambda_{0}^{3}+\left(\xi^{2}-9 \xi+10\right) \lambda_{0}^{2}+\left(3 \xi^{2}-10 \xi\right) \lambda_{0}+2 \xi^{2}\right)
\end{aligned}
$$

Seen as a quadratic polynomial in $\Omega$, the function $F_{0}$ is positive only between its real roots (if they exist) since (83) suggests that the leading coefficient of $F_{0}$ is negative. Let us investigate the existence of positive solutions for $F_{0}$.

Fix $x>0$. The discriminant of $F_{0}$ in $\Omega$ is given by

$$
\begin{aligned}
D(x)=- & \left(\lambda_{0}^{8}+(-4 \xi+12) \lambda_{0}^{7}+\left(6 \xi^{2}-42 \xi+48\right) \lambda_{0}^{6}+\left(-4 \xi^{3}+54 \xi^{2}-144 \xi+72\right) \lambda_{0}^{5}\right. \\
& +\left(\xi^{4}-30 \xi^{3}+155 \xi^{2}-180 \xi+28\right) \lambda_{0}^{4}+\left(6 \xi^{4}-70 \xi^{3}+152 \xi^{2}-56 \xi\right) \lambda_{0}^{3} \\
& \left.+\left(11 \xi^{4}-48 \xi^{3}+20 \xi^{2}\right) \lambda_{0}^{2}+\left(4 \xi^{4}+8 \xi^{3}\right) \lambda_{0}-4 \xi^{4}\right) x^{2} \\
- & 2\left(\xi-\lambda_{0}\right)^{2} \lambda_{0}^{2}\left(-\lambda_{0}+\xi-2\right)\left(\lambda_{0}+2\right)\left(-\lambda_{0}^{2}+(\xi-4) \lambda_{0}+2 \xi\right) x
\end{aligned}
$$

which is nothing but a real-rooted second-order polynomial in $x$.
Notice that, if the discriminant $D$ is negative for any positive $x$ in some sub-region of $0<\xi<1$ and $\lambda_{0}$ satisfies (83), then $F_{0}$ cannot be positive. In that case, it is clear that the MID applies.

However, if $D$ is positive then $F_{0}$ is positive only in the interval $\left(\Omega_{-}, \Omega_{+}\right)$where

$$
\begin{aligned}
\Omega_{ \pm}(x)= & -x^{2}-\frac{2\left(\lambda_{0}^{4}+(-2 \xi+6) \lambda_{0}^{3}+\left(\xi^{2}-9 \xi+10\right) \lambda_{0}^{2}+\left(3 \xi^{2}-10 \xi\right) \lambda_{0}+2 \xi^{2}\right) x}{\left(-\lambda_{0}+\xi-2\right)\left(\lambda_{0}+2\right)\left(-\lambda_{0}^{2}+(\xi-4) \lambda_{0}+2 \xi\right)} \\
& \mp \frac{\sqrt{D(x)}}{\left(-\lambda_{0}+\xi-2\right)\left(\lambda_{0}+2\right)\left(-\lambda_{0}^{2}+(\xi-4) \lambda_{0}+2 \xi\right)}
\end{aligned}
$$

By using some tedious but simple algebraic estimates, one obtains that

$$
\begin{equation*}
\Omega_{+}(x)<\Gamma(x)=(2+\sqrt{2}) x-x^{2} \tag{88}
\end{equation*}
$$

Interestingly, $\Gamma$ is a second-order polynomial in $x$, which is positive only in $x \in(0,2+\sqrt{2})$ and reaches its maximum $\Gamma_{\max }=\frac{3}{2}+\sqrt{2}$ at $x=1+\frac{\sqrt{2}}{2}$.

Clearly, in this case $\Omega=\omega^{2}<\pi^{2}$, which allows to conclude that the imaginary part of the integral factor in (86) satisfies

$$
\int_{0}^{1} \mathrm{e}^{-x t} \sin (\omega t)\left(1-\frac{\lambda_{0}\left(\xi-\lambda_{0}\right) t^{2}}{2}+\frac{\left(4 \lambda_{0}-2 \xi\right) t}{2}\right) \mathrm{d} t \neq 0 \quad \text { for } \quad(x, \omega) \in \mathbb{R}_{+}^{*} \times \mathbb{R}_{+}^{*}
$$

which ends the proof.

## 4.3 $\mathbf{P 3} \delta$ Software

The authors developed recently an intuitive Python software called Partial pole placement via delay action (P3 $\delta$ ) [19-21]. P3 (https://cutt.ly/p3delta) enables the design of delayed feedback control laws rendering the closed-loop dynamics stable with a prescribed exponential decay rate. P3 $\delta$ methodology relies on two properties of quasipolynomial's zeros distribution: (i) the MID and (ii) the CRRID [5, 10], for coexisting-real-root-induced-dominancy.

While the MID has been highlighted through this paper, the CRRID property consists in conditions on the system's parameters guaranteeing the dominance of coexistent real spectral values. When using the MID strategy on P3 $\delta$, two options are proposed: the GMID-based design and the control-oriented IMID-based design. The first option relies on the fact that a root of maximal multiplicity $M\left(\lambda_{0}\right)=n+m+1$ is necessarily dominant and the latter exploits the over-order intermediate multiplicity $M\left(\lambda_{0}\right)=n+1$, offering sufficient freedom in parameters' choice. In future software developments, the authors will integrate the result of this work into new P3 $\delta$ functionality allowing the users to exploit the IMID-based design with intermediate multiplicities $n+1 \leq M\left(\lambda_{0}\right) \leq n+m+1$, yielding further freedom for control purpose.

## 5 Concluding Remarks

The MID property defines an intriguing link between multiple spectral values and the spectral abscissa of a given plant opening promising prospects in prescribed stabilization of both finite and infinite-dimensional dynamical systems. While the GMID property $\left(M\left(\lambda_{0}\right)=m+n+1\right)$ has been fully characterized in $[15,50]$ thanks to the hypergeometric representation of the corresponding quasipolynomial, on beyond of some partial results [9,18] the over-order intermediate MID remained an open question. In this work, the over-order intermediate MID is investigated in depth. Thanks to a Hille oscillation theorem in the complex domain, we first provide a unified proof for the IMID to hold when $n+1 \leq M\left(\lambda_{0}\right) \leq m+n$ by exploiting the representation of the quasipolynomial as a linear combination of contiguous Kummer hypergeometric functions. Second, for the sake of the effectiveness of the MID-based design, an algorithmic method relying on an a-priori bound on the frequency of complex roots with positive real part. Both strategies have been illustrated through concrete control applications in the retarded as well as in the neutral cases: prescribed regulation of the Mach number in a wind-tunnel and the systematic PID control design for unstable first-order plants with input delay. Both examples, as well as some recent works such as $[3,4,35]$, show the true potential of the proposed partial pole placement not only for ODEs and DDEs but also for some classes of PDEs.

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    ${ }^{1}$ initial value problem, existence, uniqueness of the solutions, linear systems, stability theory
    ${ }^{2}$ in two classes: retarded and neutral DDEs

[^1]:    ${ }^{3}$ In this context, the terminology of differential-difference equations introduced by Bellman and Cooke [11] should be mentioned. The authors made an explicit distinction between the differential order corresponding to the highest derivative order appearing in the equation and the so-called difference order that corresponds to the number of distinct (time) arguments appearing in the (same) equation minus one; for further arguments concerning the terminology, see, for instance, [60] and the references therein.
    ${ }^{4}$ The notion of "parameters" is seen in a large sense, including all coefficients and delay terms.
    ${ }^{5}$ For instance, in the inverted pendulum case, we may have one triple characteristic root at the origin of the linearization of the dynamical system in closed-loop represented by a second-order DDE.
    ${ }^{6}$ More precisely, the reader is referred to Chapters 3 (characteristic equations), 4 (scalar DDE) and 5 (second-order DDE) in [63].
    ${ }^{7}$ Complex characteristic roots located in the right half-plane.
    ${ }^{8}$ also including the delays (seen as parameters)

[^2]:    ${ }^{9}$ complete regular splitting or not
    ${ }^{10}$ The corresponding method is known as the $\tau$-decomposition method, see, e.g., [54] for a deeper discussion on such topics.
    ${ }^{11}$ The spectral abscissa represents the rightmost characteristic root of the (corresponding characteristic) function.

[^3]:    ${ }^{12}$ To the best of the authors' knowledge, contiguous functions appear in the works of Gauss (1813) and he was the first to define the equations which relate such functions
    ${ }^{13}$ under the assumption that the flow is uniform across every cross section and the tunnel is a one-dimensional tube of varying crosssectional area, leading to a a coupled model of nonlinear partial differential equations in one space dimension
    ${ }^{14}$ For a more detailed discussion on the existing tuning rules for PID controllers, we refer to [61].

[^4]:    ${ }^{15}$ in the case when the highest order of derivation appears only in the non-delayed term $y^{(n)}(t)$

[^5]:    ${ }^{16}$ linear convection in this case
    ${ }^{17}$ Scalar closed-loop characteristic function of neutral type ( $m=n=1$ ).

